INTEGRATED SCIENCE

FOR TTCs

YEAR ONE

STUDENT'S BOOK

OPTION: SME

© 2020 Rwanda Basic Education Board

All rights reserved

This book is property of the Government of Rwanda. Credit must be given to REB when the content is quoted.

FOREWORD

Dear Student- teacher,

Rwanda Basic Education Board is honoured to present to you this Integrated Science book for Year One of TTC SME option which serves as a guide to competence-based teaching and learning to ensure consistency and coherence in the learning of Integrated Science subject. The Rwandan educational philosophy is to ensure that you achieve full potential at every level of education which will prepare you to be well integrated in society and exploit employment opportunities.

The government of Rwanda emphasizes the importance of aligning teaching and learning materials with the syllabus to facilitate your learning process. Many factors influence what you learn, how well you learn and the competences you acquire. Those factors include the instructional materials available among others. Special attention was paid to the activities that facilitate the learning process in which you can develop your ideas and make new discoveries during concrete activities carried out individually or with peers.

In competence-based curriculum, learning is considered as a process of active building and developing knowledge and meanings by the learner where concepts are mainly introduced by an activity, a situation or a scenario that helps the learner to construct knowledge, develop skills and acquire positive attitudes and values. For effective use of this textbook, your role is to:

- Work on given activities including laboratory experiments which lead to the development of skills;
- Share relevant information with other learners through presentations, discussions, group work and other active learning techniques such as role play, case studies, investigation and research in the library, from the internet or from your community;
- Participate and take responsibility for your own learning;
- Draw conclusions based on the findings from the learning activities.

I wish to sincerely extend my appreciation to the people who contributed towards the development of this book, particularly REB staff who organized the whole process from its inception. Special gratitude goes to teachers, illustrators and designers who diligently worked to successful completion of this book.

Dr. MBARUSHIMANA Nelson

Director General, REB

ACKNOWLEDGEMENT

I wish to express my appreciation to all the people who played a major role in development of this Integrated Science textbook for Year One of TTC. It would not have been successful without active participation of different education stakeholders.

I owe gratitude to different Universities and schools in Rwanda that allowed their staff to work with REB in the in-house textbooks production project. I wish to extend my sincere gratitude to lecturers, teachers, tutors, illustrators, designers and all other individuals whose efforts in one way or the other contributed to the success of writing of this textbook.

Finally, my word of gratitude goes to the Rwanda Basic Education Board staff particularly those from the Curriculum, Teaching and Learning Resources Department who were involved in the whole process of in-house textbook writing.

Joan MURUNGI

Head of Curriculum, Teaching and Learning Resources Department

TABLE OF CONTENT

FOREWORD iii

ACKNOWLEDGEMENT iv

UNIT 1: THE CONCEPT OF INTEGRATED SCIENCE AND MEASUREMENTS OF PHYSICAL QUANTITIES 1

1.1. Introduction to Integrated science 2
1.2. Measurement of physical quantities 7
1.3. Dimensions of physical quantities 15

UNIT 2: INTRODUCTION TO BIODIVERSITY 23

2.1. Ecological terms: Species, Ecosystem, Niche, Population, Community and Biodiversity 24
2.2. Determination of the distribution and abundance of organisms in an area 29
2.3. Spearman's rank, Pearson's linear correlation and Simpson's Index of Diversity 33

UNIT 3: INTRODUCTION TO CLASSIFICATION 39

3.1.Taxonomic hierarchy and Three domains of life: Archaea, Bacteria and Eukarya 40
3.2. Characteristic features of the kingdoms 44
3.3. Classification of viruses and their economic importance 54
3.4. Dichotomous keys for identification of organism 55

UNIT 4: SOLUTIONS AND TITRATION 62

4.1. Definition of standard solution and primary standard solution 63
4.2. Properties of a primary standard solution 64
4.3. Preparation of standard solutions 65
4.4. Titration process focussing on precise measurements 71

UNIT 5: THE LENSES 87

5.1. Types of lenses and their characteristics 88
5.2. Refraction of light through lenses 90
5.3. The thin lens equations 94
5.4. Defects of lenses, their corrections and applications of lenses 101
5.5. Refraction through prisms 103

UNIT 6: SIMPLE AND COMPOUND OPTICAL INSTRUMENTS 115
6.1. Introduction to optical instruments and their image formation 116
6.2. Magnifying power of optical instruments and its Calculation 122
6.3. Astronomical Telescope 127
6.4. Human eye as single lens system 130
6.5. Defects of vision and their correction 135
UNIT 7: PROPAGATION OF MECHANICAL WAVES 141
7.1. Introduction to waves 142
7.2. Properties of waves: reflection, refraction, diffraction 147
7.3. Progressive waves and its mathematical treatment 154
7.4. Young's double slit experiment and its mathematical treatment 157
7.5. Stationary waves 160
UNIT 8: COMMON DISEASES AND HYGIENE 169
8.1. Common diseases 169
8.2. Hygiene practices and their importance 208
8.3. Human immune system: Immunity, structure and role of antibodies 211
8.4. Common addictive substances and their effects 218
8.5. Balanced diet 229
8.6. Basic food service technics 234
UNIT 9: MICROSCOPE AND CELL STRUCTURE 239
9.1. Cell theory and microscopes 239
9.2. Eukaryotic and Prokaryotic cells 247
9.3. Plant and animal cells 249
9.4. Organelles and their functions 251
9.5. Specialized cells 261
UNIT 10: CELL AND NUCLEAR DIVISION 270
10.1. Cell cycle 271
10. 2. Mitosis and Meiosis: Stages and results 273
10.3. Mitosis and Meiosis roles in living organisms 282

UNIT 11: AUTOTROPHIC NUTRITION 292
11.1. Types of autotrophic nutrition 292
11.2. Structure of the chloroplast and Adaptations for photosynthesis.. 294
11.3. Absorption, action spectra and other carbon dioxide fixation pathways 298
11.4. Rate of photosynthesis: limiting factors of photosynthesis 309
11.5. Importance of autotrophic nutrition and Tests for starch in terrestrial plants and for oxygen in aquatic plants 312

UNIT 12: THE CHEMICAL BASIS OF LIFE 319
12.1. Biological molecules 319
12.2. Test for the presence of different biological molecules in variety of context 339
12.3. The structure of DNA 344
12.4. Water and Enzymes 349

UNIT 13: KIRCHHOFF'S LAWS IN ELECTRIC CIRCUITS 369
13.1. Simple electric circuits and its construction 370
13.2. Measurement of electric current and voltage in a simple electric circuit 374
13.3. Different sources of electric current 377
13.4. Connection of electrical current source and resistors 380
13.5. Kirchhoff's laws and its applications in solving problems in electric circuits 388

UNIT 14: ASEXUAL AND SEXUAL REPRODUCTION IN PLANTS 398
14.1. Comparison between sexual and asexual reproduction 399
14.2. Asexual reproduction in plants 400

UNIT 15: STRUCTURE OF AN ATOM AND MASS SPECTRUM 435
15.1. The constituents of an atom, their properties and the outline of their discovery 436
15.2. Concept of atomic number, mass number, isotopic mass and relative atomic mass 443
15.3. Calculations of the relative atomic masses of elements 445
15.4. Mass spectrometer 448

UNIT 16: ELECTRON CONFIGURATIONS OF ATOMS AND IONS 464
16.1. Bohr's atomic model: Concept of energy levels and spectra 464
16.2. Hydrogen spectrum and spectral line series 471
16.3. Quantum theory of the atom 476
16.4. Electronic configuration of atoms and ions 486

UNIT 17: IONIC AND METALLIC BONDS 503
17.1. Explanations of why atoms of elements form bonds 504
17.3. Ionic bonding 507
17.4. Metallic bonding 513
18.1. Historical background of the Periodic Table 518

UNIT 18: VARIATION IN TRENDS OF THE PHYSICAL PROPERTIES 518
18.2. Comparison of Mendeleev's table with the modern Periodic Table 522
18.3. Location of the elements in the Periodic Table based on the electronic configuration 524
18.4. Classification of the elements into blocks (s, p, d, f-block) 526
18.5. Factors that influence the change of each physical property of the elements across a period and down a group 529
18.6. Variation of the physical properties down the group and across the period 530

UNIT 19: COVALENT BOND AND MOLECULAR STRUCTURES 541
19.1. Theories on the formation of covalent bond 542
19.2. Coordinate or dative covalent bond 550
19.3. Overlap of atomic orbitals to form covalent bonds 552
19.4. The concept of valence bond theory and formation of (σ) and (π) bonds 554
19.5. Hybridisation and its types 557
19.6. VSEPR theory 562
19.7. Polarity of the covalent bond in relation to difference in electronegativity 571
19.8. Physical properties of covalent structures 575
19.9. Intermolecular forces 580

UNIT 20: MOMENTS AND EQUILIBRIUM OF BODIES 594
20.1. Vector and scalar quantities 595
20.2. Types of equilibrium: stable, unstable and neutral 598
20.3. Condition for equilibrium of a body about an axis and Stevinus proof 600
20.4. Forces and moments in equilibrium 605
20.5. Archimedes' principle of the lever and centre of gravity of bodies 612

UNIT 21: PROJECTILE AND UNIFORM CIRCULAR MOTION 623
21.1. Projectile Motion 624
21.2. Motion of a projectile thrown obliquely 631
21.3. Circular motion: Meaning and related key terms 637
21.4. Relationship between angular and linear parameters 639
21.5. Acceleration in circular motion: Constant and tangential acceleration 643
21.6. Centripetal force 647
21.7. Applications of circular motion 648

UNIT 22: GENERAL STRUCTURE OF SOLAR SYSTEM 657
22.1. Astronomical scales 658
22.2. Sun-earth-moon system: eclipses and phases of the moon 660
22.3. Solar system 665
22.4. Kepler's Laws 674
22.5. Star Patterns: Constellations 677

REFERENCES 682

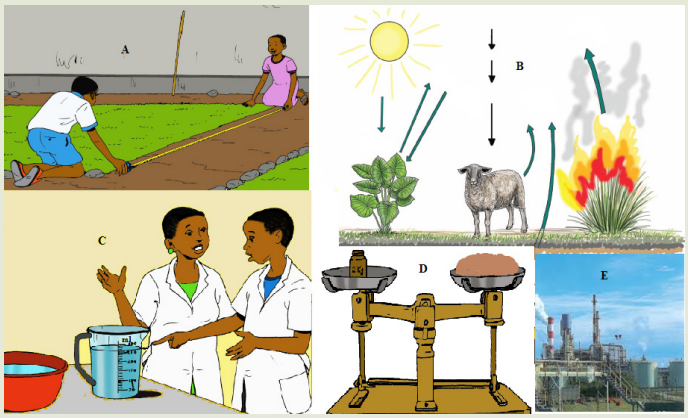
UNIT 1:

THE CONCEPT OF INTEGRATED SCIENCE AND MEASUREMENTS OF PHYSICAL QUANTITIES

Key Unit competence: Explain the concept of Integrated science and use accurately different tools to measure physical quantities in sciences.

Introductory Activity 1

Look carefully the following illustrations and answer the questions below:



Questions:

a). Describe the illustration A, B, C, D.

b). Based on your knowledge from O-level, what are scientific concepts can you associate to each of those illustrations? Group the noted concepts in their science subject areas.

c). Is there any one illustration in which you find application of many science subjects area? Justify your answer by providing other examples found in everyday life.

d). Can you explain how and why every person should have integrated understanding of those science subject areas?

e). What kind of physical quantities that can be measured in the illustration above? Suggest the names of the tools used in the illustration above?

f). Outline other examples of physical quantities and the corresponding measuring tools

g). What can be considered to select the best tool(s) to be used in measuring a given measurable quantity?

1.1. Introduction to Integrated science

Activity 1.1

Task 1

It is known that an Integrated Science course serves the purpose of unifying sciences in a whole one subject covering both the physical and life sciences. These courses are integrated in that the fields of science are not segmented. For example, in describing the physics of light, we show how this applies to the inner workings of our eyes, which, in turn, are sensitive to visible light in great part because of the chemical composition of our atmosphere.

Use the paragraph above to answer the following questions:

a). What does the term integrated science mean?

b). Explain why Integrated Science is very important in finding appropriate solutions in various complex situations? Justify your answer based on the paragraph above and other examples observed in everyday life.

Task 2

Suppose you visited two industries and took the photos A and B below and saw that distinguished science subjects are involved in the process of production. Write a paragraph about your visit identifying how Physics, Biology and Chemistry are integrated in the process.

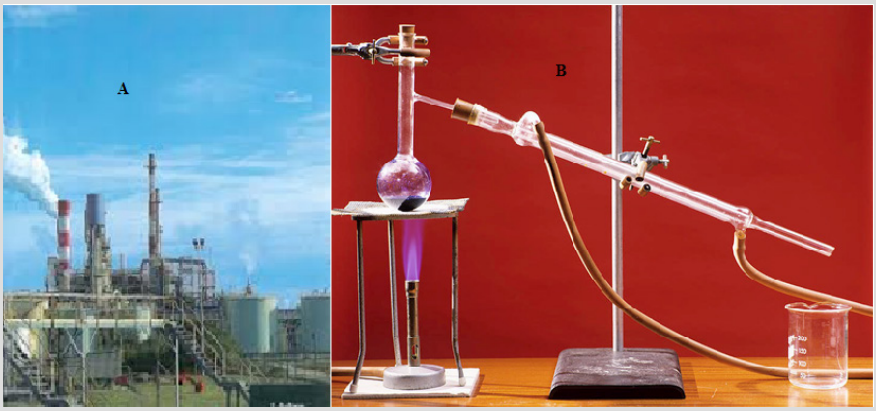


Figure 1.1.Photos took in field visit

1.1.1. Definition and rationale of Integrated science

Human survival depends on knowledge through the exploration of the environment. Science provides knowledge while technology provides ways of using this knowledge. It is therefore very important to be aware of the global dimension of science needed in our lives in order to effectively deal with every day situation.

The word "integrated" means *"to restore the whole, to come together, to be a part of, to include."* Integrated science is a subject which incorporates the knowledge base of all the science fields, both physical and life sciences and these science fields are included in one subject as a whole *"integrated science"* in that the fields of science are not segmented. It is a subject which offers experiences which help people to develop an operational understanding of the structure of science that should enrich their lives and make them more responsible citizens in the society.

Hence, integrated approach of learning science is appropriate as science knowledge is a tool to be used by every person to effectively deal with real world problems and life.

For examples, when you are studying digestion process of animals, you will need the knowledge of chemical processes. Another example, in describing the physics of light, we show how this applies to the inner workings of our eyes, which, in turn, are sensitive to visible light in great part because of the chemical composition of our atmosphere.

Aims of Integrated Science subject

The overall aim of the integrated science subject is to enable students

develop scientific literacy so that students can participate actively in the rapidly changing knowledge based society, prepare for further studies or careers in fields where the knowledge of science will be useful.

However, the broad aims of integrated science subject are to enable students to:

- Develop interest in and maintain a sense of wonder and curiosity about the natural and technological world;
- Acquire a broad and general understanding of key science ideas and explanatory framework of science and appreciate how the ideas were developed and why they are valued;
- Develop skills for making scientific inquiries;
- Develop the ability to think scientifically, critically and creatively and to solve problems individually or collaboratively in science related contexts;
- Use the language of science to communicate ideas and views on science – related issues;
- Make informed decisions and judgments about science related issues;
- Be aware of the social, ethnical, economic, environmental and technological implications of science and develop an attitude of responsible citizenship; and
- Develop conceptual tools for thinking and making sense of the world.

1.1.2. Interconnection between science subjects

The purpose of science is to produce useful models of reality which are used to advance the development of technology, leading to better quality of life for human being and the environment around him or her.

There are many branches of science and various ways of classifying them. One of the most common ways is to classify the branches into natural sciences, social sciences, and formal sciences.

Natural sciences: the study of natural phenomena (including cosmological, geological, physical, chemical, and biological factors of the universe). Natural science can be divided into two main branches: physical science and life science (or biological science). Social sciences: the study of human behavior and societies. The social sciences include, but are not limited to: anthropology, archaeology, communication studies, economics, history, musicology, human geography, jurisprudence, linguistics, political science, psychology, public health, and sociology. Formal science is a branch of

science studying formal language disciplines concerned with formal systems, such as logic, mathematics, statistics, theoretical computer science, artificial intelligence, information theory, game theory, systems theory, decision theory, and theoretical linguistics.

Note:

- Chemistry mainly deals with the study of matter's properties and behaviors as well as reactions between them to produce new useful products. For a physicist to understand the working mechanism of chemical cells, help is sought from a chemist. On the other hand, the reasons behind the various colours observed in most of the chemical reactions are explained by a physicist.

  Petroleum products are dealt with by the chemist, but the transportation of such products make use of the principles of physics.

- In Biology, the study of living cells and small insects by a biologist requires magnification. The concept of magnification using simple or compound microscope is a brain child of a physicist. A good physicist needs to have good health.

1.1.3. Relationship between Integrated science with other subjects

As science is about observation and experimentation of things in the physical and natural world, the relationship of Integrated science and other subjects might be explained in broader senses and will also predict much broader interconnections as applications of science are useful in human daily life. Below are some examples of relationship between Integrated science with other subjects:

Science with Mathematics:

A large number of scientific principles and rules are represented in the form of mathematical expressions, for which it is very necessary for person intending to get advanced study of science subjects to have sound mathematical basis. Without making use of mathematical expressions and rules, it is not possible to learn science in effective manner. Therefore, mathematics is considered to be sole language of science because of which real understanding of science is considered to be impossible without adequate knowledge of mathematics. Some of the useful mathematical tools which are generally used in the science are algebraic equations, geometrical formulas, graphs etc. For example, Astrology is an advanced branch of science in which it is predicted or enumerated that which planet revolves at which speed and when it will get appeared to the people of earth.

Science with History:

It sounds quite amazing that some kind of correlation can exist in between the science and history as earlier subject is practical in nature while nature of later subject is purely theoretical. However, it is possible to co-relate these subjects with each other. For example, in History, the determination of age fossils by historians and archaeologists use the principle developed by physicists. The medicine science lists the incidences which inspired various scientists to found out the medical remedies of various diseases.

Science with Geography:

Geography is the subject in which various concepts relating to earth on which we live are dealt with. Everything existing on earth, on different planets of the universe are also main subjects of geography. Which kind of crop should be sown in which kind of soils, how many kinds of rocks are found on the earth are some of the main topics which are covered by Geography. These topics are also covered by the subject of Science.

In science, there are various concepts relating to the atmosphere and earth in which living and non-living beings. For this reason, temperature, wind directions and measurement of rainfall are conducted in the subject of science by making use of various apparatus. For example, in Geography, weather forecast, a geographer uses a barometer, wind gauge, etc. which are instruments developed by a physicist.

Results obtained by the science in terms of climate and the manner in which it affects the human beings and earth are being interpreted by subject of Geography. The manner in which it is mentioned by the geography how soil gets produced through crushing process of rocks makes the subject a special branch of science.

As there are various topics which are of common interest for geographers and scientists, it can be said that both of these subjects are very near to each other and complementary to each other.

Science with Social Studies:

Various evidences can be found in our life which can show the significant way in which life style of human beings have got affected by inclusion of scientific developments in their life. Today, there are various kinds of machines for performing different functions, about which primitive men even did not think.

As a result of these machines, human life has become very easy and smooth and now we can accomplish complex functions within short period of time, which were considered to be very time consuming. Again, scientific

researches have led to development of various medicines with the help of which physicians have found the remedies of various diseases, which were once considered to be incurable and were responsible for bringing about heavy loss of life in earlier times.

Science with Physical education and sports:

In games and sports, different instruments developed by physicists are used for accurate measurement of time, distance, mass and others.

Application activity 1.1

1. Write a paragraph to convince someone that science is related to other subjects. Use clear examples to support your arguments and reasoning.
2. How can you describe the interconnections between science and technology, using at least three specific examples?

1.2. Measurement of physical quantities

Activity 1.2

Task 1:

Look around the place and identify possible physical quantities that can be measured? Explain the meaning of the physical quantities you have identified? Mention the SI units of the identified physical quantities?

Task 2:

It is possible to determine the nature and magnitude of the physical quantities that are measurable. Which of the following situations can be determined with the guidance of measurements? Support your answer with explanations and mention the physical quantity to be measured if possible.

a). Love between a boy and girl.

b). Size of the body

c). Size of the garden?

d). Amount occupied by water in a tank.

1.2.1. Physical quantities and their measurements

A quantity is any observable property or process in nature with which a number may be associated.

A physical quantity is defined as a property of a material that can be quantified by measurement.

Physical quantities are classified into fundamental and derived quantities.

Fundamental physical quantities

A *quantity* may be defined *as any observable property or process in nature with which a number may be associated*. This number is obtained by the operation of measurements. The number may be obtained directly by a single measurement or indirectly, say for example, by multiplying together two numbers obtained in separate operations of measurement. *Fundamental quantities* are those quantities that are not defined in terms of other quantities. In physics there are 7 fundamental quantities of measurements namely length, mass, time, temperature, electric current, amount of substance and luminous intensity.

Derived physical quantities

Quantities which are defined in terms of the fundamental quantities via a system of quantity equations are called *derived quantities*. Examples of derived quantities include area, volume, velocity, acceleration, density, weight and force.

The SI units of derived quantities are obtained from equations using mathematical expressions

Note that some derived units have been given special names. For example, force is measured in kg m/s^2 and has been given a named unit called a newton (N).

1.2.2. International system of units (SI)

In order to measure any quantity, a *standard unit* (base unit) of reference is chosen. The standard unit chosen must be unchangeable, always reproducible and not subject to either the effect of aging and deterioration or possible destruction.

In 1960, an international system of units was established. This system is called the *International System of Units* (SI).

The *International System of Units* is an internationally agreed metric system of units of measurement. The value of a physical quantity is usually expressed as the product of a *number* and a *unit*.

Table 1: Common SI base units

Quantity	symbol	SI unit	Unit symbol
length	l	metre	m
mass	m	kilogram	kg
time	t	second	s
Electric current	I	ampere	A
Thermodynamic temperature	T	kelvin	K
Amount of substance	n	mole	mol
Luminous intensity	I_v	candela	cd

Name, Symbol and factor of metric prefixes in everyday use at workplace.SI prefixes used to form decimal multiples and submultiples of SI units (table 2 below).

Table 2: SI prefixes

Name	Symbol	Factor
deci	d	10^{-1}
centi	c	10^{-2}
milli	m	10^{-3}
micro	μ	10^{-6}
nano	n	10^{-9}
pico	p	10^{-12}
femto	f	10^{-15}
anto	a	10^{-18}

Name	Symbol	Factor
deca	da	10^{1}
hecto	h	10^{2}
kilo	k	10^{3}
mega	M	10^{6}
giga	G	10^{9}
tera	T	10^{12}
peta	P	10^{15}
exa	E	10^{18}

Example for length

- 10 mm= 1cm
- 1m= 106μm
- 1m=10-9Gm
- 1m2=(1012pm)2=1024pm2

Note: Numbers in the SI system are based on the number 10. Units in the SI system can therefore be multiplied or divided by 10 to form larger or smaller units.

1.2.3. Measuring fundamental physical quantities

Measuring length and distance

We use different tools for measuring length: metre rule, ruler, tape measure, vernier caliper and the micrometer screw gauge based on the kind of length to measure. Straight distances that are less than one metre in length are generally measured using metre rules. Straight distances that are more than one metre in length are generally measured using tape measure.

A tape measure or measuring tape is a flexible ruler and used to measure distance. A tape measure is in form of a strip of metal, plastic or cloth that has numbers marked on it as shown in figure below and is used for measuring. The instruments A and B in the figure 1.4. below represent examples of tape measures:

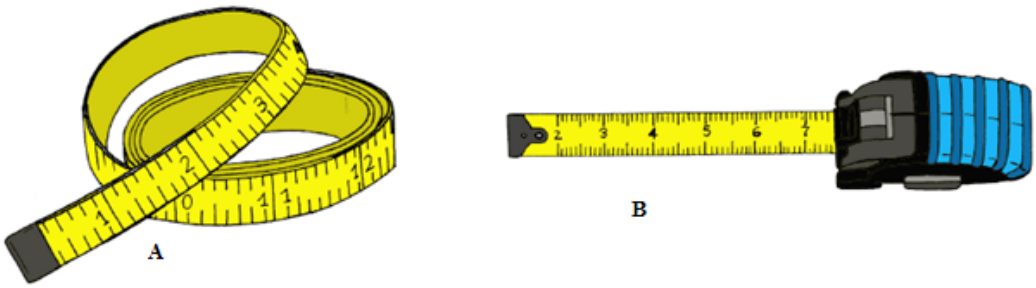


Figure 1.2: Tape measures used in measuring length or distance

It is a common measuring tool purposely designed to allow for a measure of great length to be easily carried out and permits one to measure around curves or corners. Surveyors use tape measures in lengths of over 100 m.

Metre rules are graduated in millimetres *(*mm*)*. Each division on the scale represents 1 mm unit (Fig 1.5. below).

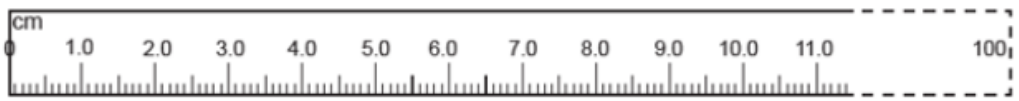


Figure 1.3: A metre rule

The direct way to measure length is by means of the straight edge of a ruler or metre ruler.

The ruler is placed alongside the object to be measured, and the number of unit intervals of the ruler equal to the length of the object is then noted.

Metre rule is used to measure lengths up to about 100 cm and has a sensitivity of 0.5 mm. Vernier calipers is an instrument used to measure outer dimensions of objects inside dimensions and depths.The figure 1.6 shows the vernier calipers:

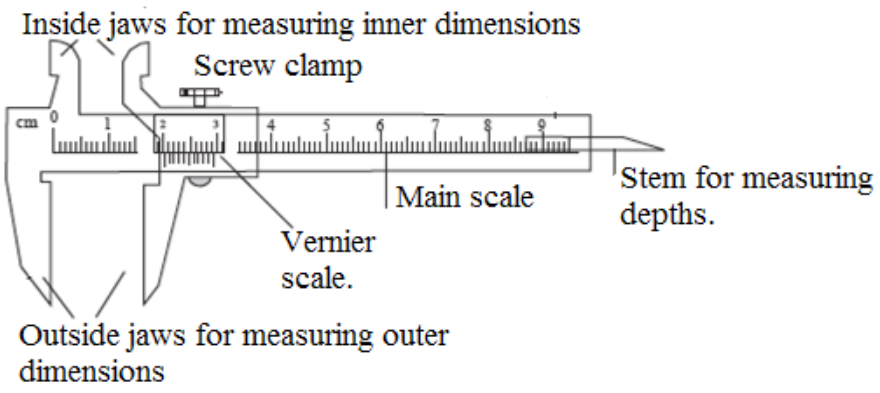


Figure 1.4.Vernier caliper

We can measure outer dimensions of objects (using the main jaws), inside dimensions (using the smaller jaws at the top), and depths (using the stem). The vernier calipers have a main scale and a sliding vernier scale that can allow readings to the nearest 0.02 mm.

To measure outer dimensions of an object, the object is placed between the jaws, which are then moved together until they secure the object.

The screw clamp may then be tightened to ensure that the reading does not change while the scale is being read.

The first significant figures are read immediately to the left of the zero of the vernier scale and the remaining digits are taken as the vernier scale division that lines up with any main scale division. The internal diameter of the test tube is given by (MSR + (VC × LC) Whereby the main scale reading (MSR), the vernier coincidence (VC) and The smallest reading called the least count (LC) that can be read from vernier callipers is 1 mm – 0.9 mm = 0.1 mm or 0.01 cm .

The main scale called the vernier coincidence *(VC)* and multiplying it with the least count i.e 0.01 cm. Therefore, the external diameter of the cylindrical object is (MSR + (VC × LC)

A micrometer screw gauge is an instrument for measuring very short length such as the diameters of wires, thin rods, and thickness of a paper.

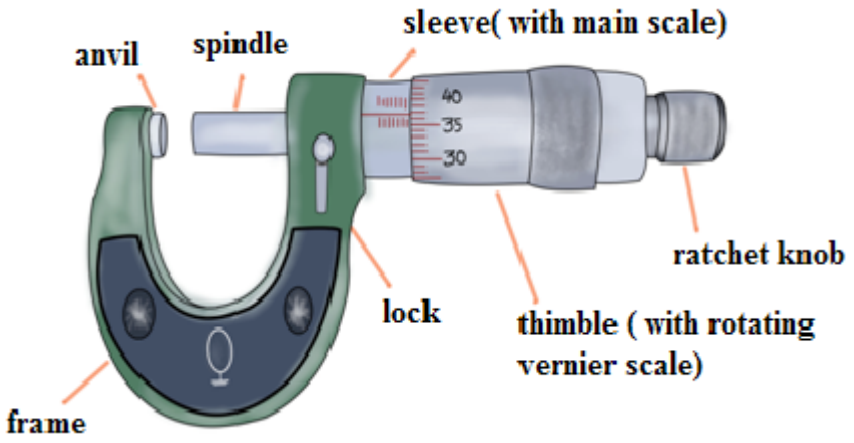


Fig 1.5. A screw gauge

The micrometers have a pitch of 0.50 mm (two full turns are required to close the jaws by 1.00 mm). The rotating thimble is subdivided into 50 equal divisions. The thimble passes through a frame that carries a millimetre scale graduated to 0.5 mm. Thimble, which has a circular rotating scale that is calibrated from 0 to either 50 or 100 divisions. This scale is called the head scale (thimble scale). When the thimble is rotated, the spindle can move either forward or backwards. Ratchet which prevents the operator from exerting too much pressure on the object to be measured. The least count = 0.01 mm. The micrometer screw gauge reading = MSR + (HSC × LC).

When the pitch is 1 mm, the thimble has 100 divisions called head scale divisions. In this case each division represents 0.01 mm. This is the least count (LC) of this screw gauge.

The thimble reading called the head scale coincidence (HSC) is the value of the mark on the thimble that coincides with the horizontal line on the sleeve. Main scale reading is taken by considering the reading of a mark on the fixed scale that is immediately before the sleeve enters the rim of the head scale.

The jaws can be adjusted by rotating the thimble using the small ratchet knob. This includes a friction clutch which prevents too much tension being applied. The thimble must be rotated through two revolutions to open the jaws by 1 mm.

In order to measure an object, the object is placed between the jaws and the thimble is rotated using the ratchet until the object is secured. The ratchet knob must be used to secure the object firmly between the jaws, otherwise the instrument could be damaged or give an inconsistent reading. The lock may be used to ensure that the thimble does not rotate while you take the reading.

Measuring mass

The mass of an object can be measured using a beam balance and a set of standard masses. It is noticed that the volume of the displaced water in measuring cylinder is equal to volume of an object lowered in the cylinder. There are many kinds of balances used for measuring mass illustrated below:

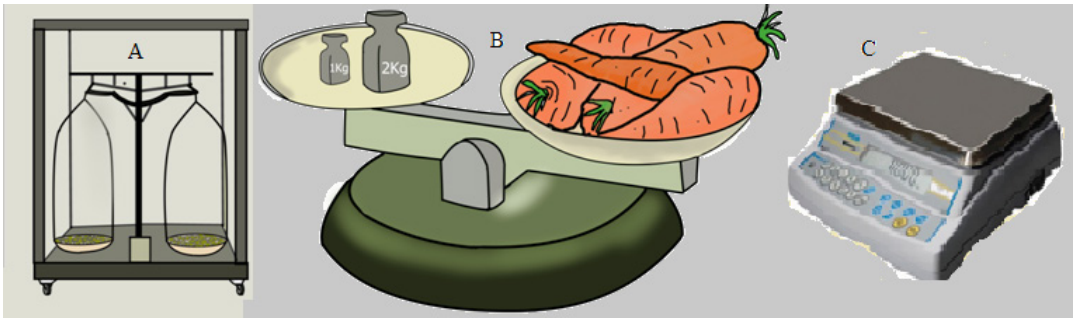


Fig 1.6: Different types of measuring instruments.

Measuring time

Time is measured using either analogue or digital watches and clocks and illustrated in figure below:

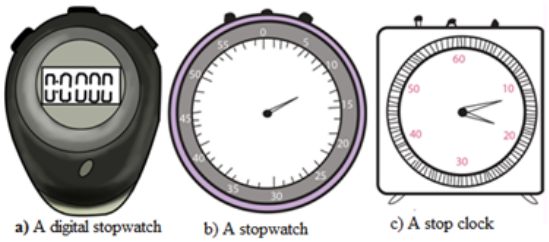


Figure 1.7: Stop watches and a stop clock

Application activity 1.2

1. Mention the appropriate instruments you would use to measure each of the following:
   a. The length of a football field.
   b. The mass of an object.
   c. The circumference of your waist.
   d. The time someone uses to cover a certain length.
   e. The diameter of a small ball.

2. It is possible to read and record the readings using a scale of a vernier caliper in order to measure the external diameter of the rod.

   Steps followed in using vernier

   a. Place the object to be measured between the outside jaws as shown in the figure below. Slide the jaw until they touch the rod.

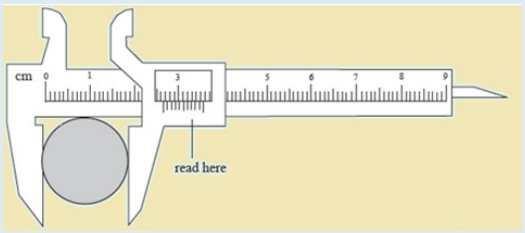


Figure 1.8: Measurement of external diameter using vernier caliper

   b. Record the readings on the main scale and the vernier scale. The main scale reading is the mark on the main scale that is immediately before the zero mark of the vernier scale.

   c. Multiply the vernier scale reading by 0.01 cm.

   d. Add the main scale reading (in cm) and the vernier scale reading (in cm) to get the diameter of the rod.

3. What is the diameter of the ball bearing shown in Figure below?

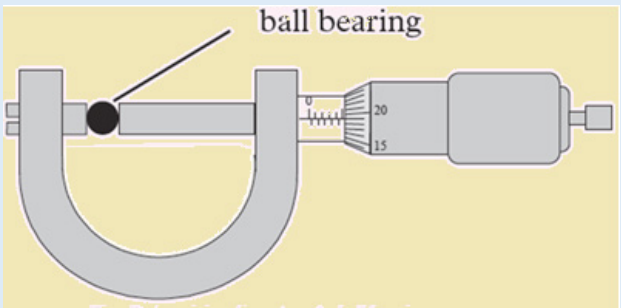


Fig 1.9: Determining diameter of a ball bearing

4.

   a). What does S I Units stands for?

   b). Explain why it is correct to say that SI units are very important in measurements?

   c). Suppose you wish to know the length of a big garden. How do you get the length of your garden?

d. Look at the following physical quantities: Mass, density, length, and time. Do all these quantities represent the fundamental quantities? Justify your decision by identifying the ones included in the category mentioned above.

4. Look at the table below and try to complete it based on the skills gained in the previous activities done;

Physical quantity	Name of the unit	Symbol of the S I units.
length		m
mass	kilogram	.....
.....	second	.....
volume	.....	m^3
area	.....	m^2

5. Choose two physical quantities with which you are familiar. Imagine that you are skilled in physical quantities and its measurements. Explain briefly how the values of these quantities can be obtained?
6. Express the following the indicated units and fill in blank spaces:

a). 250 m in …..cm.

b). 320 mg in ……g.

c). 5µg in ……g.

d). 7200 cm in …..m.

e). 3 kg in …… g.

1.3. Dimensions of physical quantities

Activity 1.3

Given the formulas for the following derived quantities, try go get the dimensions of each quantity.

a). velocity = displacement/time

b). acceleration = change of velocity/time

c). momentum = mass x velocity

d). force = mass x acceleration

e). work = force x displacement

1.3.1. Introduction to dimensions of physical quantities

The nature of physical quantity is described by nature of its dimensions. When we observe an object, the first thing we notice is the dimensions.

In fact, we are also defined or observed with respect to our dimensions that is, height, weight, the amount of flesh. The dimension of a body means how it is relatable in terms of base quantities. When we define the dimension of a quantity, we generally define its identity and existence. It becomes clear that everything in the universe has dimension, thereby it has presence.

Note: Dimensions are responsible in defining shape of an object.

1.3.2. Definition of dimensions of physical quantities

The dimension of a physical quantity is defined as the powers to which the fundamental quantities are raised in order to represent that quantity. The seven fundamental quantities are enclosed in square brackets [ ] to represent its dimensions.

Examples of assigning dimensions to physical quantities

Dimension of Length is described as [L], the dimension of time is described as [T], the dimension of mass is described as [M], the dimension of electric current is described as [A] and dimension of the amount of quantity can be described as [mol]. Adding further dimension of temperature is [K] and that dimension of luminous intensity is [Cd]

Consider a physical quantity Q which depends on base quantities like length, mass, time, electric current, the amount of substance and temperature, when they are raised to powers a, b, c, d, e, and f. Then dimensions of physical quantity Q can be given as:

$[Q] = [L^aM^bT^cA^dmol^eK^f]$

It is mandatory for us to use [ ] in order to write dimension of a physical quantity. In real life, everything is written in terms of dimensions of mass, length and time. Look out few examples given below:

1. **The volume** of a solid is given is the product of length, breadth and its height. Its dimension is given as:

   Volume = Length × Breadth × Height

   Volume = [L] × [L] × [L] (as length, breadth and height are lengths)

   Volume = $[L]^3$

As volume is dependent on mass and time, the powers of time and mass will be zero while expressing its dimensions i.e. $[M]^0$ and $[T]^0$

The final dimension of volume will be $[M]^0[L]^3[T]^0 = [M^0L^3T^0]$

2. In a similar manner, dimensions of **area** will be $[M]^0[L]^2[T]^0$
3. **Speed** of an object is distance covered by it in specific time and is given as:

   Speed = Distance/Time

   Dimension of Distance = [L]

   Dimension of Time = [T]

   Dimension of Speed = [L]/[T]

   $[Speed] = [L][T]^{-1} = [LT^{-1}] = [M^0LT^{-1}]$
4. **Acceleration** of a body is defined as rate of change of velocity with respect to time, its dimensions are given as:

   Acceleration = Velocity / Time

   Dimension of velocity = $[LT^{-1}]$

   Dimension of time = [T]

   Dimension of acceleration will be = $[LT^{-1}]/[T]$

   $[Acceleration] = [LT^{-2}] = [M^0LT^{-2}]$

Fundamental Quantity	Dimension
Length	[L]
Mass	[M]
Time	[T]
Temperature	[K]
Electric Current	[A]
Luminous Intensity	[Cd]
Amount of substance	[mol]

5. **Density** of a body is defined as mass per unit volume, and its dimension is given as:

   Density = Mass / Volume

   Dimension of mass = [M]

   Dimension of volume = $[L^3]$

Dimension of density will be = [M] / [L^3]

[Density] = [ML^{-3}] or [$ML^{-3}T^0$]

6. Force applied on a body is the product of acceleration and mass of the body

Force = Mass × Acceleration

Dimension of Mass = [M]

Dimension of Acceleration = [LT^{-2}]

Dimension of Force will be = [M] × [LT^{-2}]

[Force] = [MLT^{-2}]

1.3.3. Rules for writing dimensions of a physical quantity

We follow certain rules while expression a physical quantity in terms of dimensions, they are as follows:

- Dimensions are always enclosed in [ ] brackets
- If the body is independent of any fundamental quantity, we take its power to be 0
- When the dimensions are simplified we put all the fundamental quantities with their respective power in single [ ] brackets, for example as in velocity we write [L][T]$^{-1}$ as [LT^{-1}]
- We always try to get derived quantities in terms of fundamental quantities while writing a dimension.
- Laws of exponents are used while writing dimension of physical quantity so basic requirement is a must thing.
- If the dimension is written as it is we take its power to be 1, which is an understood thing.
- Plane angle and solid angle are dimensionless quantity that is they are independent of fundamental quantities.
- Therefore, some of the examples of dimensions of physical quantities include the following:

  Force, [F] = [MLT^{-2}]

  Velocity, [v] = [LT^{-1}]

  Charge, (q) = [AT]

Specific heat, (s) = $[L^2T^2K^{-1}]$

Gas constant, [R] = $[ML^2T^{-2}K^{-1}\ mol^{-1}]$

Benefits of Dimensions

Before writing dimensions of a physical quantity, it is must know a thing to understand why do we need dimensions and what are benefits of writing a physical quantity. Benefits of describing a physical quantity are as follows:

- Describing dimensions help in understanding the relation between physical quantities and its dependence on base or fundamental quantities, that is, how dimensions of a body rely on mass, time, length, temperature and others.
- Dimensions are used in dimension analysis, where we use them to convert and interchange units.
- Dimensions are used in predicting unknown formulae by just studying how a certain body depends on base quantities and up to which extent.
- It makes measurement and study of physical quantities easier.
- We are able to identify or observe a quantity just because of its dimensions.
- Dimensions define objects and their existence.

Limitations of Dimensions

Besides being a useful quantity, there are many limitations of dimensions, which are as follows:

- Dimensions can't be used for trigonometric and exponential functions.
- Dimensions never define exact form of a relation.
- We can't find values of certain constants in physical relations with the help of dimensions.
- A dimensionally correct equation may not be the correct equation always.

Dimension Table

It consumes a lot of time while deriving dimensions of quantities. So in order to save time, we learn some basic dimensions of certain quantities like velocity, acceleration, and other related derived quantities.

For Example, suppose you're asked to find dimensions of Force and you remember dimension of acceleration is $[LT^{-2}]$, you can easily state that the

dimension of force as [MLT^{-2}] as force is the product of mass and acceleration of a body.

The table below depicts dimensions of several derived quantities which one can use directly in problems of dimension analysis.

Quantity	Formulae	Dimensions
Velocity	Displacement/Time	[LT^{-1}]
Acceleration	Velocity/Time	[LT^{-2}]
Work done	Force × displacement	[ML^2T^{-2}]
Pressure	Force/ Area	[$ML^{-1}T^{-2}$]
Power	Work done/ Time	[ML^2T^{-3}]
Density	Mass / Volume	[$ML^{-3}T^0$]
Area	Length × Breadth	[L^2]

Application activity 1.3

1. i) What are four uses of dimensional analysis? Explain with one example for each.

   ii). What are three limitations of dimensional analysis in physics?

2. Show that $\frac{1}{2}$ gt2 has the same dimensions of distance.
3. What are the missing words in the following statements?

   a. The dimensions of velocity are …… .

   b. The dimensions of force are ……

4. a) What does the term dimension mean in Physical quantities?

   b) Given the formulas for the following derived quantities, calculate the dimensions of each quantity.

   i. Velocity = $\frac{Displacement}{Time}$

   ii. Acceleration = $\frac{\Delta Velocity}{Time}$

   iii.Momentum = mass x velocity

   iv. Force = mass x acceleration

   v. Work = force x displacement

Skills lab 1

Conduct a survey, collect and analyze data about when, where, and why people use different measuring instruments or devices and physical laws.

To complete this project you must:

- Develop a survey sheet about physical quantities, measuring instruments or devices, physical laws needed, appropriate SI units and metric prefixes used in everyday life.
- Distribute your survey sheet to other student-teachers, family members and neighbors.
- Compile and analyze your data.
- Create a report to display your findings in your sheet.

Plan it! To get started, think about the format and content of your survey sheet. Brainstorm what kinds of questions you will ask. Develop a plan for involving student-teachers in your class or other classes to gather more data.

End unit assessment 1

1. Differentiate a fundamental quantity and a derived quantity. Give one example of each and its corresponding SI units.
2. Express the following in millimetres:

   (a) 2.7 m (b) 26.9 cm (c) 356 μm.
3. What is the length of the glass rod shown in Figure below?

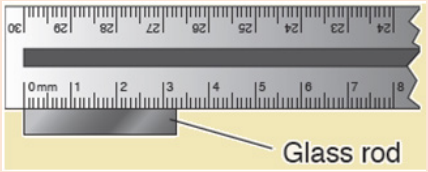


Figure 1.10: Measuring length of a glass rod

4. Use the knowledge and skills gained from the previous concepts to complete the following sentences:

   a) A quantity may be defined as any …….. in nature with which a number may be associated.

   b) Physical quantities are classified into …….. and …….. quantities.

c)……are those quantities that are not defined in terms of other quantities.

d) The value of a physical quantity is usually expressed as the product of a ……and a ……

e) The SI units stands for ……

5. Kaneza conducted an experiment on the growth of plants and recorded the results in a table. He used four plants of the same type and size and measured their growth after one month.

Table of results based on each plant type.

Plant	Amount of fertilizer (ml)	Frequency of fertilizer	Fertilizer Brand	Growth after 1 month (in cm)
1	2	Once per week	Brand A	3
2	2	Three times per week	Brand A	4
3	2	Once per week	Brand B	2
4	2	Twice per week	Brand B	6

Questions on scientific report above:

a) Identify the possible data types that were considered in Muneza's experiment?

b) Tell whether each data type was controlled effectively?

c) Explain what is wrong with Muneza's experiment?

d) What could he change to allow better conclusions to be drawn in the scientific investigation above?

UNIT 2:

INTRODUCTION TO BIODIVERSITY

Key Unit competence: Explain how biodiversity is threatened by climate change and human activities.

Introductory Activity 2

Study the figure below and answer the asked questions:

Figure 2.1: A community

a). List all living things that are available in figure 2.1.

b). Explain how each organism in the figure can be affected by another.

c). Talk on how organisms in the figure can be affected by the physical environment in which they live.

d). What would you suggest to study in this unit?

2.1. Ecological terms: Species, Ecosystem, Niche, Population, Community and Biodiversity

Activity 2.1

Based on other biological concepts learnt in previous studies, give the meaning of the following ecological terms:

a) Species b) Ecosystem c) Niche d) Population e) Community f) Biodiversity?

Species is a group of closely related organisms which are capable of interbreeding to produce fertile offspring. Occasionally two organisms which are genetically closely related but not of the same species can interbreed to produce infertile offspring. For example, a cross between a donkey and a horse, produces a mule, which is infertile. Hence, a donkey and a horse do not belong in the same species. Another example includes lions and tigers belonging in different species. However, when a male tiger mates with a female lion they can have fertile offspring called tiglons, although the offspring of female tigers and male lions called ligers are not fertile. Note that normally tigers are forest dwellers and lions are plains dwellers and they are ecologically isolated. Breeding has only been observed in captivity.

Ecological population is a group of individuals of the same species which live in a particular area at any given time.

Ecological community consists of populations of different species which live in the same place at the same time, and interact with each other.

A **Habitat** is a specific area or place in which an individual organism lives. When a habitat is very small it is regarded as a microhabitat.

Within the habitat, **an ecological niche** is the status or the role of an organism in its habitat or the mode of life of an organism within its habitats. For example, insects are pollinating agents and preys of insectivores.Two important aspects of a species' niche are the food it eats and how the food is obtained

The picture below is of birds that occupy different niches. Each species eats a different type of food and obtains the food in a different way.Example: Each of these species of birds has a beak that suits it for its niche. For example, the long slender beak of the nectarivore allows it to sip liquid nectar from flowers. The **short** sturdy beak of the granivore allows it to crush hard, tough grains.

Figure 2.2: Bird niches

In an environment, communities are influenced either by abiotic components, also called abiotic factors. These are the non-living physical aspects of the environment such as the sunlight, soil, temperature, wind, water, and air. Communities are also influenced by biotic components, or biotic factors. These are the living organisms in the environment.

An **ecosystem** is a collection of all the organisms that live together in a particular place, together with their nonliving, or physical environment.

A biome is a group of ecosystems that have the same climate and similar dominant communities.

The **biosphere** is the whole of the earth's surface, the sea and the air that is inhabited by living organisms. The highest level of organization is the entire biosphere.

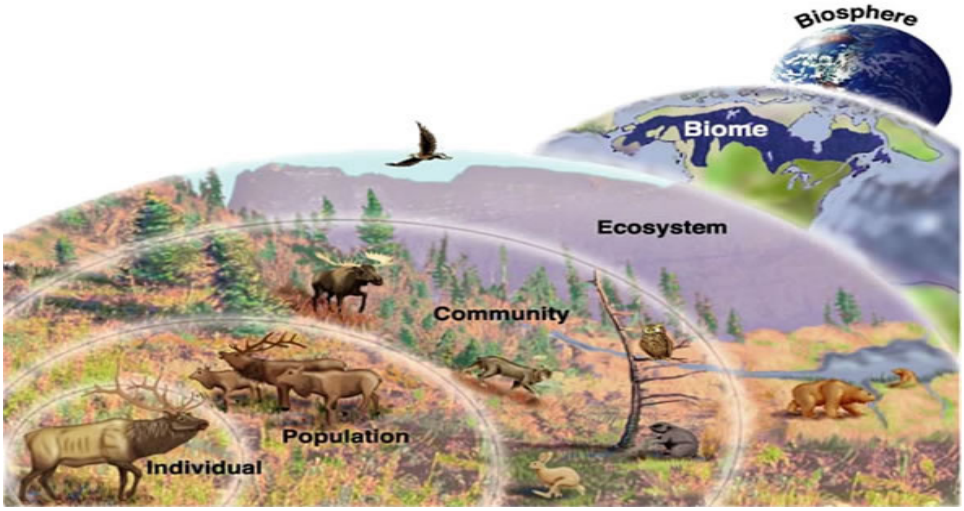


Figure 2.3: Ecological organizational levels

Biodiversity is defined as the full range of variety and variability within and among living organisms and the ecological complexes in which they occur. In other words, biodiversity is the variety of life. It refers to the totality of the species including the genetic variation represented in the species populations, across the full range of terrestrial organisms, including vertebrates and invertebrates, Protista, Bacteria and plants.

Importance of biodiversity

Biodiversity contributes to ecosystem goods and services. The ecosystem goods and services include:

- Provision of food, air, fire wood, medicines, energy, fresh water.
- Nutrient cycling such carbon, water and nitrogen cycles by microorganisms and primary production by photosynthesis.
- Cultural or aesthetic service recreation, ecotourism, cultural and religious inspiration.

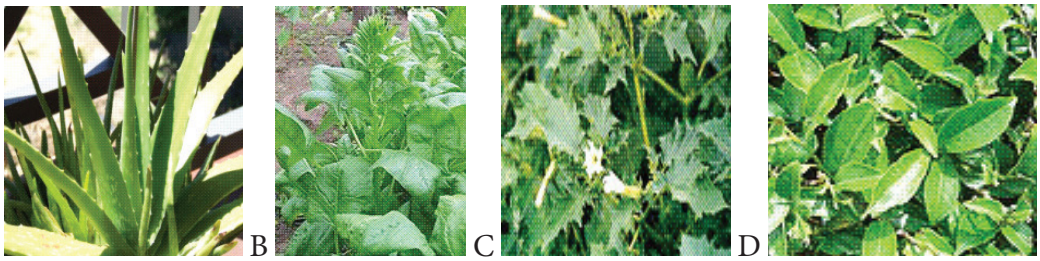


Figure 2.4: Some of the common medicinal plants (A: Aloe vera, B: Spinacia oleracea, C: Daturastramonium, D: Camellia sinensis)

Threats and consequences of biodiversity loss

Observe the figure below, and describe the human activities illustrated in them in relation to the biodiversity loss

Figure 2.5: Deforestation and bush burning cause biodiversity loss

Threats of biodiversity

The main causes of biodiversity loss can be attributed to the influence of human activities on ecosystems. Threats to biodiversity may include:

a). Habitat loss and the degradation of the environment

The habitat loss and the degradation of the environment occur in different ways.

The most occurring, are tree cutting, agriculture and fires. These human activities lead to the alteration and loss of suitable habitats for biodiversity. As a consequence, there is a loss of plant species as well as the decrease in the animal species associated to this plant diversity.

b). Introduction of invasive species and genetically modified organisms

Species originating from a particular area are harmful to native species also called endemic species when they are introduced into new natural environments. They can lead to different forms of imbalance in the ecological equilibrium, so that endemic species may fail to compete with introduced species, and they may affect the abundance and distribution in natural habitat.

c). Pollution

Human activities such as excessive use of fertilizers, and increased pollutants from industries and domestic sewage affect biodiversity. They contribute to the alteration of the flow of energy, chemicals and physical constituents of the environment and hence species may die as a result of toxic accumulation.

d). Overexploitation of natural resources

Increased hunting, fishing, and farming in particular areas lead to the decrease and loss of biodiversity due to excessive and continuous harvesting without leaving enough time for the organisms to reproduce and stabilize in their natural habitat.

e). Climate change

This is a change in the pattern of weather, related changes in oceans, land surfaces and ice sheets due to global warming resulting from man's activities. Increasing global temperatures have resulted into melting of icebergs raising sea levels and so flooding coastal areas eventually affecting the niche, and these may take the lives on many living things.

Consequences of loss of biodiversity

They are various consequences of loss of biodiversity that include:

- Desertification, is thought by scientists to be a consequence of climate change, has been considered to be related to deforestation. Disrupting water cycles and soil structure results into less rainfall in an area.
- Floods as a result of rising sea levels.

- Habitat destruction for extensive farming, timber harvesting and infrastructure and settlement.
- Decrease in food production as result of change in pattern of weather that affects productivity
- Large scale deforestation has a negative effect on nutrient recycling and can accelerates soil erosion.
- Diseases that come as effects of floods and malnutrition due to famine

Identification of biodiversity

Biodiversity can be categorized into three groups:

- **Genetic diversity:** The combination of different genes found within a population of a single species, and the patterns of variation found within different populations of the same species. These variations are caused by the gene mutations or chromosomal mutations which create differences in individuals of the same species.
- **Species diversity**:This is concerned with variation innumber of species and their relative abundance in an area in which they inhabit. All species are different from each other. These could be structural differences, such as the difference between a mango tree and a cow. They could also be functional differences, such as the differences between bacteria that cause decay and those that help us to digest food. The variation in the relative abundance of species within a habitat may be caused by different factors, mainly environmental factors which can affect their rate of reproduction.
- **Ecosystem diversity:** This is concerned with variations in ecosystems or habitats that occur within a region. Environmental factors like climate change may cause diversity of habitats or systems within a region.

Application activity 2.1

What do you understand by:

a). Genetic diversity

b). Species diversity

c). Ecosystem diversity

2.2. Determination of the distribution and abundance of organisms in an area

Activity 2.2

Observe the figures below, and relate the sampling technique with figure and discuss on methods used to determine distribution and abundance of organism in an area

(a)

(b)

Figure 2.6: Illustration of sampling techniques

Biologists use different sampling techniques when they want to determine the distribution and abundance of organisms in area. The commonly used techniques are:

a). Random sampling method

A random sampling method is a sampling method where samples are taken from different positions within a habitat and those positions are chosen randomly. Random sampling in important to avoid the bias.

b). Quadrat sampling method

A quadrat is a square area that is marked using a pre-made square of plastic, or stakes and string and it can range in size. Different species and their numbers within the quadrat are counted. Counting is repeated many times in different places in the habitat to get an accurate representation of biodiversity.

c). Frame quadrats

Frame quadrats are small plot used to isolate a standard unit of area for the study of the distribution of an item over a large area. While originally

rectangular, modern quadrats can be rectangular, circular, and /or irregular. The quadrat is suitable for sampling plants, slow-moving animals such as millipedes, insects and some aquatic organisms.

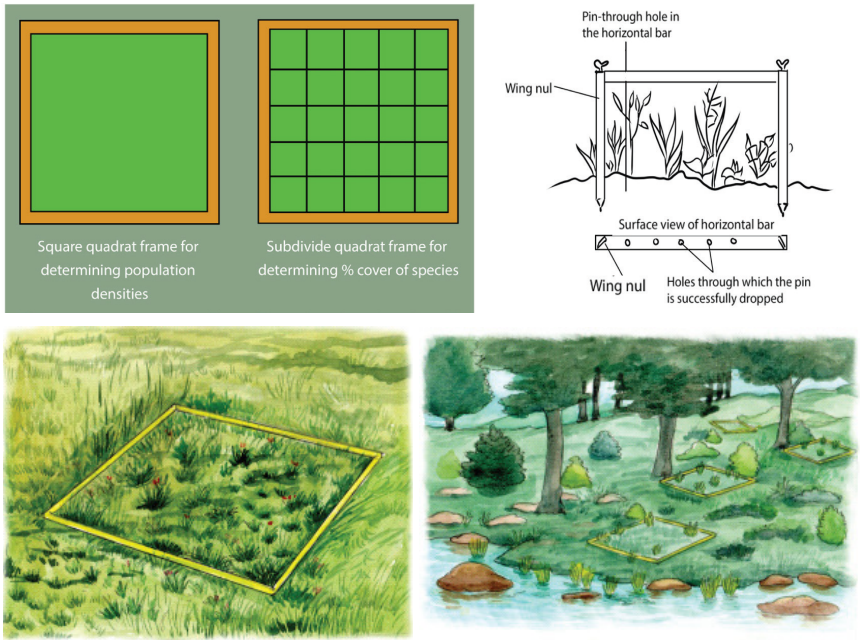


Figure 2.7: Different frame quadrats

d). Transect sampling

Transect sampling is done using a transect line, which is usually a rope or measuring tape that has been marked at set intervals, such as every meter. The line is unrolled within the habitat. At every interval, the type and number of species along the line are recorded. A measured line is laid across the area in the direction of the environmental gradient. The species touching the line can be recorded along the whole length of the line (continuous sampling) or at specific points along the line (systematic sampling).

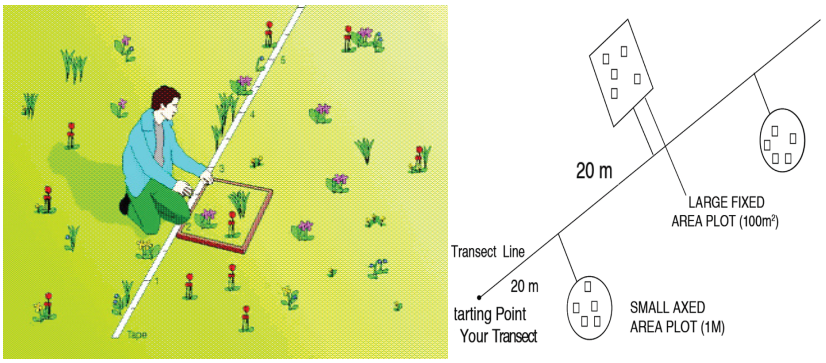


Figure 2.8: Line transect

e). Belt transects method

Belt transects method is the same as the line transects but widens the sampling area.. The samples are taken and the abundance, percentage cover in a defined area determined. Samples can be taken within the belt.

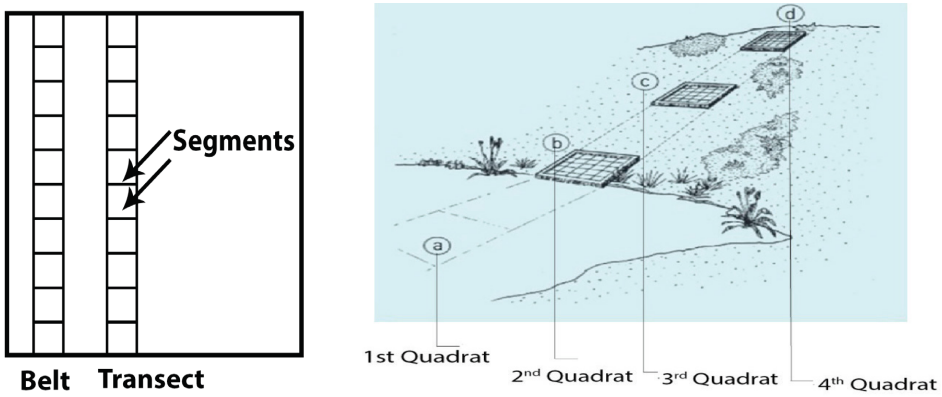


Figure 2.9: Belt transect

f). Netting

Netting is a sampling method where fine mesh nets are used to capture different organisms that include insects, birds and bats. The technique is also used for sampling small aquatic organisms like daphnia, and water boatman.

Figure 2.10: Netting method of sampling.

g). Capture -recapture technique

This method is useful for sampling non-fixed population and is suitable for animal such as fishes, birds, lizards and insects. A sample of the population to be studied is first captured and each individual is marked with a spot for identification. These are then released and given enough time to mix up with the rest of the members in the habitat. After a certain period of time, another sample is taken. During the mark-release-recapture technique, the total population can be estimated by the use of the formula:

$$\frac{n_1 \times n_2}{n_3}$$

where n_1 is a number caught and marked in first sample, n_2 is a number caught in second sample n_3 is a number in the second sample that had been marked.

To understand this application, let us use the following **example:**

A team of students used a sweep net to sample brown grasshoppers and each collect insect was marked with a very small spot of non-toxic waterproof paint and then they were released in the field. The next day, a second large sample was conducted and data were recorded as follows: number of caught and marked in first sample (n_1) = 247, number of caught in second sample (n_2) = 269, and the number in the second sample that had been marked (n_3) = 16. What is the number of estimated population?

Solution: The estimated number:

$$= \frac{n_1 x n_2}{n_3} = \frac{247x269}{16} = \frac{66443}{16}$$

= 4152 grasshoppers

Application activity 2.2

1. Explain the advantages of the random sampling techniques.
2. Use suitable methods, such as frame quadrats, line transects, and belt transects, to assess the distribution and abundance of insect species in a school garden. Record your data and use the Simpson index of diversity (D) to calculate the diversity of collected insects.
3. Suggest the benefits of using the following sampling techniques:
   a) Quadrats
   b) Transect
   c) Mark-capture-recapture
4. State the conditions in which quadrats, transect and mark recapture are suitable sampling methods.

2.3. Spearman's rank, Pearson's linear correlation and Simpson's Index of Diversity

2.3.1. Spearman's rank, Pearson's linear correlation

Activity 2.3.1

A student reads a questionnaire and one question was difficult for him to answer. The question asked to explain the process of Spearman's ranking and Pearson's linear correlation. As the one who is learning biology, assist him to answer correctly such question by doing research in different necessary resources and prepare complete related information in your exercise.

To decide if there is an association between collected data, a correlation coefficient is calculated and plot scatter graph drawn in order to make a judgment. The strongest correlation is present for studied items when all the points lie on a straight line. In this case, there is linear correlation, and the correlation coefficient equals 1.

If a given variable X increases so does another variable Y, the relationship is a positive correlation. If a variable X increases while the variable Y decreases, then the relationship is a negative correlation. A correlation coefficient of 0 means that there is no correlation at all. These correlation coefficients are ways to test a relationship observed and recorded to see if the variables are correlated and, if so, to find the strength of that correlation.

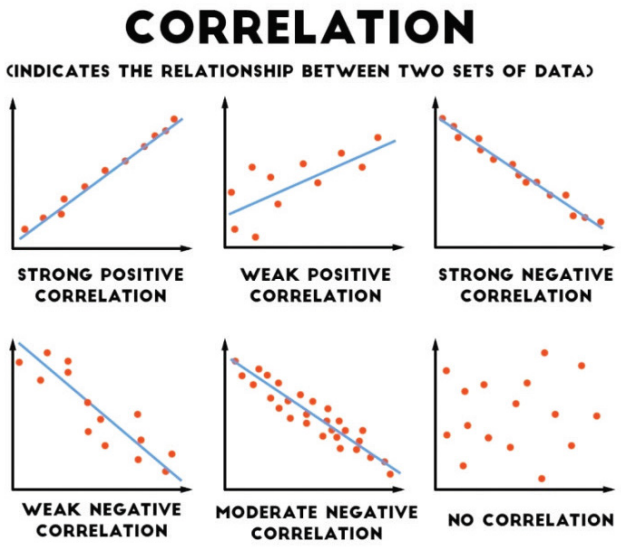


Figure 2.11: Forms of correlation

a). Spearman's rank correlation coefficient

When collected data are not quantitative, but used an abundance scale or when the researcher is not sure if quantitative data are normally distributed. It might also be possible that a graph of results shows that the data are correlated, but not in a linear fashion. In this case, the Spearman's rank correlation coefficient is used. It involves ranking the data recorded for each variable and assessing the difference between the ranks. You should always remember that correlation does not mean that changes in one variable cause changes in the other variable.

b). Pearson's correlation coefficient

Pearson's correlation coefficient can only be used where there might be a linear correlation and when there are collected quantitative data as measurements (for example, length, height, depth, and light intensity, mass) or counts (for example number of plant species in quadrats). The data must be normally distributed.

$$r = \frac{\sum xy - n\bar{x}\bar{y}}{ns_x s_y}$$

Where:

r is the correlation coefficient

x is the number of species in a quadrat

y is the number of species in the same quadrat

n is the number of readings (From1 to n)

$\bar{x}$ is the mean number of species

$\bar{y}$ is the mean number of species

s_x is the standard deviation for x

s_y is the standard deviation for y

2.3.2. Simpson's Index of Diversity (D)

Activity 2.3.2

Visit your smart classroom and download and analyze the videos that explains

a). The purpose of calculating the Simpson diversity index.

b). How to calculate Simpson index D, Simpson index of diversity and Simpson reciprocal index.

The Simpson diversity index is among indices used to measure diversity. It is expressed in three related indices namely Simpson index, Simpson index of diversity and Simpson reciprocal index.

a). Simpson index D

Simpson index D can be expressed in two ways and takes into consideration the total number of organisms of a particular species and the total number of organisms of all species. It is calculated as follows:

$$D = 1 - \sum (n/N)^2$$

with n: the total number of organisms of a particular species and N: the total number of organisms of all species. When the index equals or is nearby 0 there is an infinite diversity of considered species. When it equals or is nearby 1, this means that there is no diversity. The bigger the value of D, the lower the diversity and small is D, bigger is the diversity.

b). Simpson index of diversity 1 - D

The value of this index also ranges between 0 and 1, but now, the greater the value, the greater the sample diversity. This makes more sense. In this case, the index represents the probability that two individuals randomly selected from a sample will belong to different species.

c). Simpson reciprocal index 1 / D

Another way of overcoming the problem of the counter-intuitive nature of Simpson's index is to take the Simpson's reciprocal index 1/D. The value of this index starts with 1 as the lowest possible figure. This figure would represent a community containing only one species. The higher the value, the greater the diversity.

Example:

1. In a woodland, a quadrat was sampled for ground vegetation. Data collected were recorded in the table 1.3.2. Find out the value of the Simpson index and draw the conclusion about the biological diversity of the sampled area.

Table 2.1: Recorded data on the vegetation from a woodland

Species	Number (n)
Woodrush	2
Holly (seedlings)	8
Bramble	1
Yorkshire Fog	1
Sedge	3
Total (N)	**15**

Solution:

Start by calculating all parameters in the formula for Simpson's Index

$D = 1 - \Sigma(\frac{n}{N})^2$ in the following table:

Species	Number (n)	n/N	$(n/N)^2$
Woodrush	2	0.133333	0.017777778
Holly (seedlings)	8	0.533333	0.284444444
Bramble	1	0.066667	0.004444444
Yorkshire Fog	1	0.066667	0.004444444
Sedge	3	0.2	0.04
Total	N = 15		$\Sigma (n/N)^2 =$ 0.351111111

Putting the figures into the formula for Simpson's Index and replace each letter by its respective value, the Simpson index shall be: **D = 1 – 0.3511111= 0.648888889**

Based on the theory above, the quadrat presents a higher diversity because the value of D is nearby zero

Application activity 2.3

1. In which conditions results can you conclude that there is:
   a) A positive correlation?
   b) A negative correlation?
   c) Non-correlation.
2. Explain the difference between species richness and species evenness
3. Suggest what precautions you may need to take when measuring populations of aquatic animals or plants.
4. explain why a habitat with high diversity tends to be more stable than one with lower diversity.
5. In a survey of trees in a tropical forest, students identified five tree species (A to E). They counted the numbers of trees in an area 100m×100m and found these results:

Tree species	Number of individual species
A	56
B	48
C	12
D	6
E	3

Calculate the Simpson's Index diversity for identified species and explain the advantage of using data on species diversity and abundance when calculating an index of diversity.

6. The Simpson's Index of diversity for vegetation in an open area inhabited by grasslands was 0.8. For a similar sized area of vegetation beneath some conifer trees it was 0.2. What do you conclude from these results?

Skills lab 2

Visit your school garden and sample into two areas A and B each with 25 cm^2 by using one of the sampling technics. Calculate the Simpson's diversity index for the two samples, and indicate which one is more diverse than another. For each sample, consider 5 more abundant species only.

End unit assessment 2

1. Explain what is meant by a habitat.
2. Make a list of all the habitats you can see in your school compound.
3. Explain why we share so many of our genes with plants.
4. Discuss the contribution of ecosystems to cultural traditions in Rwanda.
5. Pollution is one of the causes of aquatic biodiversity loss:
   a) What do you understand by water pollution?
   b) Outline human activities that contribute to water pollution
   c) Discuss how polluted water affects aquatic living organisms?
6. Relate desertification with biodiversity loss.
7. Discuss on importances of biodiversity.

8. Distinguish between:
   a) Community and population
   b) Ecological niche and habitat
9. Describe the two main components of an ecosystem.
10. Calculate the value of Simpson's Diversity Index (D) for a single quadrate sample of ground vegetation in woodland from which the following sampling date was obtained, and conclude about the diversity of plants in the woodland:

Collected specimens	Number of individual species (n)
Potato	2
Maize	8
Mango tree	1
Banana	1
Beans	3

UNIT 3:

INTRODUCTION TO CLASSIFICATION

Key Unit competence: Apply the basic knowledge of classification to group living organisms in three domains

Introductory Activity 3

Observe organisms in the figure below and answer asked questions:

a). Group the organisms in figure above in seven groups (two by two) based on how they are closely related.

b). Explain how their features have helped you to classify them in their correct groups.

3.1.Taxonomic hierarchy and Three domains of life: Archaea, Bacteria and Eukarya

3.1.1.The taxonomic hierarchy

Activity 3.1.1

You are provided with cards written on a list of words such as continent, district, country, cell, province, sector, village and family.

1. Arrange the above words in increasing size
2. What is your opinion about the people of the same family and those in the whole country?
3. Compare your arrangement in 1. above with possible groupings of organisms in biological taxonomic hierarchy.

Taxonomy is the study of classification of living organisms in taxonomic levels called taxa (singular: taxon). In biological classification, these taxa form a hierarchy. Each kind of organism is assigned to its own species, and similar species are grouped into a genus (plural: genera). Similar genera are grouped into a family, families into an order, orders into a class, classes into a phylum (plural: phyla) and phyla into a kingdom. The hierarchy classification starts from the largest group, the domain.

The eight taxonomic levels of classification are known as taxa (taxon in singular), these include: **Domain, Kingdom, phylum, class, order, family, genus** and **species**. As one moves down the taxonomic hierarchy, it follows that the number of individuals decreases but the number of common features increases.

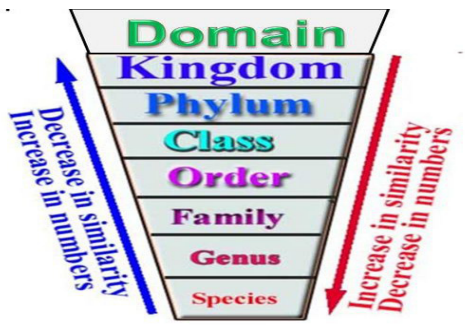


Figure 3.1: Levels of classification

Table 3.1 below shows how human being, earthworm and hibiscus are classified.

Table 3.1: Classification of human being, earthworm and hibiscus

Taxa	Human	Earthworm	Hibiscus
Domain	Eukarya	Eukarya	Eukarya
Kingdom	Animalia	Animalia	Plantae
Phylum	Chordata	Annelida	Angiospermae
Class	Mammalia	Oligochaeta	Dicotyledonae
Order	Primata	Terricolae	Malvales
Family	Homonidae	Lumbricidae	Malvaceae
Genus	Homo	Lumbicus	Hibiscus
Species	*H. sapiens*	*L. terrestris*	H. rosa sinesis

Application activity 3.1.1

1. What are the three domains of living things?
2. Discuss the ways in which a domain differs from a kingdom?

3.1.2. Three domains: archaea, bacteria and eukarya

Activity 3.1.2

Visit the computer lab and search the characteristics archaea, bacteria and eukaryadomains and present them on manila paper.

There are three domains used by biologists to divide organisms into three large groups based on their cell structure. The domain is the highest taxon in the hierarchy. The prokaryotes are divided between the domains Eubacteria and Archaebacteria, while all the eukaryotes are placed into the domain Eukarya.

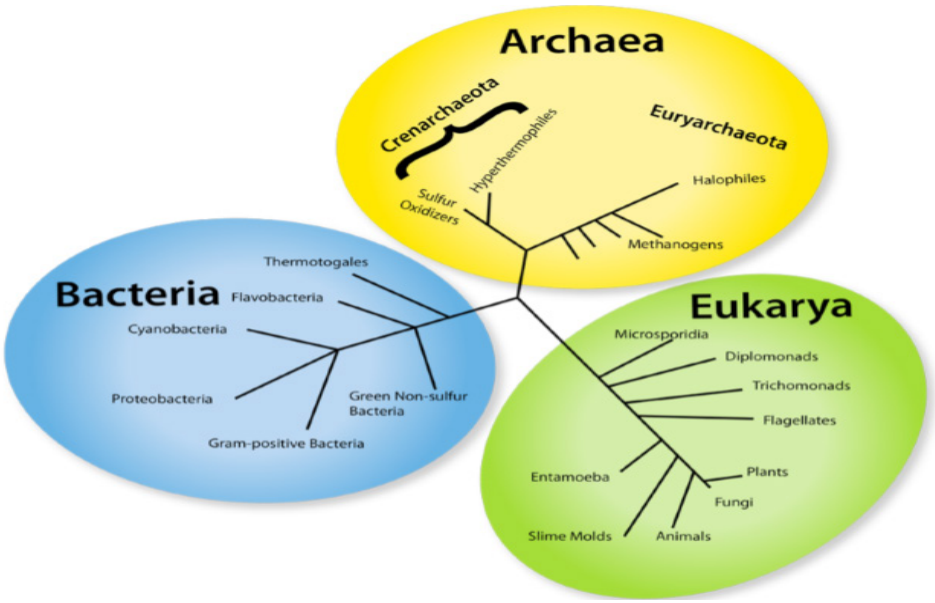


Figure 3.2: The three domains of organisms

a). Domain Eubacteria

Domain bacteria include prokaryotic organisms as their cells have no true nucleus.

They are all microscopic that vary in size between 0.2 to 10 micrometers. The characteristic features of bacteria are:

- Cells with no true nucleus
- DNA exists in circular chromosome and does not have histone proteins associated with it.
- No membrane-bound organelles (mitochondria, endoplasmic reticulum, Golgi body, chloroplasts)
- Contain mesosomes as infolding of membrane and acts as sites for respiration as they lack mitochondria.
- Ribosomes (70 S) are smaller than in eukaryotic cells
- Cell wall is always present and contains peptidoglycans in place of cellulose
- Cells divide by binary fission
- Usually exist as single cells or colonies.

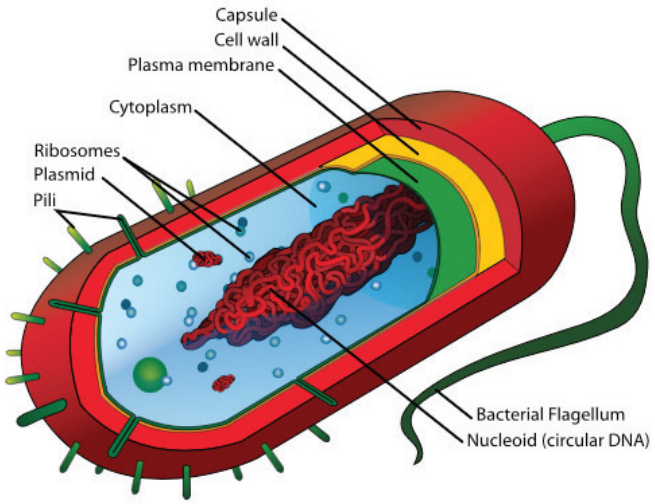


Figure 3.3: Structure of a typical bacteria

b). Domain Archaea (Archaebacteria)

This includes bacteria that live in extreme environments where few other organisms can survive, like in volcanic hot springs and black organic mud totally devoid of oxygen.

They are classified according to the environments they live in:

- Methanogenic bacteria that live in habitats deprived of oxygen and give off methane as a product of metabolism, for example those that live in the guts of ruminant animals.
- Halophilic bacteria live only in salty conditions
- Thermoacidophilic bacteria tolerate extreme acid and temperature that exceed boiling point of water and a pH below 2.

c). Domain Eukarya

All the organisms classified into this domain have cells with true nuclei and membrane-bound organelles. It includes the four remaining kingdoms: protists, fungi, plantae and Animalia. Their characteristic features are:

- Cells with a nucleus and membrane-bounded organelles
- linear DNA associated with histones arranged within a chromosome in the nucleus
- Ribosomes (80S) in the cytosol are larger than in prokaryotes, while chloroplasts and mitochondria have small ribosomes (70S ribosomes), like those in prokaryotes.

- Chloroplast and mitochondrial DNA is circular as in prokaryotes suggesting an evolutionary relationship between prokaryotes and eukaryotes
- A great diversity of forms: unicellular, colonial and multicellular organisms
- Cell division is by mitosis.
- Many different ways of reproduction including asexually and sexually.

Application activity 3.1.2

1. What are the six kingdoms of living things as they are now identified?
2. List three domains of living organisms.
3. Which kingdoms include only prokaryotes? Which kingdoms include only heterotrophs?
4. How do domains and kingdoms differ?
5. Suppose that you discover a new single-celled organism which has a nucleus, mitochondria and a giant chloroplast. In which kingdom would you place it? What are your reasons?
6. It is confirmed that: "Some bacteria can survive in extreme temperatures such as hot springs". Justify this statement.

3.2. Characteristic features of the kingdoms

Activity 3.2

Observe carefully the photo of organisms below and answer asked questions.

1. Identify the characteristics of organism found on photo.
2. Among other organisms you know, apart from that you are seeing on photo, search their characteristics and present them in your exercises book

Living organisms are classified in five kingdoms namely Protoctista, Fungi, Plantae, Monera and Animalia.

3.2.1. Characteristic features of the kingdom Protoctista

This kingdom is made up of a very diverse range of eukaryotic organisms, which includes those that are often called protozoans and algae. Living things such as paramecium, amoeba, euglena, algae and plasmodia belong to the kingdom Protoctista.

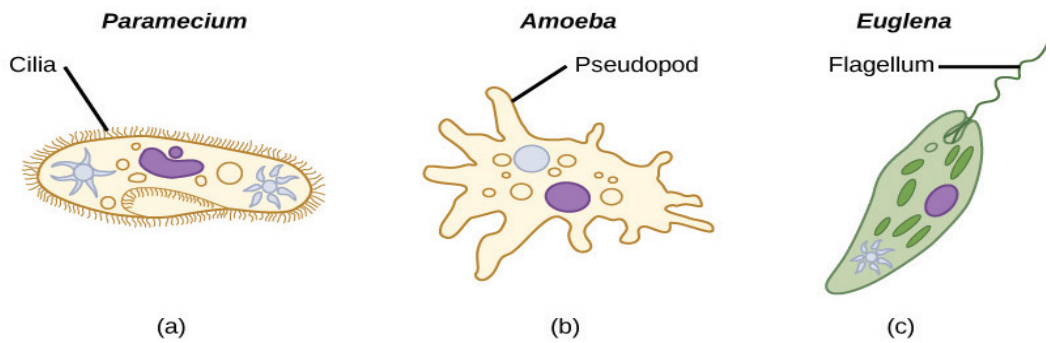


Figure 3.4: Structures of some of the examples of organisms in protoctista showing some common features: paramecium, Amoeba and Euglena.

The characteristic features of protoctists are listed according to the different phyla due to their diverse range:

- Rhizopus that have pseudopodia for locomotion. Example, amoeba.
- Flagellates which are protoctista which move by using flagella. Example, Trypanosoma.
- Sporozoans which are mainly parasitic organisms that reproduces by multiple fission. Example plasmodium.
- Ciliates are protoctista which move with cilia. Example paramecium.
- Euglenoid flagellates which are organisms with flagella but with a biochemistry quite distinct from that of flagellates. Example Euglena.
- Green algae are photosynthetic protoctista with chlorophyll pigments. Example chlorella.

- Red algae are photosynthetic protoctista with red pigment as well as chlorophyll. Example, chondrus
- Brown algae which are photosynthetic protoctista with brown pigments as well as chlorophyll. Example Fucus and sea weed.

3.2.2. Characteristic features of the kingdom fungi

Fungi are all heterotrophic, obtaining energy and carbon from dead and decaying matter or by feeding as parasites on living organisms. There is a vast range in size from the microscopic yeasts to macroscopic fungi.

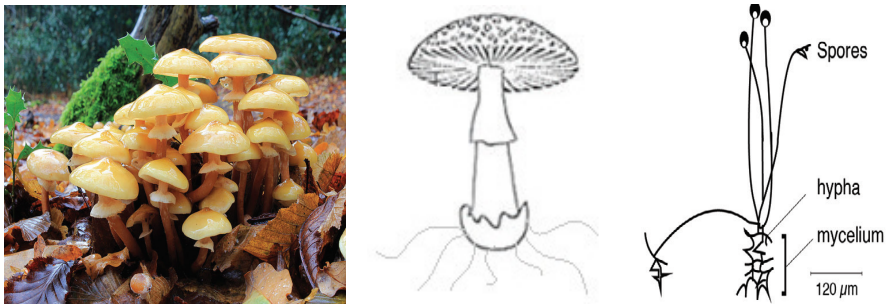


Figure 3.5: Forms of fungi

Other characteristic features of fungi are:

- Heterotrophic nutrition.
- They use organic compounds made by other organisms as their source of energy and source of molecules for metabolism.
- Reproduce asexually by means of spores and sexually by conjugation.
- Simple body form, which may be unicellular or made up of long threads called hyphae (with or without cross walls).
- Large fungi such as mushrooms produce large compacted masses of hyphae known as fruiting bodies to release spores.
- Cells have cell walls made of chitin or other substances.

3.2.3. Kingdom Plantae

Figure 3.6: Variety of plants

Plants are all multicellular photosynthetic organisms. They have complex bodies that are often highly branched both above and below the ground. Characteristic features of plants are:

- Multicellular eukaryotes with cells that are differentiated to form tissues and organs.
- Few specialized cells.
- Cells have large and often permanent vacuoles for support with cell walls made of cellulose.
- Most plants store carbohydrates as starch or sucrose.

3.2.4. Kingdom Animalia

Animals are multicellular organisms that are all heterotrophic with different methods of obtaining their food.

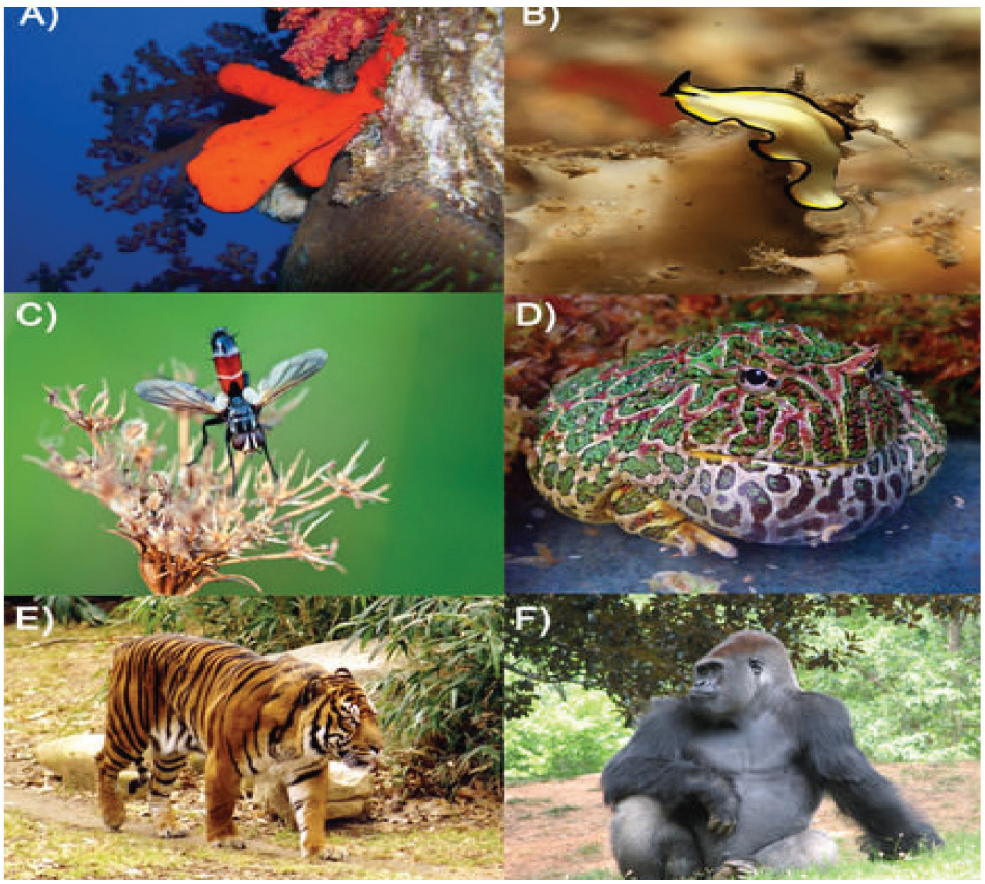


Figure 3.7: A variety of animals (A) Sponge, (B) Flatworm (C) Housefly (D) Frog (E) Tiger (F) Gorilla.

Organisms in this kingdom have other additional features:

- Different types of specialized cells.
- Cells do not have chloroplasts and cannot photosynthesize (although some, such as coral polyps have photosynthetic protoctists living within their tissues).
- Cell vacuoles are small and temporary (for example lysosomes and food vacuoles).
- Cells do not have cell walls.
- Communication is by the nervous system

3.2.5. Kingdom Monera

Organisms in this kingdom are unicellular, that do not have a nucleus. They are prokaryotic. They are the smallest and simplest organisms. Some of them stick together to form chains or clusters while others are single cells. The figure below shows a typical structure of a bacterial cell which contains all the main features of prokaryotes.

Although some of them are harmful in causing human diseases, others are beneficial species that are essential to good health, as they are involved in food industry, medicine and in pharmacy.

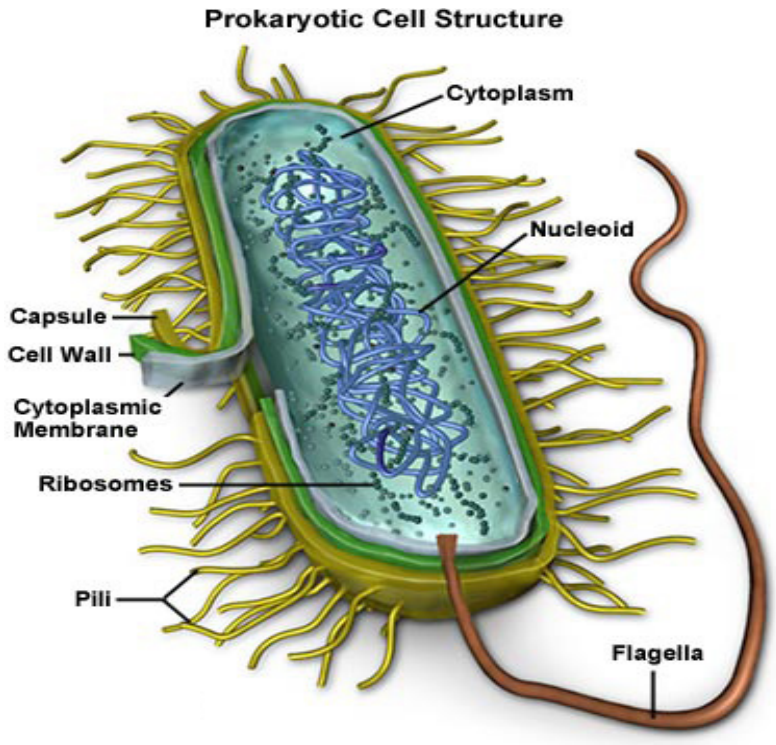


Figure 3.8: Procaryota cell structure

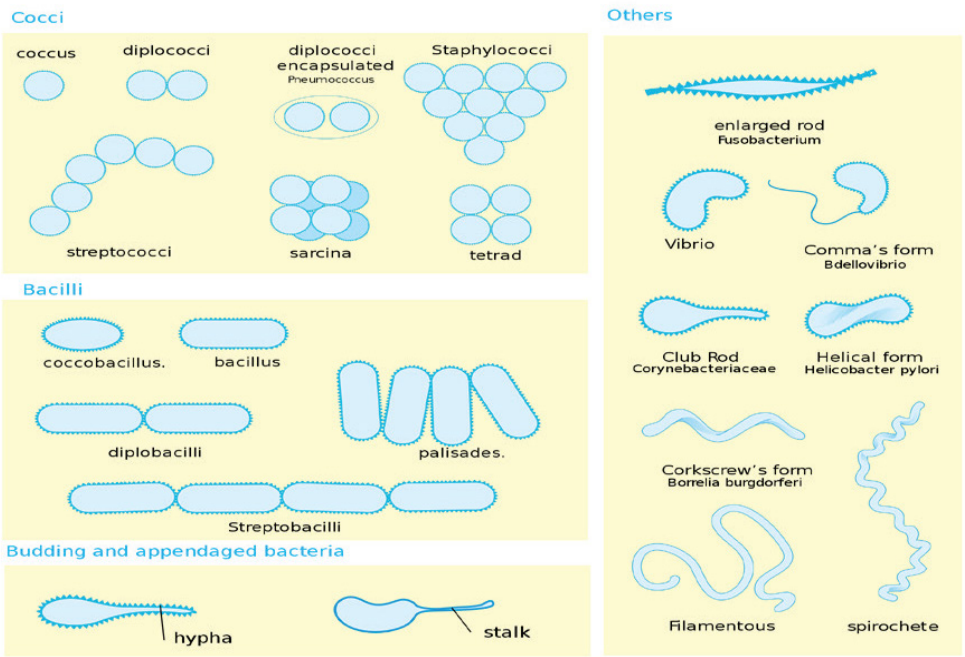


Figure 3.9: Structure of a typical bacterial and bacterial features

3.2.6.Common bacterial diseases in plants and animals

Activity 3.2.6

Suppose there is cholera outbreak in your village and the executive secretary invited you to sensitize people about preventive measures against cholera. Prepare a brief presentation for this purpose and include causes, mode of transmission and then preventive measures

The bacteria that causes diseases are harmful to humans and other animals and are referred to as pathogenic bacteria. The body is a home to many millions of bacteria. Some are useful while others are harmful to humans. A bacterial disease is caused by entry of bacteria into a host which can grow and flourish in the host, causing harm to the host. Bacteria cause diseases like cholera, tuberculosis (TB), typhoid fever, pneumonia, tetanus, and diphtheria, and bacterial meningitis, tooth decay in humans and anthrax in cattle.

Table 3.2: Common bacterial diseases in humans

Disease	Cause	Mode of transmission	Signs and symptoms	Prevention and control
Cholera	*Vibrion cholerae*	Through food and drinks contaminated by *Vibrio cholera*. Also water borne and spread by vectors such house flies from human faeces.	Running diarrhoea, painless, watery excretion that can quickly cause severe dehydration and death if the treatment is not immediate. Vomiting by most patients.	Washing hands after leaving the toilets and before meals, cleaning the body, purification or boiling drinking water, quick disposal of household waste, defecation in latrines, wash thoroughly all edibles when ready for eating.
Dysentery or bloody diarrhea	*Shigella* bacterium	Food and drinks contaminated by Shigella bacterium	Frequent excretion of watery stool (diarrhoea), often mixed with blood, or mucus or accompanied by abdominal cramps, pressing need of bowls, constriction of rectum, and constipation.	Cleaning the body, purification or boiling drinking water, quick disposal of household waste, defecation in latrines, wash thoroughly all edibles when ready for eating
Typhoid Fever	*Salmonella typhi*	Food and drinks contaminated by *Salmonella typhi*	Headache, insomnia, fatigue, continuous fever sometimes rising up to 40 °C, without raising the heart beats, possible nose bleed, feeling backache.	Cleaning the body, purification or boiling drinking water, quick disposal of household waste, defecation in latrines, wash thoroughly all edibles when ready for eating
Meningitis	Two main forms: bacterial meningitis, and viral meningitis.	Transmission through the air from a sick person to healthy person through coughing and sneezing.	High fever that stops a sick person from eating or drinking, strong headache, stiff neck, nausea and vomiting, and confusion.	Meningitis is potentially a deadly disease. Preventive measures are required, and the vaccine against meningitis is available, especially for the A and C.

	With the first one affecting fewer people. It is fatal and survivors have a severe disability caused by brain injury.			meningocoque type in case of epidemic.

Table 3.3: Common Bacterial Diseases in Plants

Name of disease	Cause	Signs and symptoms
Common scab	*Streptomyces scabies*	Scabs or patches on surface of potato tubers.
Bacterial blight of coffee	*Pseudomonas sp.*	Dark swellings or lesions and scotching of leaves.
Black in cotton	Bacteria	Leaves become black.
Black rot of cotton	Bacteria	Rotting of the whole plant
Blight of beans	Bacteria	Scotching of leaves

3.2.7. Economic importance of bacteria

Activity 3.2.7

When an animal dies in a forest, it decays after a certain period of time. Once a farmer grows beans in the soil with such dead animal decay, beans grow well.

1. What cause the dead animal to decay?
2. Why the beans have grown well?

Bacteria are economically important as these microorganisms are used by humans for many purposes and are harmful in causing disease and spoiling food. Bacteria are useful in many ways:

a. Biotechnology

Bacteria are used in biotechnology for example in the manufacturing industries. They are used to manufacture products such as ethanol, acetone,

organic acid, enzymes, and perfumes. In the chemical industry, bacteria are most important in the production of pharmaceuticals. For example, *Escherichia coli* is used for commercial preparation of riboflavin and vitamin K.

b. Genetic engineering

Bactria are used in genetic engineering through the manipulation of genes, also called recombinant DNA technology. In this case, bacterial cells are transformed and used in production of commercially important products for example, production of human insulin used against diabetes.

c. Decomposition of dead organisms

In addition, bacteria are important in decomposition of dead organisms and animal wastes such as feces to form organic matter. This process improves soil fertility and plays an important role in mineral recycling in an ecosystem.

d. Fibre retting

Some bacteria including *Clostridium butylicum* are used to separate fibres in a process called retting. In this process, fibres are formed to make ropes and sacks.

e. Nitrogen fixation

Some other bacteria are used to fix nitrogen in form of nitrates into the soil. For example, *Rhizobium* bacteria which lives in root nodules of leguminous plants. Such bacteria help in improvement of soil fertility especially during nitrogen cycle.

f. Digestion

Some bacteria living in the gut of ruminant animals such as cattle, horses and other herbivores secrete cellulase, an enzyme that helps in the digestion of the cellulose of plant cell walls. Cellulose is the major source of energy for these animals. Another example is *Escherichia coli* that live in the human large intestine which synthesizes vitamin B and releases it for human use.

g. Biological control

Some bacteria are used as biological agents in biological pest control such as *Bacillus thuringiensis* (also called BT) instead of pesticides. Because of their specificity, these bacteria are regarded as environmentally friendly, with little effect on humans, wildlife, pollinators, or other beneficial insects.

Application activity 3.2

1. Discuss 4 characteristics for each among the five kingdoms
2. What are the three methods that protists use to obtain food?
3. Identify three characteristics of protists
4. The following is a list of organisms belonging to various kingdoms: housefly (*Musca domestica*), maize (*Zea mays*), Frog (*Rana spp*), Bat and Eagle.
   a) Classify these organisms into their kingdoms.
   b) Name any two organisms that are not closely related and give a reason.
5. How are fungi different from members of kingdom plantae?
6. Mr. Green lives in one of the slums in a certain city. He prepares and sells chapattis on street. He is usually very clean, but one morning, he is late for work so he does not bother to wash his hands after visiting the toilet. That day he prepares 400 chapattis all of which are sold. Few hours later, his customer Sandra suffered from a disease with the following signs and symptoms: severe diarrhea, excessive loss of water leading to dehydration, and vomiting, after five days. Later, all his customers were rushed and admitted in hospital due to the same problem.
   a) Suggest the disease that Mr. Green's customers were suffering from and what caused the disease?
   b) Name three other ways this disease might be spread around city.
   c) After reading this scenario, what message do you have for people who are like Mr. Green?
   d) Suppose you were the health officer for the area in town with such a problem. What steps would you take to prevent the disease from spreading further?
   e) House flies are described as vectors. Describe, how houseflies transmit diseases tohumans.
7. Discuss 6 economic importances of bacteria.
8. Explain how bacteria are used as biological control.

3.3. Classification of viruses and their economic importance

Activity 3.3

Visit the internet and conduct a research to explain the following:

a). The classification of viruses

b). The economic importance of viruses.

3.3.1. Classification of viruses

Viruses can be classified according to:

- Type of nucleic acid molecules they have. Most animal viruses contain RNA while plant viruses contain DNA.
- Type of host cell: plant or animal viruses as they are specific to their hosts.
- Presence or absence of the envelope: Plant viruses' bacteriophage are no enveloped while animal viruses like HIV and influenza virus are enveloped.

3.3.2.Characteristics of viruses

Viruses are microorganisms whose structure is only visible with electron microscopes. A typical virus consists of DNA or RNA within a protective protein coat called capsid which provides protection. Viruses become active in metabolism only once insidethe host cell.When they infect cells, they use biochemical machinery and proteins of the host cellto copy their nucleic acids and to make proteins coats often leading to destructionof the host cells. The energy for these processes is provided by the ATP from the hostcell.

Because viruses do not consist of cells, they also lack cell membranes, cytoplasm, ribosomes, and other cell organelles. Without these structures, they are unable to make proteins or even reproduce on their own. Instead, they must depend on a host cell to synthesize their proteins and to make copies of themselves. Viruses infect and live inside the cells of living organisms. They are also regarded as parasites since they depend entirely on living cells for their survival. Although viruses are not classified as living things, they share two important traits with living things: They have genetic material, and they can evolve.

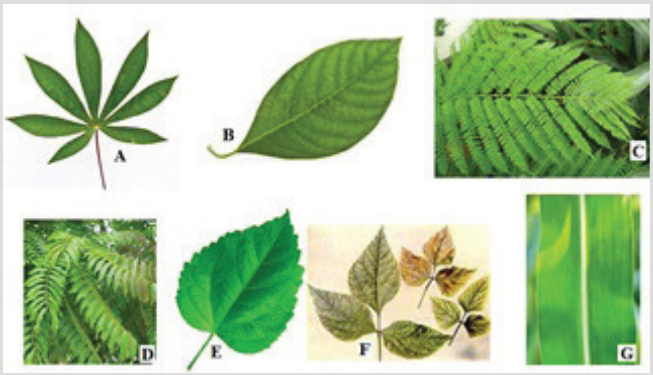
Application activity 3.3

1. What is meant by the term virus?
2. State the main components of a virus.
3. Describe the two ways how viruses cause an infection
4. Differentiate between a bacteriophage and a retrovirus?
5. Do you think viruses should be considered as a form of life? Give reasons for your answer.

3.4. Dichotomous keys for identification of organism

Activity 3.4

The figure below represents different types of plant leaves. Make a classification of these plants based on the external structure of the leaves.



The dichotomous key is also called biological identification key. It is made up of a series of contrasting statements called leads indicated by the numbers 1, 2, 3…where each lead deals with a particular observable characteristic. The characteristics used in keys should be readily observable morphological features which may be either qualitative, such as shape of abdomen, nature of legs and color, or quantitative, such as number of antennae, number of pairs of legs and length of the antennae in case of arthropods. It is essential to note that size and color are often less considered to both can be influenced by the environment, the season, the age or state of the organism at the time of identification.

3.4.1. Guidelines used in construction of a dichotomous key

The following guidelines must be considered while constructing a dichotomous key.

- Use morphological characteristics which are observable as much as possible such as leaf venation, nature of margin, apex, lamina and nature or length of the petiole (leaf stalk).
- Start with a major characteristic that divide the organism or the specimen into two large groups such as the type of a leaf.
- Select a single characteristic at a time and identify it using a number for example: simple leaf……go to 2, compound leaf……go to 5. This means that in 2 you will deal with only simple leaves and 5 only compound leaves.
- Use similar forms of words for two contrasting statements for example at 2: leaf with parallel venation ……go to G and leaf with network venation ……go to 3.
- The first statement should always be positive.
- Avoid generalizations or overlapping variations, be specific and precise to the point.

Example

- Collect leaves from the following plants: cassava, avocado, jacaranda, cassia, hibiscus bean, maize or paspalum grass,
- Label different leaves collected as, A, B, C, D, E, F and G
- Observe and familiarize with the specimens before starting the experiment to minimize errors during the identification process
- Make a table summarising the specimens and steps followed to identify each of them.
- Construct a dichotomous key basing on the observable features (characteristics) and table of steps followed.

Solution: The dichotomous key of specimens A, B, C, D, E, F and G.

1a) Simple leaves ----------go to 2

b) Compound leaves ---------- go to 5

2 a) Parallel venation ----------G

b) Network venation ---------- go to 3

3 a) Simple digitate ----------A

b) Non simple digitate ----------go to 4

4 a) Leaf with serrated margin ----------E

b) Leaf with smooth margin ----------B

5 a) Leaf with three leaflets (compound trifoliate)----------F

b) Leaves with more than three leaflets ----------go to 6

6 a) Pinnate leaf ---------- D

b) Bipinnate leaf ---------- C

3.4.2. Common features used for identification of animals

Animals are classified basing on the following features:

- Locomotory structures such as legs, wings and fins
- Antennae (presence, nature and number)
- Presence or absence of eye and eye type
- Number of body parts for example insects have three body parts
- Body segments (nature and number)
- Body surface structures such as fur, hair, feathers and scales
- Feeding structures such as mouth parts in arthropods for example in insects
- Type of skeleton present such as endoskeleton, exoskeleton and hydrostatic

3.4.3. Common features used for identification of plants

Plants can be classified basing on the following features:

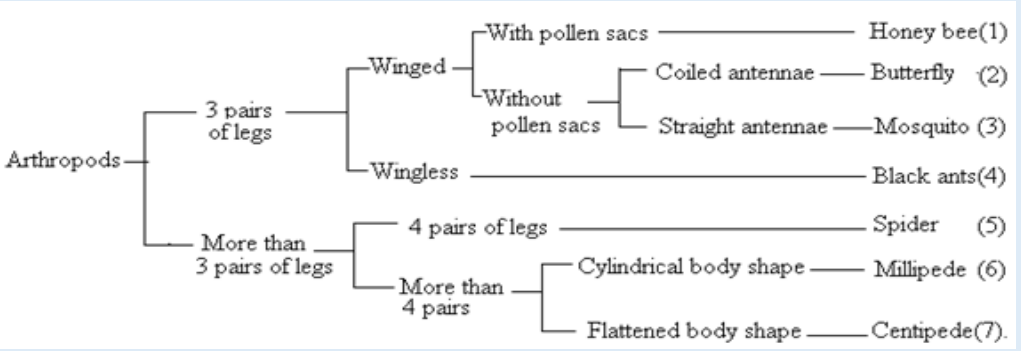
- The leaf structure such as: nature of apex, margin, venation, lamina and petiole
- The flower structure including inflorescence type, flower shape and number of floral parts
- The type of stem (woody, fleshy and herbaceous), shape (rectangular, cylindrical) and texture of the stem (smooth, spiny and thorny) …
- The type of root system, tap root, storage root, fibrous roots…

Precaution

- Care must be taken while collecting and handling some organisms because some are poisonous, have thorns and others are able to sting
- Collection of specimen should be done a day or few days before the experiment depending on nature of the experiment
- Avoid and try to minimize where possible, uprooting, cutting down or plucking and pruning of plants as this may threaten the biodiversity as well as result into environmental degradation

Application activity 3.4

Read and interpret the dichotomous tree below and use it to answer the following questions.



1. Specify the phylum of kingdom animalia represented by the above dichotomous tree? Give one observable reason to support your answer.
2. According to this dichotomous tree, which characteristic feature was used to classify different insects?
3. Which observable characteristic feature distinguishes between a spider and a mosquito?
4. How does a millipede differ from a centipede?
5. To which classes do a millipede and a centipede belong?
6. Which class of arthropods is not represented on figure 2.12?

Skills lab 3

- Move around the school compound and select a small plant which you know the scientific name and the the taxonomic hieracrchy.
- Place the plant between two sheets of newspaper and in between some heavy books. ...
- When your plant is pressed, remove it from the newsprint and carefully glue it to the 11x16 paper. ...
- Take your printer paper, cut a 3" x 4" piece of paper and glue this onto the bottom right corner of your herbarium sheet.
- Label your work (Taxonomic hierarch, Sientific name, Local name, date and your name)

End unit assessment 3

1. Which one of the following living organisms belongs to Domain Bacteria?
   a). Euglena
   b). Vibrio cholerae
   c). Paramecium
   d). Moulds
2. The group of classification where organisms resemble one another and are capable of interbreeding together to produce viable offspring is known as:
   a). Species
   b). Kingdom
   c). Genus
   d). Phylum
3. Which one of the following is not a kingdom of living organisms?
   a). Monera
   b). Animalia
   c). Annelida
   d). Protoctista
4. Which one of the following is a characteristic feature common to fish, reptiles and birds but absent in mammals?
   a). Possession of scales
   b). Has no limbs
   c). Possession of feathers
   d). Undergo internal fertilization

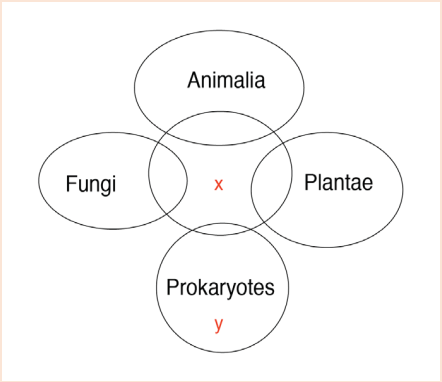
5. Which one of the following statements about fish is not correct?
   a). Fish live both in water and on land and undergo external fertilization.
   b). Most fish have bones while others are cartilaginous
   c). Most fish have streamlined body, lateral line and swim bladder.
   d). Gills are organs for gaseous exchange in fish
6. Which one of the following is not a characteristic of all insects?
   a). They have three body parts namely head, thorax and abdomen.
   b). They have three pairs of jointed legs attached on segment of the thorax.
   c). They have four pairs of jointed legs
   d). They have a pair of antennae attached on the head.
7. The following are characteristics of all mammals except;
   a). They have mammary glands to secrete milk feed their young ones.
   b). Their skin is covered with hair.
   c). Undergo internal fertilization and internal development of the embryo.
   d). They have a pair of wings made up feathers.
8. The point where the leaf joins the stem is called;
   a). Apex
   b). Margin
   c). Leaf base
   d). Lamina
   e). Length of petiole.
9. Which of the following is less considered while identifying feature to construct
   a). dichotomous key of leaves?
   b). Nature of margin
   c). Nature of apex
   d). Size and color of leaf

10. The following are characteristics of arachnids except;
    a). Four pairs of jointed legs
    b). Two body parts
    c). Three body parts
    d). Do not have wings

11. Match the structures with the organisms which possess them.

Structures	Organisms
Antennae	Fungus
Flagella	Snail
Spores	Housefly
Coiled shell	Euglena
Pseudopodia	Amoeba
Cilia	Paramecium

12. A group of Y1 students drew a Venn diagram below to summarize the five kingdoms into which organisms are classified. Study the diagram and answer the questions that follow:



    a). Which kingdoms are represented by the letters x and y?
    b). State one characteristic that organisms of x may share with:
        i. Prokaryotes
        ii. Fungi
        iii. Plantae

13. What is the significance of classification of living organisms?

UNIT 4:

SOLUTIONS AND TITRATION

Key Unit competence: To prepare standard solutions and use them to determine concentration of other solutions by method of titration.

Introductory Activity 4

Observe the photo below and attempt the questions that follow:

1. For the bottle of beer 1, and that of Beer 2, we find on the bottle of its labels 5% alcohol and 6% alcohol respectively.
   a) Explain the meaning of 5% and 6% alcohol.
   b) 5% alcohol corresponds to which volume of alcohol of the total volume of 72cl of the bottle of beer 1.
   c) Calculate the volume of alcohol corresponding to 6% alcohol of the bottle of Beer 2 of the total volume of 65cl.
2. On the label of the bottle of fruit juice we find that dilution is 1:5. Explain this ratio and explain also the purpose of diluting substances.

4.1. Definition of standard solution and primary standard solution

Activity 4.1

1. Explain the concepts of:
   a) Solution
   b) concentration
2. You are given a basic solution, NaOH (aq), and you are requested to determine its concentration. How could you proceed to know its concentration?

A homogeneous mixture is known as a **solution**.

An **aqueous solution** is a solution in which the solvent is water. Its concentration may be estimated in many ways.

In analytical chemistry, a **standard solution** is a solution containing a precisely known concentration of an element or a substance and used to determine the unknown concentration of other solutions. A known weight of solute is dissolved to make a specific volume. It is prepared using a standard substance, such as a primary standard.

Standardization is the process by which the concentration of a solution is determined by measuring accurately the volume of the solution required to react with an exactly known amount of a **primary standard**.

A primary standard is defined as a substance or compound used to prepare standard solutions by actually weighing a known mass, dissolving it, and diluting to a definite volume.

Or a substance, which is chemically stable in aqueous solution and its concentration remains constant with change in time such that it can be used to standardize other solutions.

Some important examples of primary standard are;

- Sodium Carbonate, Na_2CO_3
- Potassium dichromate, $K_2Cr_2O_7$
- Benzoic acid, C_6H_5COOH
- Oxalic acid, $H_2C_2O_4$
- Iodine, I_2

- Sodium oxalate, $Na_2C_2O_4$
- Butanoic acid, C_3H_7COOH
- Sodium tetraborate, Borax, $Na_2B_4O_7$
- Potassium chloride, KCl
- Arsenic(III)oxide, As_2O_3
- Silver nitrate, $AgNO_3$

Standard solutions are normally used in titrations to determine unknown concentration of another substance.

Application activity 4.1

Differentiate between standard solution and primary standard solution

4.2. Properties of a primary standard solution

Activity 4.2

1. Recall the concept of "primary standard solution"
2. Do research and find out the role and characteristics of a good primary standard solution.

A good primary standard solution meets the following criteria:

- High level of purity
- High stability
- Be readily soluble in water
- High equivalent weight (to reduce error from mass measurements)
- Not hygroscopic (to reduce changes in mass in humid versus dry environments)
- Non-toxic
- Inexpensive and readily available
- React instantaneously, stoichiometrically and irreversibly with other substances i.e. should not have interfering products during titration.
- It should not get affected by carbon dioxide in air

Molar concentrations are the most useful in chemical reaction calculations because they directly relate the moles of solute to the volume of solution.

The formula for molarity is:

$$\text{Molarity} = \frac{\text{moles of solute}}{\text{volume of solution}}$$

Application activity 4.2

Discuss the properties of a good primary standard solution.

4.3. Preparation of standard solutions

Many of the reagents used in science are in the form of solutions which need to be purchased or prepared. For many purposes, the exact value of concentration is not critical; in other cases, the concentration of the solution and its method of preparation must be as accurate as possible.

Thus, the preparation of solutions is one of the most fundamental tasks performed by a chemist.

A solution may be prepared by two methods: **dissolution method and dilution method**.

4.3.1. Preparation of standard solution by dissolution of solids

Activity 4.3.1

Some chemicals of high interest used in scientific laboratories, hospitals, pharmacies, are found in the form of solutions with precise concentration.

1. What is the most common unit of the solution concentration?
2. Using your knowledge in chemistry acquired so far and available resources identify the requirements and the procedure for the preparation of such solutions starting with solids.

The most common unit of solution concentration is molarity (M).

The molarity or molar concentration of a solution is defined as the number of moles of solute per one litre (mol/L) of solution.

Molarity, therefore, is a ratio between moles of solute and litres of solution.

To prepare laboratory solutions, usually a given volume and molarity are required. To determine molarity, the formula weight or molar mass of the solute is needed.

When preparing a solution starting with a solid, the following steps may be followed:

- Determine the mass in grams of one mole of solute, the molar mass, Mm.
- Decide volume of solution required, in litres, V.
- Decide molarity of solution required, M.
- Calculate grams of solute required.

In the preparation of solution, glasses, volumetric flask, pipette, droppers, glass rod, measuring cylinder, analytical balance, spatula, beakers, magnetic stirrer and other laboratory devices are used.

The preparation of a solution by dissolution follows the steps illustrated by figure 4.1

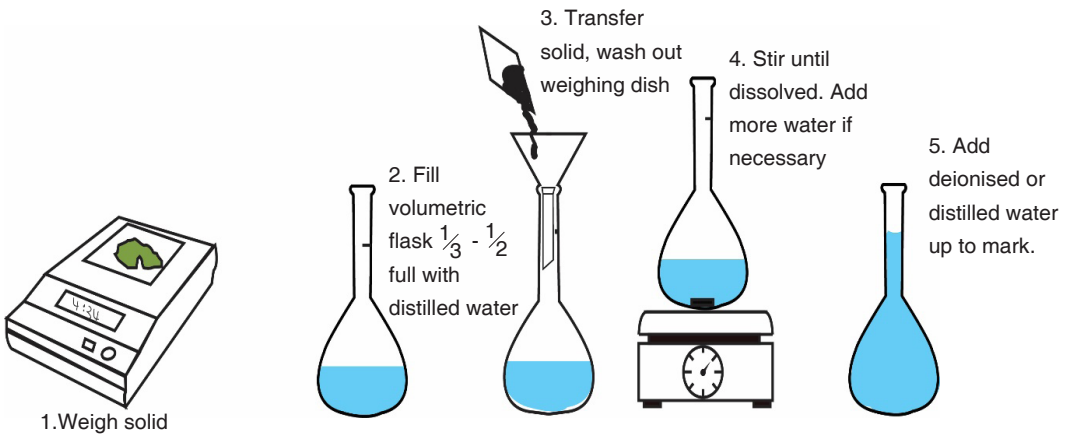


Figure 4.1: Preparation of solution by dissolution of solids

The following examples illustrate the calculations for preparing solutions.

Examples:

1. Describe in details how you can prepare the following solution: 50 mL of NaOH, 10%.

Answer:

Preliminary calculations

10% means that in 100 mL of solution only 10 g are pure NaOH. So 50 mL of NaOH will be prepared by taking the mass of NaOH, dissolving it in water and making up to 50 ml.

$$\text{Mass of NaOH} = \frac{10\text{kg} \times 50\text{ml}}{100\text{ml}} = 5\text{g of NaOH}$$

Procedure

- Weigh 5g of NaOH accurately using an analytical balance.
- Carefully dissolve it in a beaker containing about 30mL of distilled water and stir using a glass rod or a magnetic stirrer till you get homogeneous mixture.
- Transfer the solution and the washings in a 50 mL volumetric flask and mix.
- Add more distilled water until the level is about 2 cm below the graduation mark on the volumetric flask.
- Top up using distilled water from a dropper until the bottom of the meniscus is at the level of mark when viewed at eye level.
- Stopper the flask and shake again to homogenise.
- Label your solution: 10% NaOH; 50 mL
- The solution prepared is 10% NaOH

Usually, the minimum label requirements are: (1) Identity of contents (2). Concentration (3) Your name and (4) date of preparation.

2. Describe in details the preparation of 250 cm^3 of a 0.1M Na_2CO_3 solution.

Answer:

Step 1: Preliminary calculations

Calculate the amount of anhydrous sodium carbonate required to be dissolved in 250cm^3 of solution. i.e:

Molar mass of Na_2CO_3= (23 x 2) +12+ (16 x 3) = 106 g/mol

Thus, 1 mole of Na_2CO_3 has a mass of 106 g.

0.1 mole of Na_2CO_3 solution will have a mass = 106g/mol x 0.1 mol = 10.6 g.

1000 cm^3 of 0.1M Na_2CO_3 solution contains 10.6 g.

250 cm^3 of 0.1M Na_2CO_3 = $\frac{10.6g \times 250}{1000}$ = 2.65g

Step 2: Procedure

- Weigh 2.65g of sodium carbonate and carefully transfer in a beaker containing 150mL of distilled water.
- Stir using a glass rod until complete dissolution.
- Using a filter funnel, transfer the solution and the washings, in a 250 mL volumetric flask.

- Using a wash bottle, add more distilled water up to about 1cm of the graduation mark.
- Using a dropper pipette, complete the volume the graduation mark.
- Stopper the flask and shake to homogenise.
- Label the solution.

Application activity 4.3.1

Give a detailed description, including calculations involved, for the preparation of the following solutions:

a). 1L of a 0.01M $KMnO_4$ solution

b). 250 mL oxalic acid 10g/L.

4.3.2. Preparation of standard solution by dilution

Activity 4.3.2

Some chemicals of high interest used in scientific laboratories, hospitals, pharmacies, are found in the form of solutions with precise concentration.

Some are obtained by dissolution of solids and others may be obtained by dilution of stock solution.

1. What is meant by:

   a) stock solution? b) dilution?

2. Do research to identify the requirements and the procedure for the preparation of such solutions starting with solutions or liquids.

Apart from, dissolution method, we can prepare a solution by dilution from the stock solution. This consists of reducing the concentration of a concentrated stock solution to less concentrated

A simple dilution is one in which a unit volume of material of interest (solute) is combined with an appropriate volume of a solvent to achieve the desired concentration.

The dilution factor is the total number of unit volumes in which the solute is dissolved. For example a 1:5 dilution involve combining 1 unit of the solute and 4 unit volumes of the solvent.

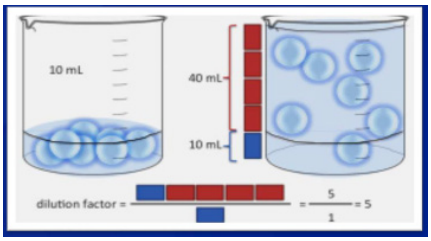


Figure 4.2: Demonstration if dilution factor

Thus, when preparing a solution starting with a solution or liquid reagent:

When diluting more concentrated solutions,

- Decide what volume (V_f) and molarity (M_f) the final solution should be.Volume can be expressed in liters or milliliters.
- Determine molarity (M_i) of starting, more concentrated solution.
- Calculate volume of starting solution (V_i) required using equation:

  $\mathbf{M_iV_i} = M_fV_f$

  Where M_i and M_f are the initial and the final concentrations (or initial and final molar concentrations or initial and final molarities);

  V_i and V_f are the initial and the final volumes.

  Note: $V_i$must be in the same units as V_f.

The following are the requirements: Graduated pipettes and /or micropipettes, volumetric flasks, Stoppers, Dropping pipettes, Stock solutions, Droppers, Distilled water.

The preparation of a solution by diluting stock solutions follows the steps illustrated by figure 4.2

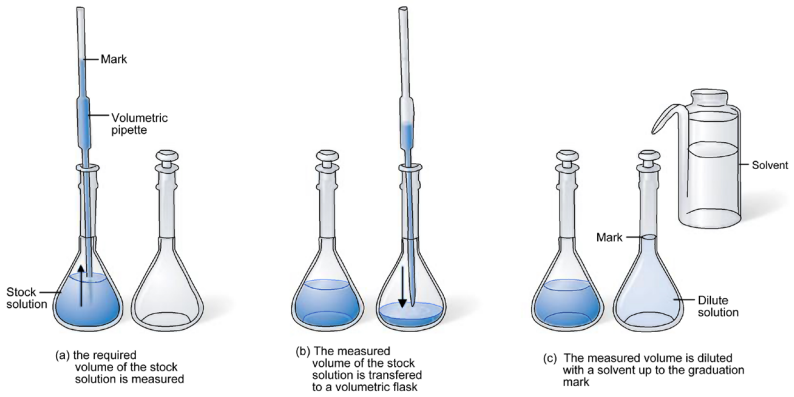


Figure 4.3: Preparation of solution by diluting a stock solution

Examples:

1. Describe how you can prepare 1L of a 1M H_2SO_4 solution from the stock solution with the following specification: density (d) =1.84; percentage (P) by mass: 98%; Mw = 98 g/mol.

Answer:

Preliminary calculations

The molarity of the stock solution may be calculated using the following relation.

$$M = \frac{10 \times d \times p}{Mw}$$

Therefore, the molarity of the concentrated sulphuric acid is equal to

$$\frac{10x1.84x98}{98} = 18.4M$$

From dilution law, we have $M_i \times V_i = M_f \times V_f$

$$V_i = \frac{Mf \times Vf}{Mi} = \frac{1Mx1L}{18.4M} = 0.05435L = 54.35mL$$

Hence the required volume of concentrated sulphuric acid to be pipetted is 54.35mL.

Procedure:

1) Using a graduated pipette or a burette, measure 54.35mL of concentrated H_2SO_4;
2) Using a funnel, carefully transfer the concentrated acid into a 1L volumetric flask containing about 700mL of distilled water;
3) Put a stopper and gently shake the mixture;
4) Using a wash bottle, add more distilled water until the level is 1cm below the mark;
5) Using a dropper pipette, make up to 1L dropwise;
6) Stopper the volumetric flask and gently shake.
7) Label the solution and specify the preparation date.

The solution prepared is 1M H_2SO_4.

2. Calculate the volume of 15M H_2SO_4 that would be required to prepare 150cm^3 of 2MH_2SO_4.

Answer:

$$M_i \times V_i = M_f \times V_f$$

$$V_i = \frac{M_f x V_f}{M_i} = \frac{2Mx150cm^3}{15M} = \mathbf{20cm^3}$$

The volume of 15M H_2SO_4 required is 20 cm^3 **or**

1000cm^3 of diluted acid contain 2mol

150cm^3 of diluted acid will contain = $\frac{2x150mol}{1000}$ = 0.3 mol

But, 15 moles of concentrated H_2SO_4 are contained in 1000 cm^3,

Then, 0.3moles of concentrated H_2SO_4 will be contained in $\frac{1000x0.3}{15}$ = 20 cm^3

Volume of 15M H_2SO_4 required 20 cm^3.

Application activity 4.3.2

Describe the preparation of 500mL 0.5M HCl solution from a stock solution of HCl with the following specifications.

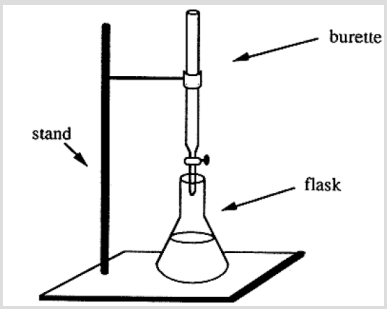
- Molar weight: 36.5g/mol
- Percentage of HCl: 36% by mass
- Density: 1.18g/mL

4.4. Titration process focussing on precise measurements

4.4.1. Simple acid-base titrations

Activity 4.4.1

1. Analyse the diagram below and answer the questions that follow.



a) What technique is illustrated by the diagram above?

b) Is it important in real life to perform operations illustrated by the above set up? Explain using at least two examples.

c) What are the necessary requirements for such operations?

2. You are provided with:

S_1: solution of HCl (aq) 0.1 M

S_2: solution of NaOH (aq)

Materials: Burettes - Indicator (phenolphthalein) - pipettes - Washing bottles - Conical flasks - Beakers - Retort stands - Funnels

Procedure

a). Using a pipette, transfer 10 mL of S_2 into a conical flask.

b). Add three drops of phenolphthalein indicator and titrate it with S1 from the burette.

c). Repeat the titration until you obtain consistent values.

d). Record your readings in the table below:

Experiment number	1	2	3
Final burette reading			
Initial burette reading			
Volume of S_1 used			

e). Calculate the average volume of S_1 used.

f). Calculate the number of moles of HCl that react with S_2.

g). Calculate the molarity of S_2 in 10 mL.

Titration is a process used in volumetric analysis, in which a solution of one reactant of known concentration, of unknown concentration, the **titrant**, is added to a known volume of another reactant of unknown concentration, the titrand, until the reaction between the two is complete.

The analyte (titrand) is the specie of interest in a titration i.e, the specie that is subjected to analysis.

If a concentration and volume of the titrant is reacted with the analyte, it is possible to calculate the analyte's concentration.

The completion of the reaction is signalled by an **indicator**, i.e a substance that has distinctly different colours in acidic and basic media. E.g phenolphthalein.

The point at which the indicator changes the colour and the titration is stopped is called **end point**.

The **equivalence point** is the point, in a titration, at which the mole ratio of reactants is exactly equal to the mole ratio in the balanced chemical equation. The difference between the end point and equivalent point is called **titration error.**

Because volume measurements play a key role in titration, it is also known as ***volumetric analysis.*** Usually it is the volume of the titrant required to react with a given quantity of an analyte that is precisely determined during a titration.

Titrations can be classified by the type of reaction. Different types of titration reaction include acid-base titrations, redox titrations, complexometric titrations, etc. In our context we will deal with acid-base titrations.

Acid-base titrations

In an acid-base titration, a standard solution of a base is added to a known volume of an acid with unknown concentration or vice versa, until the end point is reached.

At the **equivalence point**, the acid and the base that have reacted are in equivalent amounts.

Therefore, $n_a M_b V_b = n_b M_a V_a$,

Where M_a=Molarity of acid; M_b =Molarity of the base

V_b =Volume of base; n_b = stoechiometric number of base

V_a =Volume of acid; n_a = stoechiometric number of acid

Indicators and pH-meters can be used to determine the equivalence point.

An acid-base titration needs an indicator which will change colour with the concentrations of hydrogen ions (H^+).

Choice of indicators in acid-base titrations

When the technique of acid-base titration is extended to a wide variety of acidic and alkaline solution, care needs to be taken about the choice of indicator for any given reaction.

The choice of an inappropriate indicator would lead to incorrect results, and it is therefore extremely important that the indicator is chosen carefully.

The principle on which a choice of indicator is made concerns the strength of the acid or base involved in the reaction. Note that the strength of an acid or base is not to be confused with the concentration of its solution. Example of strong and weak acids and bases and choice of indicator are given in the Table below.

Table 4.1. Examples of strong/weak acids and bases

Acids		Bases	
Strong	**Weak**	**Strong**	**Weak**
Hydrochloric acid, HCl	Ethanoic acid, CH3COOH	Sodium hydroxide, NaOH	Ammonia, NH_3
Nitric acid, HNO_3	Methanoic acid, HCOOH	Potassium hydroxide, KOH	
Sulphuric acid, H_2SO_4	Carbonic acid, H_2CO_3	Sodium amide, $NaNH_2$	

Indicators which are suitable for particular types of acid-base titrations are given in the Table 4.2.

Examples of indicators

Table 4.2. Examples of acid-base indicators

Acid-base titration	Example	Choice of indicator
Strong acid/ strong base	H_2SO_4 and NaOH	Any indicator
Weak acid/ strong base	CH_3COOH and KOH	Phenolphtalein
Strong acid/ weak base	HCl and NH_3	Methyl orange
Weak acid/ weak base	CH_3COOH and NH_3	No satisfactory indicator

The common equipments used in titration are:

Burette
Pipette
Retort stand and clamp
Dropper pipette
pH-indicator/acid-base indicator
White tile: used to see a colour change in the solution(a white paper can also be used)
Conical flask (Erlenmeyer flask) or a beaker

How to perform titrations?

It is essential to be familiar with the use of pipette and burettes and how to handle them. The following points are useful in order to correctly perform a titration.

1. The apparatus should be arranged as shown in the activity 4.4.1
2. The burette tap is opened with the left hand and the right hand is used to shake the conical flask.
3. The equivalence-point is reached when the indicator just changes permanently the colour.
4. At the end-point, the level of the titrant is read on the burette.
5. The titration is now repeated, three more times are recommended. Towards the end-point, the titrant is added dropwise to avoid overshooting.

Notice: Before titration, check if the tip of the burette is filled with the titrant, and doesn't contain bulb of air. If there is a bulb of air, a quick opening and closing of the tap will expel the air out of the burette.

The results are summarised in a table as shown below:

Burette readings should be written to two decimal places (for burette having precision up to hundredth)

Table 4.3. Sample results

Experiment number	1	2	3
Final burette reading / cm^3	25.50	34.00	29.10
Initial burette reading / cm^3	0.00	10.00	5.00
Volume of solution used / cm^3	25.50	24.00	24.10

Average titre should be obtained using values which differ not by more ± 0.10cm^3.

Such values are known as **consistent values**. In the example above, consistent values are 24.00 cm^3 and 24.10 cm^3

$$\text{Average volume} = \frac{24.00+24.10}{2} = 24.05 \text{ cm}^3$$

Examples:

1. A volume of 20.00 mL of 5.0 x 10^{-3} M NaOH is required to reach the end-point in the titration of 10.0 mL of HCl of unknown concentration. Use these titration data and determine the molarity of the acidic solution.

Answer

Begin with the balanced neutralization reaction equation. From the equation, determine the chemically equivalent amount of HCl and NaOH.

$HCl\ (aq) + NaOH\ (aq) \rightarrow NaCl\ (aq) + H_2O\ (l)$

1 mole 1 mole 1 mole 1 mole

Calculating the number of moles of NaOH used in the titration;

Molarity = n/v; n = Molarity x Volume

$n = 5.0 \times 10^{-3}\ M \times 20.0 \times 10^{-3}\ L = 10^{-4}$ moles of NaOH

Because one mole of NaOH is needed to neutralize one mole of HCl, the amount of HCl in the titration must be 10^{-4} moles of HCl.

The molarity of the HCl can now be calculated.

At the end-point: MaVa = MbVb because na = nb

Ma = Molarity of acid HCl Mb = Molarity of base NaOH

Va = Volume of acid HCl Vb = Volume of base NaOH.

$$Ma = \frac{MbxVb}{Va} = \frac{0.005\frac{mol}{1}x0.02l}{0.010l}$$ or Ma = 0.01M of HCl

$Ma = 10^{-2}$ mol/L or $Ma = 10^{-2}$ M or Ma = 0.01 M of HCl

2. 20.0 cm^3 of 0.10 M sodium carbonate completely reacted with 25.0 cm^3 of hydrochloric acid using methyl orange indicator.

   a). Write the equation of reaction taking place.

   b). Calculate:

   i. The number of moles of sodium carbonate that reacted.

   ii. The number of moles of hydrochloric acid that reacted.

   iii. The concentration of hydrochloric acid in $mole.dm^{-3}$.

Answer:

a). $Na_2CO_3(aq) + 2HCl(aq) \rightarrow 2NaCl(aq) + H_2O(l) + CO_2(g)$,

b). (i) Moles of sodium carbonate used = $\frac{20.0x0.10}{1000} = 0.002mol$

(ii) Reaction ratio Na_2CO_3: 2HCl = 1:2

Moles of HCl = 0.002mol x 2 = 0.004mol

(iii) 25.0cm^3 of solution contain 0.004mol

1000 cm^3 of contain $\frac{0.004x1000}{25.000}$ = 0.16mol

The hydrochloric acid is 0.16M.

3. Standardization of sodium hydroxide using 0.1M Hydrochloric acid.

Reagents provided:

$FA_1$is 0.1M hydrochloric acid solution

NaOH pellets

Phenolphtalein indicator

Distilled water.

Procedure

- Weigh accurately 1.0 g of sodium hydroxide pellets in a weighing boat and transfer into a 250cm^3volumetric flask.
- Add a little distilled water and shake to dissolve.
- Make up the solution to the mark with distilled water dropwise.
- Label the solution FA_2.
- Pipette 20cm^3(or 25 cm^3) of FA_2into a conical flask.
- Add 2-3 drops of phenolphthalein indicator.
- Titrate with $FA_1$from the burette.
- Record your results in the spaces provided below.

Specimen results

Mass of weighing container + NaOH pellets = 11.30 g

Mass of weighing container alone = 10.30 g

Mass of NaOH pellets = 11.30 g – 10.30 g = 1.00 g

Masses should be recorded to at least two decimal places (Table 11.4)

Volume of pipette used = ……..

Table 4.4. Burette readings

Experiment number	1	2	3
Final burette reading / cm^3	36.20	25.10	30.80
Initial burette reading / cm^3	10.80	0.00	5.60
Volume of FA_1 used / cm^3	25.40	25.10	25.20

Each value or entry in the table must be recorded or written to two decimal places.

Different initial readings should be used. Initial reading in each experiment should be correctly subtracted from the final reading.

Questions:

a). Determine the values used to calculate average volume, and calculate the average volume of FA_1 used.

b). Write the equation of the reaction between NaOH and HCl

c). Calculate the number of moles of FA_1 used.

d). Calculate the concentration of FA_2 in $mol.dm^{-3}$.

e). Calculate the concentration of FA_2 in g/L.

Answer:

Volume of pipette used = 25.0 cm^3

a). The values used to calculate average volume are 25.10 cm^3 and 25.20 cm^3

Average volume = $\frac{25.10+25.20}{2}$ = 25.15

b). Equation NaOH (aq) + HCl(aq) → NaCl(aq) + H_2O(l)

c). Number of moles of FA_1 used:

1000cm^3 of FA_1 solution contain 0.1mol of HCl

25.15cm^3 of FA_1 solution used will contain $\frac{0.1x25.15mol}{1000}$ = 0.002515mol

Number of moles of FA_1 = 0.002515 mol

d). Concentration of FA_2 in mol. dm^{-3}

From the equation of the reaction in (a);

1mole of HCl neutralizes 1mole of NaOH

Number of moles of FA_2 used (reacted) = 0.002515 mol of NaOH.

25.0cm^3 of FA_2 used contain 0.002515 mol

1000cm^3 of FA_2 will contain $\frac{0.1x25.15mol}{1000}$ = 0.1006 mol

Concentration of FA_2 (NaOH) = 0.1006 mol dm^{-3}

e). Concentration of FA_2 in g/L (Na = 23, O =16, H=1)

Molecular mass of NaOH = 23+16+1= 40g/mol

Thus 1 mole of NaOH weighs 40 g

0.1006 mole of NaOH will weigh = 40 x 0.1006g = 4.024

Concentration of FA_2 in g/l = 4.024g/L

Application activity 4.4.1

1. 20.00cm^3 of 0.4M HCl was used to neutralize 40cm^3 of $Ba(OH)_2$. Determine the molar concentration of $Ba(OH)_2$.
2. 25.0cm^3 of sodium hydroxide solution is neutralized by 15.0cm^3 of a solution of hydrochloric acid of concentration 0.25M. Find the concentration of the sodium hydroxide solution.
3. Practical: Acid-base titration

   You are provided with the following:

   - BA_1, which is a 0.1 M H_2SO_4 solution
   - BA_2, which is a solution of NaOH with unknown concentration

Procedure

a). Pipette 20.0 cm^3 of BA_2 into a conical flask. Add three drops of methyl orange indicator.

b). (Titrate with BA_1 from burette

c). Repeat the titration until you obtain consistent values.

d). Record your results in the table below:

Experiment number	1	2	3
Final burette reading / cm^3			
Initial burette reading / cm^3			
Volume of BA1 used / cm^3			

Titre values used for calculating average volume of BA_1 are: ……

Average volume of BA_1 used: …….

Questions

1. The equation of the reaction is: …….
2. Calculate the number of moles of H_2SO_4 reacted with NaOH.
3. Calculate the number of moles of NaOH reacted with H_2SO_4.
4. Calculate the molarity of BA_2.
5. Calculate the concentration of BA_2 in g/L.

4.4.2. Determination of atomic masses by titration

Activity 4.4.2

The titration process has many applications such as the determination of an unknown concentration, acidity of a wine, mass atomic, etc...

Use textbooks and/or internet to find how atomic masses may be calculated using titration data. .

It is possible to find the atomic mass of the metal atom in titrating a basic compound which reacts with acidic solution.

Example:

1.0 g of metal hydroxide $X(OH)_2$ was dissolved and made up to 250 cm^3 of solution. A 25.0 cm^3 of the solution required 27.0 cm^3 of 0.1 mol dm^{-3} hydrochloric acid for complete neutralization using methyl orange indicator. Calculate the value of m_x, the relative atomic mass of X and suggest the formula of that metal hydroxide.

Answer:

Number of moles of HCl that reacted = $\frac{27x0.1}{1000}$ = 0.0027mol

Reaction ratio is $X(OH)_2$: HCl = 1 : 2

Number of moles of hydroxide in $1000cm^3$ = $\frac{0.0027mol}{2}$ = 0.00135mol

Number of moles of hydroxide in $25cm^3$ = $\frac{0.0035molx1000}{25x10^{-3}}$ = 0.054M

Mass of the metal hydroxide in $1000cm^3 = \frac{1.0x1000}{250} = 4.0mol$

Molar mass of metal of the metal hydroxide , $X(OH)_2 = \frac{4.0}{0.054} = 74.1g/mol$

$m_x + 34 = 74.1$ $m_x = 40.1$

The relative molecular mass value of X is 40.1, so X is Ca and the formula of the hydroxide is $Ca(OH)_2$.

Application activity 4.4.2

What is the molecular weight of an unknown monoprotic acid 0.4955g of the acid are neutralised by 37.00 of a 0.100M sodium hydroxide solution?

4.4.3. Determination of number of moles of water of crystallisation

Activity 4.4.3

Observe the picture below and answer the questions that follow.

What compound is represented by the picture above?

What observable changes could be recorded if the substance was heated and how do you interpret those observations?

When blue crystals of copper(II) is heated it becomes a white powder. This is due to the loss of water molecules attached to it. This water is known as water of crystallisation.

Water of crystallisation is water present in definite proportions in some crystalline compounds.

Many crystalline salts form hydrates containing 1, 2, 3, or more moles of water per mole of the compound and the water may be held in the crystal

in various ways. The water molecules may occupy lattice positions in the crystal or may form bonds with cations or anions present.

Examples of some hydrated substances

$BaCl_2.2H_2O$: Barium chloride

$CuSO_4.5H_2O$: Copper (II) sulphate

$Na_2CO_3.10H_2O$: Sodium carbonate

$Na_2S_2O_3.5H_2O$: Sodium thiosulphate

The amount of water of crystallisation can be determined by titration or by weighing a solid residue obtained when the water is driven off by heating.

1. Weighing method

This method is preferably used when the crystals are neutral, and they cannot react with acid or basic solution.

Procedure:

(For example, $CuSO_4.\ xH_2O$)

Weigh the mass of a clean dry crucible with lid.

Two or three grams of hydrated **$CuSO_4.XH_2O$** are added and the whole set up is weighed.

The crucible, lid and the content are heated on a pipe-clay triangle on a tripod stand, gently at first and later strongly, to drive off water of crystallization.

The set up is allowed to cool in a dessicator (to exclude moisture) and reweighed.

Heat again the crucible, lid and content and then cool as before and weigh again.

Repeat the process of heating, cooling and reweighing until a constant mass is reached which shows that all water has been expelled.

Using calculations, it is possible to get x moles of water of crystallization found in one mole of substance. See the following example:

Example:

4.99g of hydrated copper (II) sulphate, **$CuSO_4.x\ H_2O$** was strongly heated until all water of crystallization was eliminated. The mass of anhydrous copper sulphate was found to be 3.19g. Calculate the number of moles of

water of crystallization contained in one mole of **$CuSO_4.xH_2O$** (Cu = 63.5, S = 32, O =16, H = 1)

Answer:

Mass of water = 4.99g – 3.19g = 1.8g

Mr for $CuSO_4$= 63.5 +32 +64 = 159.5

Number of moles of anhydrous copper sulphate, $CuSO_4 = \frac{3.19}{159.5} = 0.02mol$

Number of moles of water = $\frac{1.8}{18}$ = 0.1 mol

Number of moles of water of crystallisation = $\frac{0.1}{0.02} = 5$

So, we may write **$CuSO_4.5H_2O$**

2. Using titration

Titration is used for crystals substances which are either basic or acidic such as $H_2C_2O_4.x\,H_2O$, $Na_2CO_3.x\,H_2O$. This method obeys all the titration principles used in acid-base titration.

Example

Hydrated sodium carbonate ($Na_2CO_3.xH_2O$) is sold commercially as "washing soda".

In an experiment to determine the value of x, 3.425g of $Na_2CO_3.nH_2O$ were dissolved in water and the solution made up to 250cm^3. 25cm^3 of this solution were titrated with a 0.113M HCl solution. The titration required 21.20cm^3.

Given that the equation of the reaction is:

$Na_2CO_3(aq) + 2HCl(aq) \rightarrow 2NaCl(aq) + H_2O(l) + CO_2(g)$,

a). Calculate the mass of sodium carbonate in the sample;

b). Calculate the value of n.

Answer:

Number of moles of HCl used = 0.113mol x $\frac{21.20cm^3}{1000cm^3}$ = 2.3956×10^{-3}mol

From the equation of the reaction, the number of moles of $Na_2CO_3.nH_2O$ is:

$n(Na_2CO_3.xH_2O) = 2.3956 \times 10^{-3}mol/2 = 1.1978 \times 10^{-3}mol$

Molarity of $Na_2CO_3.xH_2O = 1.1978 \times 10^{-3}mol \times \frac{1000cm^3}{25cm^3} = 4.7923 \times 10^{-2}M$

Mass concentration of $Na_2CO_3.xH_2O$ = 3.425g x $\frac{1dm^3}{0.250dm^3}$ = $13.7gxdm^{-3}$

Molar mass of $Na_2CO_3.xH_2O$ = $\frac{13.7g/dm^3}{4.7912x10^{-2}mol/dm^3}$ = 286g/mol

That is (2 x 23) + 12 +(3 x 16) +(n x 18) = 286

106 + 18 x = 286

18x = 180 n = 10

Formula of the washing soda = $Na_2CO_3.10H_2O$

Application activity 4.4.3

Calculate, x, the number of water of crystallisation in oxalic acid hydrate, $H_2C_2O_4.xH_2O$ from the following data.

5.00g of the compound is made up to exactly 250mL and 25mL of this solution requires 15.9mL of 0.50M NaOH solution for neutralization.

Skills lab 4

You are asked to investigate the acidity content of aspirin tablets.

Write a procedure including materials, set up and procedure you will use for that purpose.

Your project will be performed either during evenings weekends. You will request from the school authorities to avail some materials for you.

End unit assessment 4

1. Explain the following terms: a. Standard solution b. Primary standard.
2. Describe the characteristics of a primary standard.
3. A solution is made by dissolving 5.00g of impure sodium hydroxide in water and making it up to $1dm^3$ of solution.$25cm^3$ of this solution are neutralized by $30.0cm^3$ of hydrochloric acid of concentration 0.102M. Calculate the percentage purity of the sodium hydroxide.
4. You are provided with a white powder containing a mixture of sodium carbonate and magnesium nitrate and 1M nitric acid. Make a solution using 10g of the powder and $47.3cm^3$ of water. You need $0.05dm^3$ of the acid to fully react with the solution. What is the percentage composition of magnesium nitrate in the powder?

5. Calculate, x, the number of water of crystallisation in oxalic acid hydrate, $H_2C_2O_4.xH_2O$ from the following data.

   5.00g of the compound is made up to exactly 250mL and 25mL of this solution requires 15.9mL of 0.50M NaOH solution for neutralization.

6. Ethanedioic acid is a solid at room temperature. It is soluble in cold water and is toxic.

   In an exercise to determine the number of moles of water of crystallisation, x, in one mole of ethandioic acid, $(COOH)_2.xH_2O$, 1.48 g of ethanedioic acid was dissolved in water and made up to exactly 250 cm^3 of solution.25.0 cm^3 of the ethanedioic acid solution was then titrated against sodium hydroxide of concentration 0.100 mol dm^{-3}. A volume of 23.5 cm^3 of sodium hydroxide was required.

   The equation of the reaction that takes place is: $(COOH)_2 + 2NaOH \rightarrow (COONa)_2 + 2H_2O$

   a). Suggest a suitable indicator for the titration and give its colour change at the end-point of the titration.

   b). Describe how the ethanedioic acid would be dissolved and made up to exactly 250 cm^3 of solution. In your answer include the names of each piece of apparatus required and how each would be prepared for the procedure. Include a brief description of each step in the process.

   c). Calculate the number of moles of sodium hydroxide used in the titration.

   d). (i) Calculate the number of moles of ethanedioic acid in 25.0 cm^3 of solution.

   (ii) Calculate the number of moles of ethanedioic acid in 250 cm^3 of solution.

   e). (i) From your answer to (d)(ii) and the mass of ethanedioic acid used calculate the relative molecular mass of ethanedioic acid.

   (ii) Calculate a value for x.

   f). Suggest one appropriate safety precaution that should be taken in this exercise, based on the fact that ethanedioic acid is toxic.

7. Practical activity: Determination of the number of moles of water of crystallization by titration.

Procedure:

1. Dissolve 6.3 g $H_2C_2O_4.xH_2O$ (oxalic acid or ethanedioic acid) to make 1.0 litter of solution and prepare 0.1 M of solution of sodium hydroxide.
2. i). Pipette 25 cm^3 of $H_2C_2O_4.xH_2O$ into a conical flask and add two drops of phenolphthalein indicator.

   ii). Titrate with 0.1 M NaOH from burette.

   iii). Record your results in the table below:

Experiment number	1	2	3
Final burette readings in cm^3			
Initial burette readings in cm^3			
Volume of 0.1M NaOH used			

3. Calculations

a). Average volume of NaOH used = …….

b). To determine the number of moles of water of crystallization in $H_2C0_4.xH_2O$, write the chemical equation of reaction taking place.

c). Calculate the number of moles of NaOH that reacted

d). Mole ratio of $H_2C_2O_4$: NaOH = …….. : ……..

e). Calculate the number of moles of $H_2C_2O_4$ that reacted

f). Molarity of $H_2C_2O_4$ = ……..

g). Relative molecular mass of $H_2C_2O_4.xH_2O$ = ……

$$\text{molarity} = \frac{\text{concentration in g/l}}{\text{relative molecular mass}}$$

h). Calculate the number of moles of water of crystallization in one mole of $H_2C_2O_4.xH_2O$.

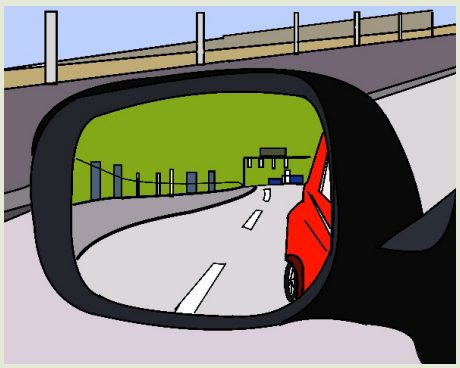
UNIT 5:

THE LENSES

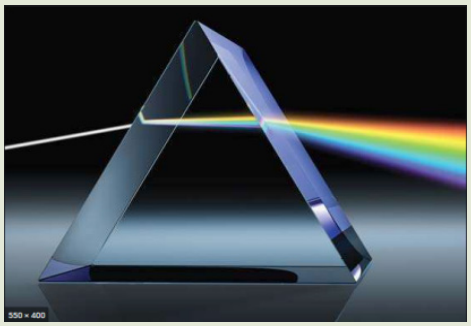
Key Unit competence: Interpret and solve problems solving in thin lenses and glass prism

Introductory Activity 5

Observe the images below and answer the following questions

A 

B

C 

D

a). Name image A,B,C and D and identify where they can be used in our daily life activities

b). What is the effect of light on A, B, C and D?

c). Analyze the phenomenon in figure C and write your observations in your notebook

d). How the images are formed on each of the above figures? Explain.

5.1. Types of lenses and their characteristics

Activity 5.1

1. Observe closely the lenses below and answer the asked questions:

a). Predict the materials used to make the above lenses. Explain your answer

b). How are the lenses different?

c). Classify these lenses according to your observation and explain the criteria used to classify them.

A lensis an optical device with perfect or approximate axial symmetry which transmits and refractslight, by converging or diverging the beam. Most lenses are spherical lenses: their two surfaces are parts of the surfaces of spheres. Lenses are classified by the curvature of the two optical surfaces.

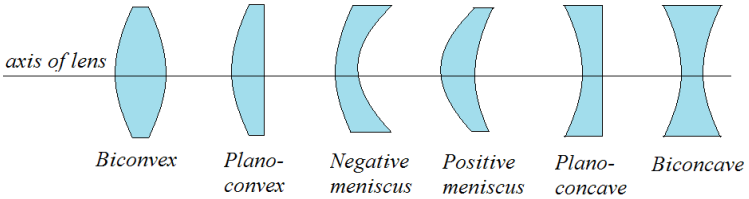


Figure 5.1: Classification of lenses

The lens which is thicker at the centre than at the edges is called a convex lens while the one which is thinner at its centre is known as a concave lens. The curved surface of the lens is called a meniscus. The lens in the human eye is thicker in the centre, and therefore it is a convex lens.

The light rays from the ray box change the direction after passing through the lens. They are therefore refracted by the lens. Hence, lenses form images of objects by refracting light.

You can see that the rays from the convex lens are getting closer and closer to a point. The rays are thus converging, and hence a convex lens is called a converging lens. You can also see that the refracted rays from the concave lens are spreading out. This kind of lens is called diverging lens. When light passes through a lens, refraction occurs at the two lens surfaces.

Using Snell's law and geometry, you can predict the paths of rays passing through lenses. To simplify such problems, assume that all refraction occurs on a plane, called the principal plane that passes through the center of the lens. This approximation, called the thin lens model, applies to all the lenses.

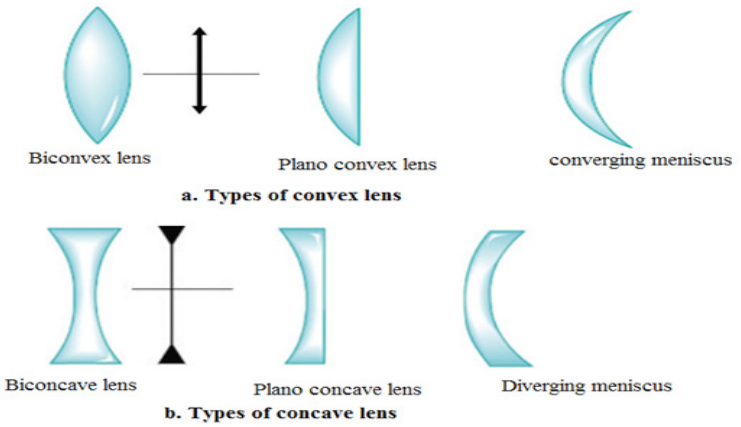
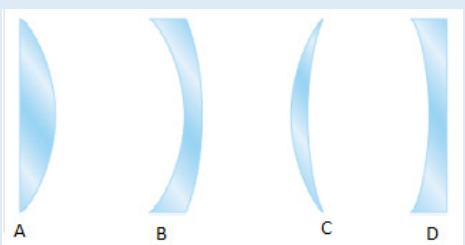


Figure 5.2: Types of lenses

Application activity 5.1

1. The cross sections of four different thin lenses are shown in Figure below.
   a). Which of these lenses, if any, are convex, or converging lenses?
   b). Which of these lenses, if any, are concave, or diverging lenses?
   c). Why do you make such decision at a and b?



2. Make a deep research on where this types of lenses are used in our daily life activities

5.2. Refraction of light through lenses

Activity 5.2

i. Hold a hand lens about 2 m from the window. Look through the lens. (CAUTION: Do not look at the sun).

   What do you see?

ii. Move the lens farther away from your eye.

   What changes do you notice?

iii. Now, hold the lens between the window and a white sheet of paper, but closer to paper.

iv. Slowly move the lens away from the paper towards the window. Keep watching the paper.

   What do you see? What happens as you move the paper?

   Do you see that an inverted image of trees outside is formed on the paper? How do you think the image is formed?

Lenses can be thought of as a series of tiny refracting prisms, each of which refracts light to produce an image. These prisms are near each other (truncated) and when they act together, they produce a bright image focused at a point.

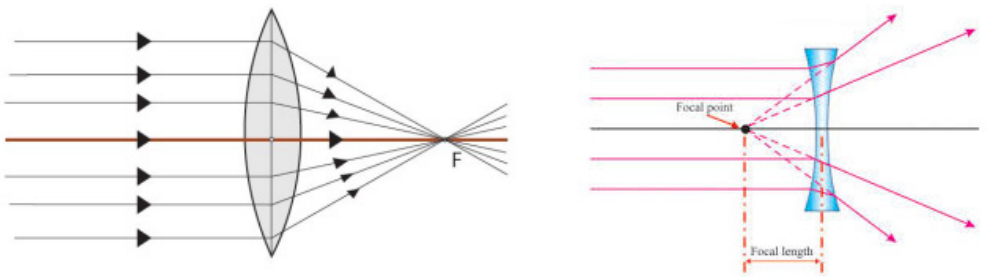


Figure 5.3: Action of converging and diverging lenses

5.2.1. Ray diagrams and properties of images formed by lenses

Notice that an image cannot be seen on the screen irrespective of the position of the object.

The nature of the image formed by a convex lens depends on the position of the object along the principal axis of the lens.

The principal focus of a lens plays an important part in the formation of an image by a lens since parallel rays from the object converge to it, and thus, we consider points F and 2F when describing the nature of the images formed by the lens.

These images can be larger or smaller than the object or same size as the object. When an image is larger than the object, we say that it is magnified and when it is smaller, we say that it is diminished.

Images which can be formed on the screen are Real images. Because light rays pass through these images, real images can be formed on the screen. All real images formed by the convex lens are inverted.

To determine an image point, we need to consider only the three rays indicated if Fig 4, which uses an arrow (on the left) as the object, and a converging lens forming an image to the right. These rays, emanating from a single point on the object, are drawn as if the lens were infinitely thin, and we show only a single sharp bend at the centre line of the lens instead of refraction at each surface. These three rays are drawn as follows:

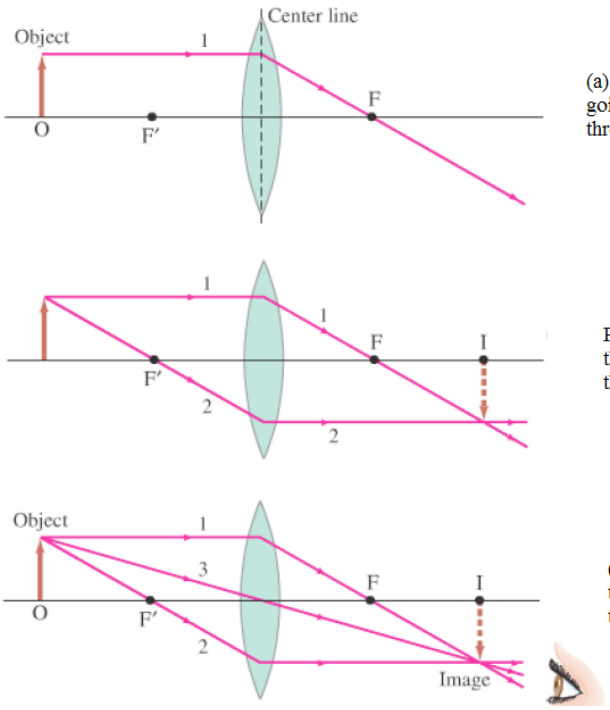
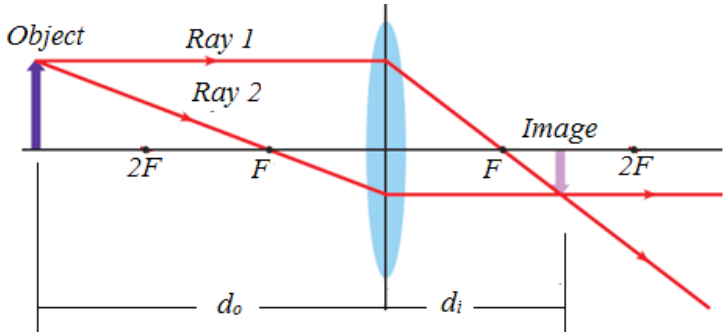


Figure 5.4: Ray tracing for convex lens

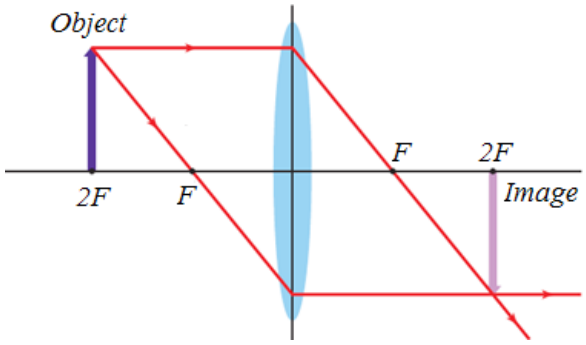
The point where these three rays cross is the image point for that object point. Actually, any two of these rays will suffice to locate the image point, but drawing the third ray can serve as a check.

5.2.2. Graphical determination of focal length of lenses

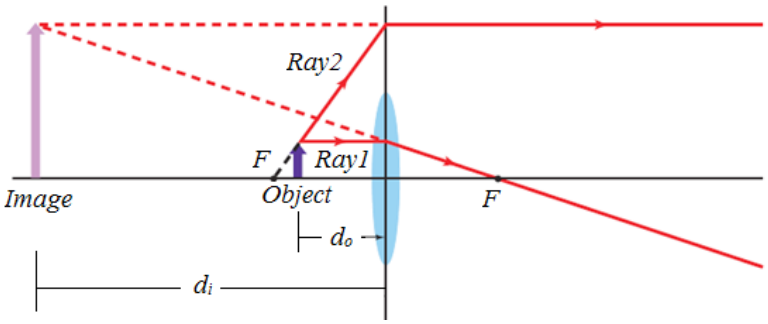
If the lens is biconvex or plano-convex, a collimated parallel or diverging beam of light travelling through the lens will be converged to a spot behind the lens. In this case, the lens is called a converging lens.



When an object is placed at a distance greater than the focal length from the lens, the image is real and inverted on the other side of the lens. When an object is placed at a distance equal to twice the focal length from the lens, the image is the same size as the object and inverted.



A convex lens forms a virtual image that is upright and larger compared to the object when the object is located between the lens and the focal point.



5.2.3. Graphical construction of images by diverging lens

If the lens is biconcave or plano-concave, a beam of light travelling to the lens axis and passing through the lens will be diverged.

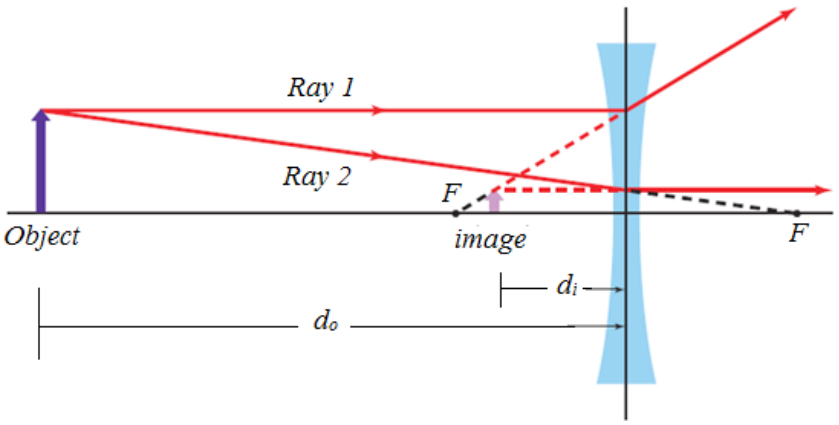


Figure 5.5: Finding image by ray tracing for a diverging lens

Concave lenses produce only virtual images that are upright and smaller compared to their objects.

When an object is between F and the lens, there is no image formed on the screen. The image formed is not real and is only seen by removing the screen and placing an eye in its position. We say that it is a virtual image. For a virtual image, rays appear to come from its position.

Unlike for a convex lens where the nature of the image depends on the position of the object, a concave lens gives only an upright, small, virtual image, and is situated between the principal focus and the lens for all positions of the object.

Application activity 5.2

1. Design an experiment to study images formed by convex lenses of various focal lengths. How does the focal length affect the position and size of the image produced?
2. An object of length 5 cm is placed at a distance 25 cm in front of a lens of focal length 10 cm.

   Use a ray diagram to construct the image of this object and state its properties if the lens is:
   a). Converging or convex;

   b). Diverging or concave.

5.3. The thin lens equations

Activity 5.3

Task 1:

i. Place a converging lens on a table while facing a window.

ii. Place a white screen behind the lens. Move the screen to and fro (forwards and backwards) until a sharp image of a distant object is seen on the screen.

iii. Discuss and write down the observation in your notebook.

iv. Measure the distance from the lens to the screen. What is this distance called?

Task 2:

v. Draw a ray diagram to determine the nature and position of the image of an object placed 10cm from a diverging lens of focal length 15cm.

vi. Using the above information, find the nature and position of the image using a lens formula. (Assign f a negative sign during your substitution).

What is the location of the image?

5.3.1. Convex lens

We now derive an equation that relates the image distance to the object distance and focal length of a thin lens. This equation will make the determination of image position quicker and more accurate than doing ray tracing. Let $\mathbf{d_o}$ be the object distance, the distance of the object from the center of the lens, and $\mathbf{d_i}$ the distance of the image from the center of the lens. And let $\mathbf{h_i}$ and $\mathbf{h_o}$ refer to the heights of object and image. Consider two rays shown in Figure below for a converging lens, assumed to be very thin.

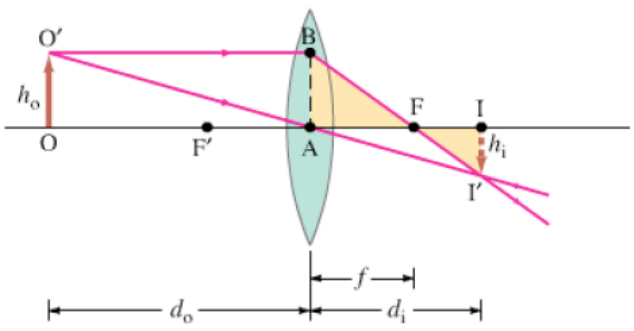


Figure 5.6: Deriving the lens equation for a converging lens

The right triangles FI'I and FBA are similar because angle AFB equals angle IFI'; so

$$\frac{h_i}{h_o} = \frac{d_i - f}{f}$$

Since length **AB= h_o**. Triangles OAO' and IAI' are similar as well. Therefore,

$$\frac{h_i}{h_o} = \frac{d_i}{d_o}$$

We equate the right sides of these equations (the left sides are the same), and divide by to obtain

$$\frac{1}{f} - \frac{1}{d_i} = \frac{1}{d_o} \quad \text{Or} \quad \frac{1}{f} = \frac{1}{d_i} + \frac{1}{d_o}$$

This is called the thin lens equation. It relates the image distance d_i to the object distance d_o and focal length **f**. It the most useful equation in geometric optics. If the object is at infinity, then $\frac{1}{d_o}=0$, so $\frac{1}{d_i}=f$. Thus the focal length is the image distance for an object at infinity, as mentioned earlier.

5.3.2. Concave lens

We can derive the Lens equation for diverging lens using the following figure.

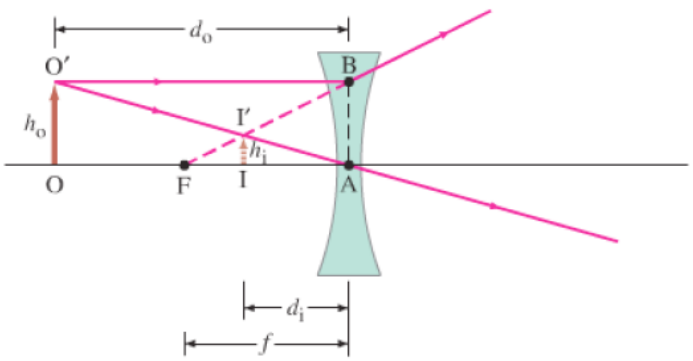


Figure 5.7: Deriving the lens equation for a diverging lens.

Triangles IAI' and OAO' are similar; and triangles IFI' and AFB are similar. Thus (noting that length **AB=h_o**)

$$\frac{h_i}{h_o} = \frac{d_i}{d_o} \text{ and } \frac{h_i}{h_o} = \frac{f - d_i}{f}$$

When we equate the right sides of these two equations and simply, we obtain

$$-\frac{1}{f} = \frac{1}{d_o} - \frac{1}{d_i}$$

This equation becomes the same as that of convex lens if we make **f** and $\mathbf{d_i}$ negative. That is, we take **f** to be negative for a diverging lens, and $\mathbf{d_i}$ negative when the image is on the same side of the lens as light comes from.

Thus

$$\frac{1}{f} = \frac{1}{d_o} + \frac{1}{d_i}$$

Will be valid for both converging and diverging lenses, and for all situations, if we use the following **sign conventions:**

1. The focal length is positive for converging lenses and negative for diverging lenses.
2. The object distance is positive if the object is on the side of the lens from which light is coming; otherwise, it is negative.
3. The image distance is positive if the image is on the opposite side from where light is coming; if it is on the same side, it is negative. Equivalently, the image distance is positive for real image and negative for a virtual image.
4. The height of the image, is positive if the image is upright, and negative if the image is inverted relative to the object. ($\mathbf{h_o}$ is always taken as positive).

5.3.3. Magnification

The magnification, m, of a lens is defined as the ratio of the image height to object height. From the above figures and the sign conventions, we have

$$m = \frac{h_i}{h_o} = -\frac{d_i}{d_o}$$

for an upright image the magnification is positive, and for an inverted image the magnification is negative.

5.3.4. The power of lenses

Whenever a ray of light passes through a lens it bends except when it passes through the optical centre. The degree of convergence or divergence of a lens is expressed as power. A lens of short focal length deviates the rays more while a lens of large focal length deviates the rays less. Thus the power of a lens is defined as the reciprocal of its focal length.

Power of a lens $P = \frac{1}{f}$

The unit of power is dioptre (D) $1D = 1m^{-1}$

From sign convention 1, it follows that the power of converging lens, in diopters, is positive, whereas the power of a diverging lens is negative. A converging lens is sometimes referred to as positive lens, and a diverging lens as a negative lens.

In case the lenses are combined

$P = P_1 + P_2 + ...$

$\frac{1}{F} = \frac{1}{f_1} + \frac{1}{f_2} + ...$ or

$F = \frac{f_1 f_2 ...}{f_1 + f_2 + ...}$

Example 1

An object is placed 32.0 cm from a convex lens that has a focal length of 8.0 cm.

a). Where is the image?

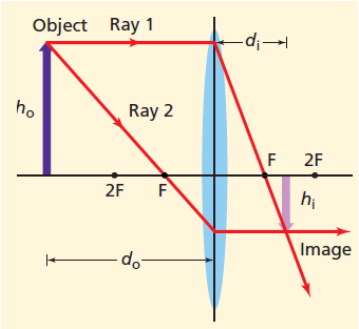
b). If the object is 3.0 cm high, how tall is the image?

c). What is the orientation of the image?

Solution

1. Analyse and sketch the problem

- Sketch the situation, locating the object and the lens.
- Draw the two principal rays



Known $\mathbf{d_o=32.0cm}$

$\mathbf{h_o=3.0cm}$

$\mathbf{f=8.0cm}$

Unknown: $\mathbf{d_i=?}$

$\mathbf{h_i=?}$

2. Solve for the unknown

a). Use the thin lens equation to determine $\mathbf{d_i}$

$$\frac{1}{f} = \frac{1}{d_o} + \frac{1}{d_i} \rightarrow d_i = \frac{fd_o}{d_o - f}$$

$$d_i = \frac{(8.0cm)(32.0cm)}{(32.0cm)-(8.0cm)} = 11cm$$

11cm away from the lens on the side opposite the object.

b). Use the magnification equation and solve for image height.

$$m = \frac{h_i}{h_o} = -\frac{d_i}{d_o}$$

$$h_i = \frac{-d_i h_o}{d_o} =$$

$$= \frac{-(11cm)(3.0cm)}{32.0cm} = -1.0cm \text{ (1.0 cm tall)}$$

c). The negative sign in part b means that the image is inverted.

5.3.5. Lens maker formula

The following assumptions are made for the derivation:

- The lens is thin, so that distances measured from the poles of its surfaces can be taken as equal to the distances from the optical centre of the lens.
- The aperture of the lens is small.
- Object point is considered.
- Incident and refracted rays make small angles.

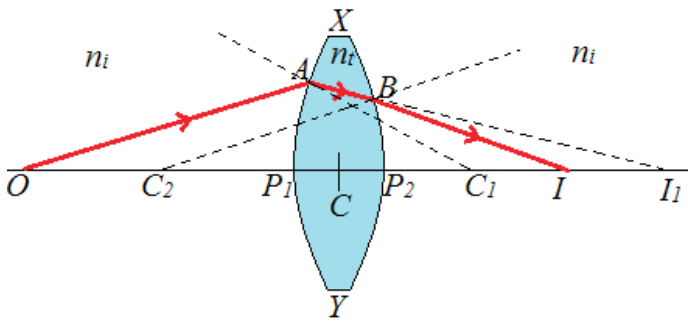


Figure 5.8: Derivation of lens maker formula.

Consider a convex lens of absolute refractive index n_t to be placed in a rarer medium of absolute refractive indexn_i.

Considering the refraction of a point object on the surface XP_1Y, the image is formed at I_1 who is at a distance of P_1I_1.

$CI_1 = P_1I_1$ (as the lens is thin)

$CC_1 = P_1C_1 = R_1$

$CO = P_1O$

The refracted ray from A suffers a second refraction on the surface XP_2Y and emerges along BI. Therefore I is the final real image of O. Here the object distance is

$$d_o = CO \approx CI_1 \approx R_1$$

It follows from the refraction due to convex spherical surface XP_1Y

$$\frac{n_i}{-d_o} + \frac{n_t}{CI_1} = \frac{n_t - n_i}{R_1}$$

Let $CI \approx P_2I = d_i$ be the Final image distance. Let R_2 be radius of curvature

of second surface of the lens.

$d_i = CI \approx CI_1 \approx R_2$

It follows from refraction due to concave spherical surface from denser to rarer medium that:

$$\frac{-n_t}{CI_1} + \frac{n_i}{d_i} = \frac{n_t - n_r}{R_2} = \frac{n_t - n_i}{-R_2}$$

Adding

$$\frac{n_i}{-d_o} + \frac{n_i}{d_i} = (n_t - n_i)\left(\frac{1}{R_1} - \frac{1}{R_2}\right)$$

$$\frac{1}{d_i} - \frac{1}{d_o} = \frac{(n_r - n_i)}{n_i}\left(\frac{1}{R_1} - \frac{1}{R_2}\right)$$

From equations $\frac{n_r}{n_i} = n$ and $\frac{1}{f} = \frac{1}{d_i} - \frac{1}{d_o}$

$$\frac{1}{f} = (n-1)\left(\frac{1}{R_1} - \frac{1}{R_2}\right)$$

This equation gives the lens maker's formula

Note:

- The lens maker's formula indicates that a convex lens can behave like a diverging one if $n_i > n_t$ i.e., if the lens is placed in a medium whose index is greater than the index of lens. Similarly a concave lens can be made convergent.
- The lens maker's formula can be derived for a concave lens in the same way.

Sign convention states that real is positive while virtual is negative. This should be put under consideration when one is using the lens formula to solve problems.

Application activity 5.3

1. Compute the composition and focal length of the converging lens which will project the image lamp, magnified 4 times, upon a screen 10.0 m from the lamp.
2. A lens has a convex surface of radius 20 cm and a concave surface of radius 40 cm and is made of glass of refractive index 1.54. Compute the focal length and the power of the lens, and state whether it is a converging or a diverging lens.

5.4. Defects of lenses, their corrections and applications of lenses

Activity 5.4

i. Place a white sheet of paper on a horizontal ground.

ii. Hold a glass ruler above the paper so as to focus rays from the sun on to the paper.

iii. Observe carefully the image formed on the sheet of paper.

iv. Repeat the above with the convex lens. What have you observed?

Aberration

A lens made of a uniform glass with spherical surface cannot form a perfect image. The spherical aberration is prominent image defect for a point source on the optical axis of such a lens. It arises because all rays through the lens are not focused to a common point.

The dependence of index of refraction of wavelength also causes the focal length, and thereby the image position, to depend on the color of the light. All simple lenses suffer from such chromatic aberration.

There are several different types of aberration which can cause the image to be an imperfect replica of the object (spherical aberration, chromatic aberration, coma, field curvature, barrel, pincushion distortion, astigmatism, etc).

Note that the image has coloured patches. This defect where by an image formed has coloured patches is called chromatic aberration. There are two kinds of defects; spherical aberration and chromatic aberration.

5.4.1. Spherical aberration

This arises in lenses of larger aperture when a wide beam of light incident on the lens, not all rays is brought to one focus.

As a result, the image of the object becomes distorted. The defect is due to the fact that the focal length of the lens for rays far from the principal axis are less than for rays closer to a property of a spherical surface and as a result, they converge to a point closer to the lens.

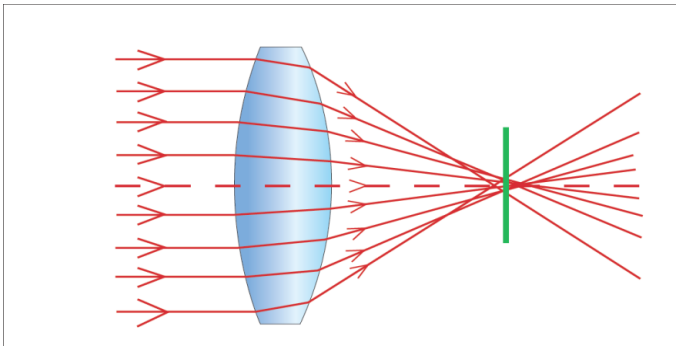


Figure 5.9: Spherical aberration

This defect can be minimized (reduced) by surrounding the lens with an aperture disc having a hole in the middle so that rays fall on the lens at a point closer to its principal axis. However, this reduces the brightness of the image since it reduces the amount of light energy passing through the lens.

5.4.2. Chromatic aberration

Chromatic aberration is caused by the dispersion of the lens material. Since, from the lens formulae, f is dependent upon n, it follows that different colours of light will be focused to different positions.

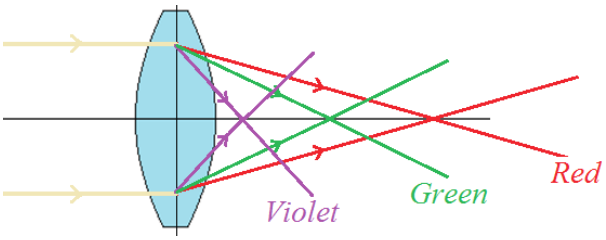


Figure 5.10: Chromatic aberration

Chromatic aberration can be minimised by using an achromatic doublet (or achromat) in which two materials with differing dispersion are bonded together to form a single lens.

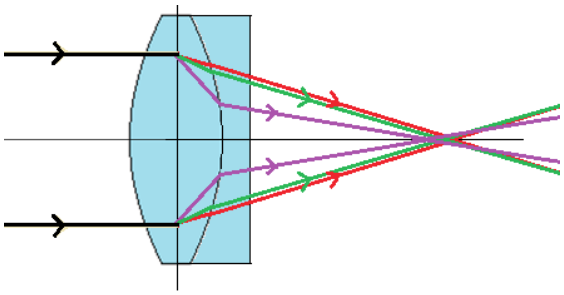


Figure 5.11: Correction of chromatic aberration

Application activity 5.4

i. How spherical and chromatic aberration can be reduced?

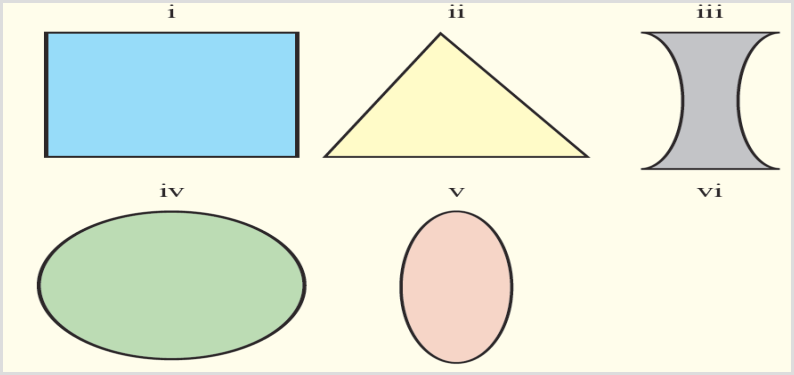
ii. If a plano-convex lens is used as objective lens in a telescope, how is its convex surface faced to minimize spherical aberration? Explain

5.5. Refraction through prisms

Activity 5.5

1. Have you ever heard of a prism? How does it look like?
2. Observe clearly the shapes of the glasses provided below and identify those with the shape of a prism.

   Explain your choices



3. With the help of a teacher, touch, observe and identify the real shape of the prism.
4. Examine the features of the one selected as a prism.

5.5.1. Terms associated with refraction through prism

In optics, a prism is transparent material like glass or plastic that refracts light. At least two of the flat surfaces must have an angle less than 90° between them. The exact angle between the surfaces depends on the application.

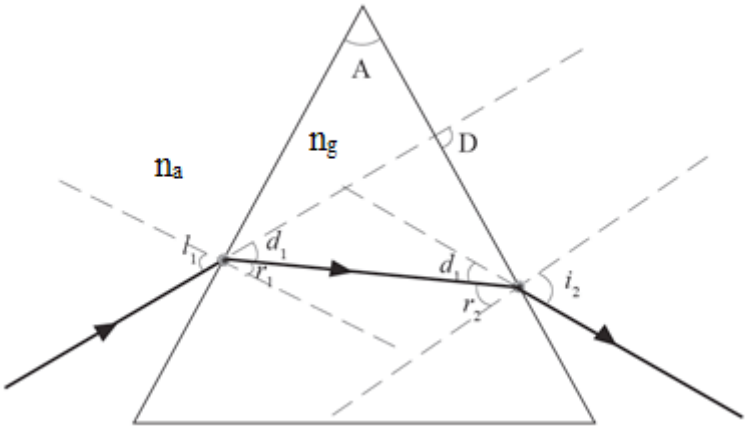


Figure 5.12: Refraction through a prism

From the figure 5.5.1, the following terms are used in refraction through prism.

Angle A: This is called refracting angle or angle of the prism. It is the angle between the inclined surfaces of the prism.

Angle i: This is the angle of incidence on the first face of the prism.

Angle r_1: This is the angle of refraction on the first face of the prism.

Angle r_2: This is the angle of refraction on the second face of the prism.

Angle I_2: This is the angle of emergence from second face of the prism. Sometimes this is denoted by letter e.

Angle D: this is the angle of deviation of rays through a prism.

n_a: is a refractive index in air.

n_g:is a refractive index in glass

Each section of a lens acts as a tiny glass prism. The refracting angles of these prisms decrease from the edges to its centre. As a result, light is deviated more at the edges than at the centre of the lens.

The refracting angles of the truncated prisms in a converging lens point to

the edges and so bring the parallel rays to a focus.

The truncated prisms of the diverging lens point the opposite way to those of the converging lens, and so a divergent beam is obtained when parallel rays are refracted by this lens because the deviation of the light is in the opposite direction.

The middle part of the lens acts like a rectangular piece of glass and a ray incident to it strikes it normally, and thus passes un-deviated.

Dispersion of light by a prism

Dispersion of light is the separation of a beam of light into its constituent colours. This takes place when a light beam passes through a dispersive medium. A beam of white light incident on a prism splits into its constituent colours to form "a visible spectrum"consisted of colours **violet, indigo, blue, green, yellow, orange and red.**

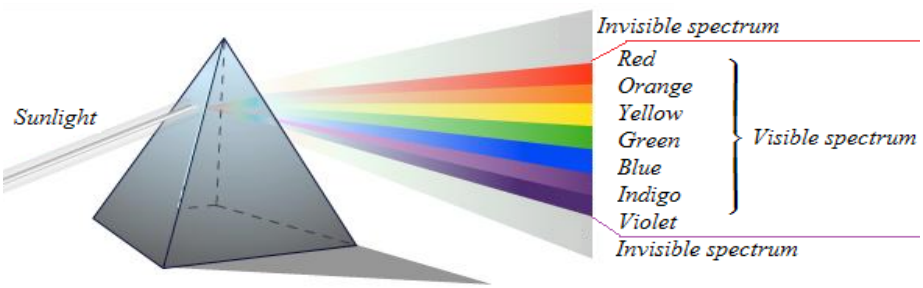


Fig 5.13: Dispersion of light by a prism

When a polychromatic light is incident on the first surface of the prism, each constituent colour gets refracted through a different angle. When these colours are incident on the second surface of the prism they are again refracted further.

Application activity 5.5.1

1. (i) Place a prism in the centre of a piece of paper so that its refracting surface is directly facing the windows in order to receive light from the sun.

   (ii) Place a white screen on the far side of the prism so that the refracted rays hit it.

   (iii) Observe what is formed on the screen.

   (iv) In brief, write in your notebook the observation.

2. Do a research to find other terms associated to the refraction light through prism.

5.5.2. Deviation by a prism and determination of refractive index

Activity 5.5.2

1. Have you ever heard of the word deviation? List down in your notebook at least two ways in which light can be deviated.
2. You are provided with a glass prism of refracting angle 60°, four optical pins, a white sheet of paper, a soft board and fixing pins.
   i. Place a prism on a white sheet of paper and mark its outline ABC.

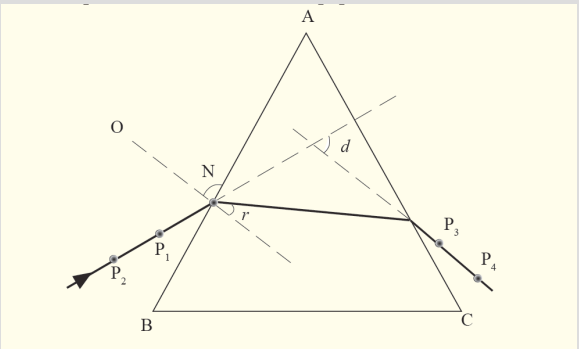


Figure 5.14: Diagram related to activity

   ii. Remove the prism and draw a normal line ON to face AB and draw a line making an angle of 10° to ON to represent the incident ray.
   iii. Place back the prism in its outline and fix pins P_1 and P_2 along the line.
   iv. While looking through the other face AC of the prism, fix pins P_3 and P_4 so that they appear in line with images of P_1 and P_2.
   v. Remove the prism and draw a line through P_3 and P_4 on to face AC of the prism.
   vi. Measure the angle of deviation d.
   vii. What is your observation on the direction of incident ray after refraction?

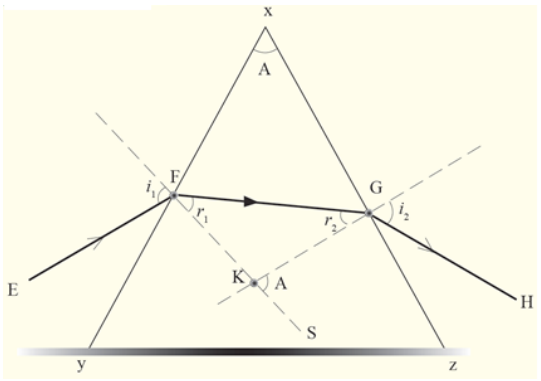


Fig 5.15: Refraction through prism

In the figure5.15, E F is a ray incident on the refracting surface YX of the prism XY Z from air and then to air from surface XZ of the prism. KF and KG are normal at the points of incidence and emergence of the ray respectively.

Now, from the geometry of quadrilateral XFKG,

< X FK + < X GK = 180^O and

A + <F K G = 180^O .....(1)

But since FKS is a straight line,

ΔFKG + ΔGKS = 180^O.....(2)

Comparing equation (1) and (2), it means that, ΔGKS = A.

Using ΔKFG, ΔGKS is an opposite exterior angle of r_1 and r_2

Thus, $r_1 + r_2$ = ΔGKS.

Hence, $r_1 + r_2$ = A

Note that given i_1, r_1 and i_2, r_2 as angles of incidence and refraction at F and G as shown and n is the prism refractive index, and then Snells law holds.

That is; sin i_1 = n sin r_1, and

sin i_2 = n sin r_2

The position and shape of the third side of the prism does not affect the refraction under consideration and so is shown as an irregular in figure.

Example

A ray of light falls from air to a prism of refracting angle 60^o at an angle of 30^o. Calculate the angle of emergence on the second face of the prism (Take refractive index of the material of glass, n_g = 1.5).

Solution

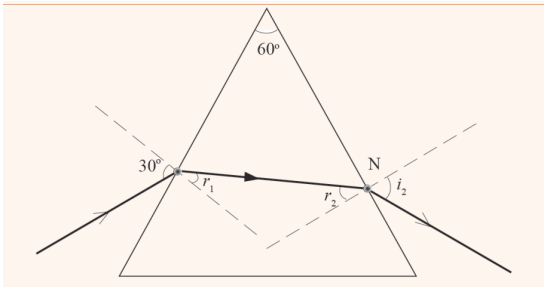


Figure 5.16: Ray through a prism

Using Snell's law,

$$n \sin i = \text{constant}$$

Thus

$$n_a \sin i_1 = n_g \sin r_1$$

$$1 \sin 30^0 = 1.5 \sin r_1$$

Therefore

$$\sin r_1 = \frac{0.5}{1.5}$$

$$r_1 = \sin^{-1} \frac{0.5}{1.5} = 19.5^0$$

But $r_1 + r_2 = A$

It follows that

$$r_2 = 60^0 - 19.5^0 = 40.5^0$$

Now, on the second face, $n_g \sin r_2 = n_a \sin$

$$n_g \sin r_2 = n_a \sin i_2$$

Thus,

$$1.5 \sin 40.5^0 = \sin i_2$$

So,

$$i_2 = \sin^{-1}(0.974) = 77^0$$

Hence the angle of emergence equals **77^0**

1. Deviation of light by a glass prism

Light can be deviated by reflection and refraction. Since a prism refracts light, it therefore changes its direction.

A prism deviates light on both faces. These deviations do not cancel out as in a parallel sided block where the emergent ray, although displaced, is parallel to the incident ray surface. The total deviation of a ray due to refraction at both faces of the prism is the sum of the deviation of the ray due to refraction at the first surface and its deviation at the second face.

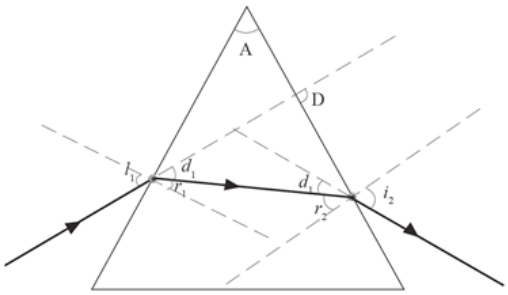


Figure 5.17: Diagram related to activity

Let d_1 and d_2 be angles of deviation at the first and second faces of the prism respectively.

Total deviation $D = d_1 + d_2$

Angle of deviation at the first face, $d_1 = i_1 - r_1$ and the angle of deviation at the second face,

$d_2 = i_2 - r_2$

Thus

$$\mathbf{D = i_1 - r_1 + i_2 - r_2 = (i_1 + i_2) - (r_1 + r_2)}$$

But $\mathbf{r_1 + r_2 = A}$

Therefore

$$\mathbf{D = (i_1 + i_2) - A}$$

N.B: The deviation D for a small angle prism is $D = A\,(n\text{-}1)$

The expression $D = A\,(n - 1)$ shows that for a given angle A, all rays entering a small angle prism at small angles of incidence suffer the same deviation.

2. Angle of minimum deviation

From the variation figure below, there is one angle of incidence which gives a minimum deviation. The experiment shows that this minimum deviation occurs when the angle of emergence is exactly equal to the angle of incidence and the two internal angles of refraction are equal. At this value, a ray passes symmetrically through the prism and the ray inside the prism is perpendicular to the directing plane.

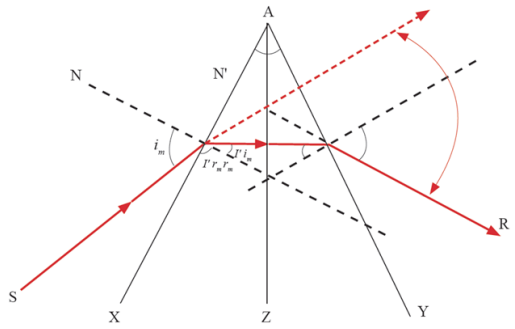


Figure 5.18: Minimum deviation

Since the angle of emergence i_2 = angle of incidence i_1, it follows that

$$i_1 = i_2 = i \text{ and } r_1 = r_2 = r$$

From deviation

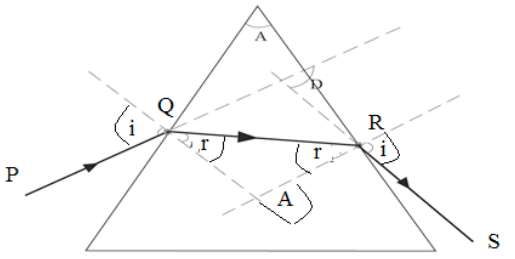
$$D = i_1 + i_2 - A$$

$$D_{min} = i + i - A$$

$$D_{min} = 2i - A$$

3. Determination of refractive index n of a material of the prism

A very convenient formula for refractive index, n, can be obtained in the minimum deviation case. The ray PQRS then passes symmetrically through the prism, and the angles made with the normal in the air and in the glass at Q, R respectively are equal. Suppose the angles are denoted by as shown. Then:



$$(i - r) + (i - r) = D_{min}$$

And (The refracting angle); so,

$$r \quad \frac{A}{2}$$

Substituting for we get:

$$2i = A + D_{min}$$

$$i = \frac{A + D_{min}}{2}$$

$$n = \frac{\sin i}{\sin r} = \frac{\sin \frac{A + D_{min}}{2}}{\sin \frac{A}{2}}$$

Example

A glass prism of refracting angle 60° has a refractive index of 1.5. Calculate the angle of minimum deviation for a parallel beam of light passing through it.

Solution:

$$n = \frac{\sin\left(\frac{A + D_{min}}{2}\right)}{\sin\left(\frac{A}{2}\right)}$$

$$1.5 = \frac{\sin\left(\frac{60 + D_{min}}{2}\right)}{\sin\left(\frac{60}{2}\right)}$$

Thus,

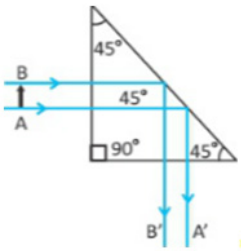
It follows that

$D_{min} = 2\sin^{-1}(0.75) - 60^0 = 97.2^0 - 60^0 = 37.2^0$

Hence, $\mathbf{D_{min} = 37.2^0}$

4. Applications of total internal reflection of light by a prism

Many optical instruments use right -angled prisms to reflect a beam of light through 90° or 180° (By total internal reflection) such as cameras, binoculars, periscope and telescope. One of the angles of right angled prism is 90°. When a ray of light strikes a face of prism perpendicular, it enters the prim without deviation and strikes the hypotenuse at an angle of 45°. Since the angle of incidence 45° is greater than critical angle of the glass which is 42°, the light is totally reflected by the prism through an angle of 90°. Two such prisms are used in periscope. The light is totally reflected by the prism by an angle of 180°. Two such prisms are used in binoculars.



Application activity 5.5.2

1. A ray of light is refracted through a 60 degree prism of ordinary glass making an incident of 35 degree with the normal i.e incident ray =35°.What will the angle of emergent ray and the angle of deviation measure?
2. What are the conditions for minimum deviation when a ray of light passes through a prism?
3. Show that the deviation produced by a small angled prism is independent with angle if incidence.

Skills lab 5

Determination of the focal length of a lens

Apparatus required:

1 torch bulb fitted in a bulb holder

1 switch

3 Torch cells

2 Cell holders

Connecting wires about 50cm long.

Wire gauze

Converging lens, of focal length 15cm.

Lens holder,

White screen

Metter rule.

Retort stand with its holding accessories.

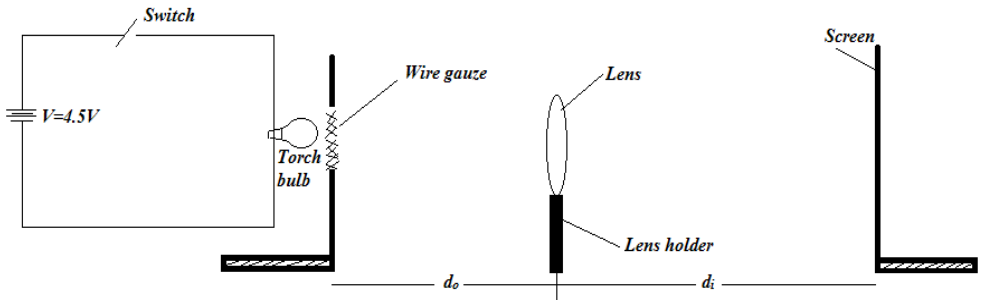
Instructions

In this experiment you will determine the focal length of the converging lens provided.

a). Mount the lens on the holder and place it facing a window

b). Place the screen behind the lens and adjust the screen until a clear image of a distant object is obtained

c). Measure and record the distance, x, between the lens and the screen.



d). Arrange the bulb, the wire gauze, lens and the screen as shown in figure.

e). Adjust the lens so that d_o equals $4.5x$.

f). Close the switch and move the screen until a clear image is formed on the screen.

g). Measure and record the distance d_i.

h). Repeat the procedures e) to g) above, for d_o equal to $4.0x$, $2.5x$, and $2.0x$.

i). Record your results in a suitable table including the values of $d_o d_i$ and $d_o + d_i$.

j). Plot a graph of $d_o d_i$ (along the vertical axis) against $d_o + d_i$ (along the horizontal axis).

k). Find the slope.

End unit assessment 5

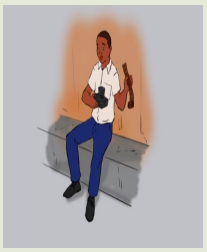
1. How does a convex lens work? Explain your reason
2. What will happen when a light ray travelling in glass is incident on an air surface? Explain your answer
3. What will happen When a light ray makes an angle of 90^0 while entering a glass slab? Explain your reason
4. What are factors responsible for the angle of deviation through prism depends?
5. A beam of monochromatic light is incident at i = 50° on one face of an equilateral prism, the angle of emergence is 40°.Calculate the angle of minimum deviation.
6. A 60^0 prism has a refractive index of 1.5. Calculate the angle of incidence for minimum deviation, the angle of emergence of light at maximum deviation and angle of maximum deviation.

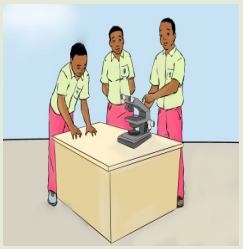
UNIT 6:

SIMPLE AND COMPOUND OPTICAL INSTRUMENTS

Key unit competence: Describe and use effectively simple and compound optical instruments

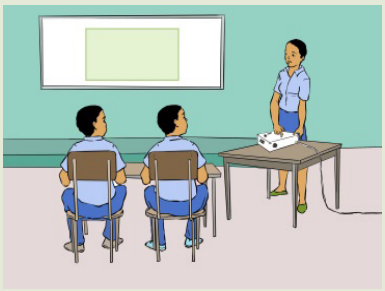
Introductory Activity 6

A 

B 

C

D

E 

1. After observing the above images, suggest the names of the available optical instruments.
2. Compared to the above images, provide other optical instrument you know.
3. Where do you think these instruments are used in our daily life activities?
4. As a student teacher, what those instruments will help you in teaching and learning process?

6.1. Introduction to optical instruments and their image formation

Activity 6.1

Task 1

i. What objects (things) do you see in the classroom?

ii. Move outside class and observe the kind of objects there, and write down at least five of them.

iii. Look at the distant objects. Are you able to examine the objects in a more detailed manner? Do you think you can be able to see these objects at night?

iv. When they are an eclipse, it's recommended to use some devices to observe it.

a). What are devices can you use to observe clearly that eclipse when it happened?

b). How can you take a photo of that eclipse?

TASK 2

i. Hold a hand lens at above the word Rwanda at a distance of about 4cm from the word.

ii. Move the lens farther away slowly from the word while observing the word through the lens.

iii. What changes do you notice after observing?

iv. write your observation in your notebook.

v. Compare the size of the word and the image of that word Rwanda.

Task 3

i. Have you ever heard or seen an instrument called a compound microscope? What is it used for?

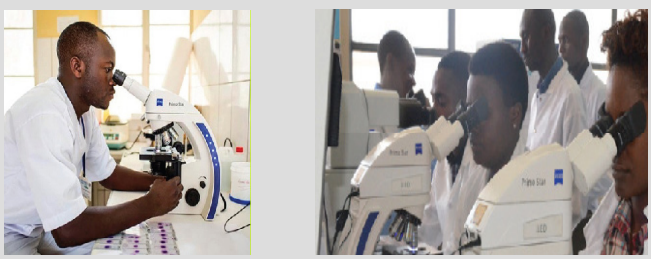
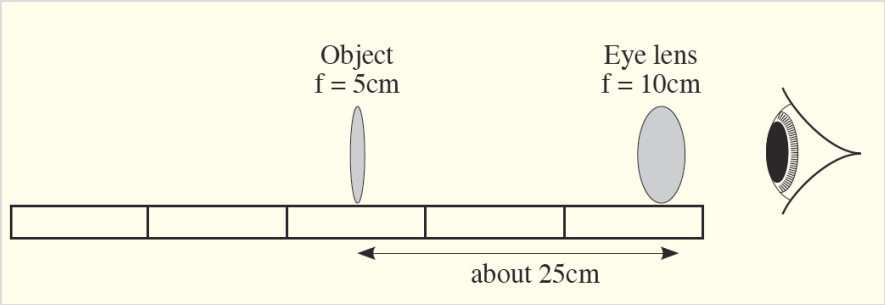


Figure 6.1: Different Rwandans using compound microscopes

ii. Observe the above pictures carefully and discuss places where a compound microscope is used in daily life.

iii. You are provided with two lenses of focal lengths 5cm and 10cm together with a half meter ruler and some plasticine.

a). Arrange the lenses as shown in the figure below.



b). Move the object until it appears in focus.

What do you notice about the image? Is it distorted? Is it coloured differently in any way?`

We use our eyes to see and view different objects. The eye cannot be used to view clearly these objects at night, and some distant objects or hidden objects. Objects which cannot be viewed by the eye can be focused using other instruments. All the instruments used to aid vision are called Optical instruments.

When the lenses or mirrors or both are arranged in a way, the arrangement can be used to observe objects in a more detailed manner. The arrangement makes what we call a compound optical instrument. The compound instruments include a compound microscope, telescopes, prism binoculars etc.

6.1.1. Image formation by a camera, simple and compound microscope

1. Lens camera

The image is upside down. The pin hole helps you to see the image of the object. This device is called a pin hole camera.

We have already seen that when an object is beyond 2F of a thin converging lens, the image formed is smaller than the object.

A camera consists of a light- tight box with a convex (converging) lens at one end and the film at the other end. It uses the convex lens to form a small, inverted, real image on the film at the back.

Figure 6.2: The lens camera

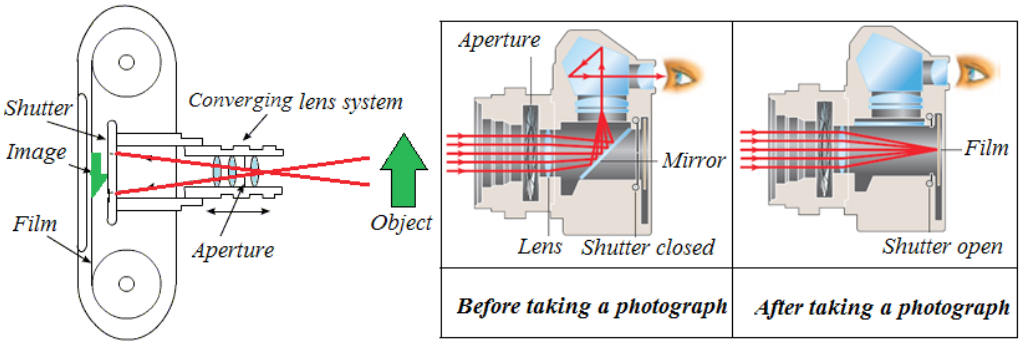


Figure 6.3: The lens camera

The lens focuses light from the object onto a light sensitive film. It is moved to and fro so that a sharp image is formed on the film. In many cameras, this happens automatically. In cheaper cameras, the lens is fixed and the photographer moves forwards and backwards to focus the object.

The diaphragm is a set of sliding plates between the lens and the film. It controls the aperture (diameter) of a hole through which light passes.

In bright light, a small aperture is used to cut down the amount of light reaching the film and in dim light, a large hole is needed. Very large apertures give blurred images because of aberrations

So the aperture has to be reduced to obtain clear images.

In many cameras, the amount of light passing through the lens can be altered by an aperture control or stop of variable width. This size of the hole is marked in f – numbers i.e 1.4, 2, 2.8, 4,5.6, 8, 11, 16, 22, 32. Defined as

$$\text{f - stop} = \frac{f}{D}$$

Where is the focal length of the lens and D is the diameter of the lens opening.

The smaller the f-number, the larger the aperture. An f-number of 4 means the diameter d of the aperture is ¼ the focal length, f of the lens. To widen the aperture, the f number should therefore be decreased.

The aperture also controls the depth of field of the lens camera. The depth of field is a range of distances in which the camera can focus objects simultaneously. This depth of field is increased by reducing the aperture.

This large depth of field ensures a large depth of focus. The depth of focus is the tiny distance the film plane can be moved to or from the lens without defocusing the image. A large depth of focus means that both near and far objects appear to be in focus at the same time which is obtained by a small hole in the diaphragm.

The shutter controls the exposure time of the film. It opens and closes quickly to let a small amount of light into the camera.

The exposure time affects the sharpness of the image. When the exposure time is short, the image is clear (sharp) but when it is long the image becomes blurred.

The film: This is where the image is formed. It is kept in darkness until the shutter is opened. It is coated with light sensitive chemicals which are changed by the different shades and colours in the image. When the film is processed, these changes are fixed and the developed film is used to print the photograph.

Note that a diminished image is always formed on the film and that the image of distant object is formed on a film at distance f from the lens. For near objects, the lens is moved further away from the film (closer to the object) to obtain a clear image. In this case, the film is at a distance greater than f of the lens.

2. Simple Microscope (Magnifying Glass)

The word Rwanda becomes larger and larger and finally disappears. This word gets larger because of the lens. We say that it is being magnified by the lens.

Notice that the hair (fur) and other small holes on the skin are seen clearly.

These parts of the skin are made bigger by the glass lens and this enables one to see them clearly. This lens which magnifies images is called a magnifying glass or a simple microscope.

A magnifying glass consists of a thin converging lens and it is used to view very small organisms or parts of organisms which cannot be easily seen by the naked eye.

It has to be held at the right distance between the eye and the object for the object to be in focus.A magnifying glass, the convex lens bends the parallel rays so that they converge and create a virtual image on your eyes' retinas.

Formation of images by a magnifying glass

We have already seen that when an object is between the lens and its principal focus, the image formed is magnified and upright. So, a magnifying glass forms a virtual, upright, magnified image of an object placed between the lens and its principal focus.

The image is at the least distance of vision since the eyes are not strained and the magnifying glass is said to be in normal adjustment.

A microscope is in normal adjustment if the final image is formed at the near point, and it is not in normal adjustment if the final image is at infinity.

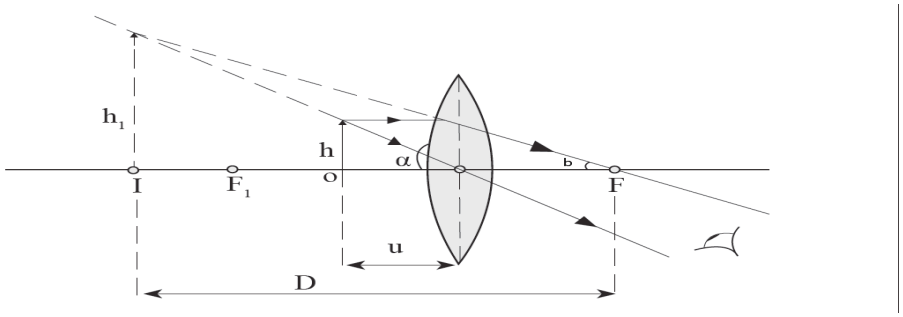


Figure 6.4. Image formed by a magnifying glass

3. The compound microscope and its images formation

The compound microscope is used to detect small objects; is probably the most well-known and well-used research tool in biology. In daily life, microscopes are used in hospitals, in biology laboratories

By arranging the lenses as above, you have actually made a compound microscope.

We have already seen how a single lens (magnifying glass) can be used to magnify objects.

However, to give a higher magnifying power, two lenses are needed. This arrangement of lenses makes a compound microscope. It produces a magnified inverted image of an object.

A compound microscope is used to view very small organisms that cannot be seen using our naked eyes for example micro organisms.

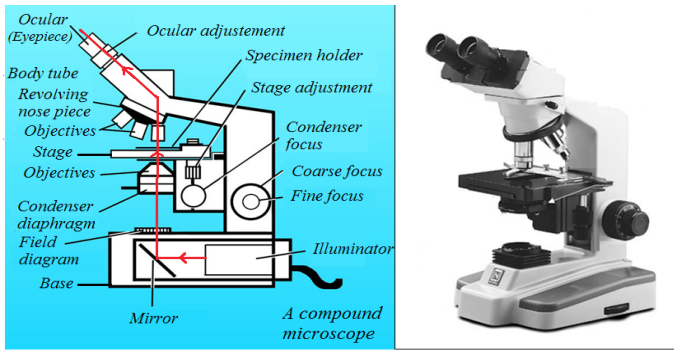


Figure 6.5. A compound microscope

A compound microscope consists of two convex lenses of short focal lengths referred to as the objective and the eye piece. The objective is nearest to the object and the eye piece is nearest to the eye of the observer.

The object to be viewed is placed just outside the focal point (at a distance just greater than the focal length) of the objective lens. This objective lens forms a real, magnified, inverted image at a point inside the principal focus of the eye piece.

This image acts as an object for the eye piece and it produces a magnified virtual image. So the viewer, looking through the eye piece sees a magnified virtual image of a picture formed by the objective i.e of the real image.

Image formation in a compound microscope

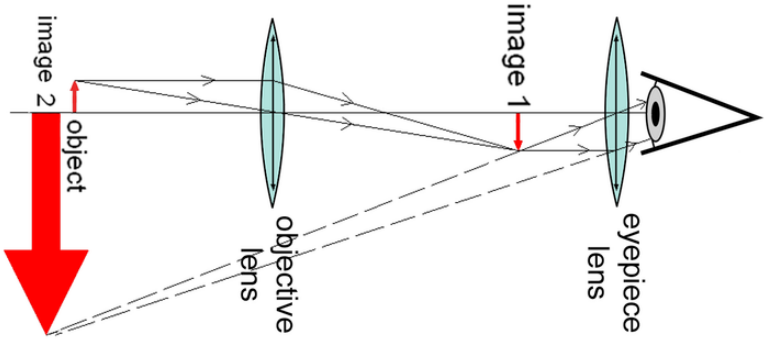


Figure 6.6: Images formed by a compound microscope

An objective lens L_1 forms a real magnified image 1, of an object O just placed outside its principal focus F_0. Image 1, is formed just inside the principal focus F_e of the eye piece L_2, which acts as a magnifying glass and produces a magnified, virtual image 2 of image 1.

Application activity 6.1

1. In groups, discuss the differences and similarities between the lens camera and the human eye.
2. (i) Make a paper box and carefully use a pin to make a tiny hole in the centre of the bottom of the paper box.

   (ii) Place a piece of wax paper on the open end of the box. Hold the paper in place with the rubber band.

   (iii) Turn off the room lights. Point the end of the box with a hole in a bright window.

   (iv) Look at the image formed on the wax paper. Which kind of image have you seen? Is it upside down or right side up. Is it smaller or larger than the actual object? What type of image is it?
3. (i) Carefully place a magnifying glass above some prints on a piece of paper and adjust it until they are seen clearly.

   (ii) Make sure that you don't feel any strain in the eye while you are observing.

   (iii) What do you think is the position of the image from the eye? Explain your reason

   (iv) Explain when the microscope can be in normal adjustment
4. Research on other types of optical instrument and the formation of their images

6.2. Magnifying power of optical instruments and its Calculation

Activity 6.2

1. What do you understand by the term magnification of an optical instrument?
2. Research how to make a magnification of camera, projector, microscope and telescope.

6.2.1. Magnifying power of a simple microscope

We have already seen that the size of the image depends on the angle subtended by the object on the eye called the visual angle. Thus, the magnifying power depends on the visual angle.

It is defined as the ratio of the angle subtended by the image to the lens to the angle subtended by the object at the near point N of the eye.

$$M = \frac{\theta'}{\theta}$$

Where θ' and θ are shown in figure below

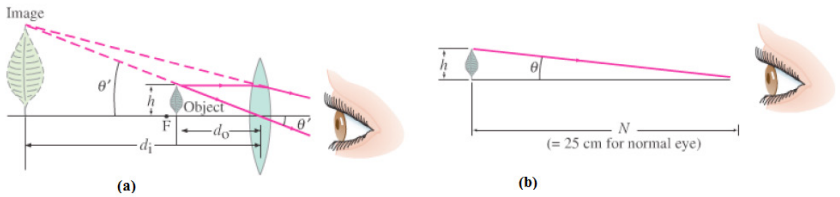


Figure 6.7: Leaf viewed (a) through a magnifying glass, and (b) with the unaided eye. The eye is focused at its near point in both cases.

We can write M in terms of the focal length by noting that $\theta = h/N$ (fig. 6.7 (a)) and $\theta' = h/d_0$ (Fig. 6.7 (b)), where h is the height of the object and we assume the angles are small so θ' and θ equal their sines and tangents. If the eye is relaxed (for least eye strain), the image will be at infinity and the object will be at infinity and the object will be precisely at the focal point.

Then $d_0 = f$ and $\theta' = h/f$. Thus

$M = \frac{\theta'}{\theta} = \frac{h/f}{h/N} = \frac{N}{f}$ (Eye focused at ∞, $N = 25\,cm$ for normal eye.)

We see that the shorter the focal length of the lens, the greater the magnification.

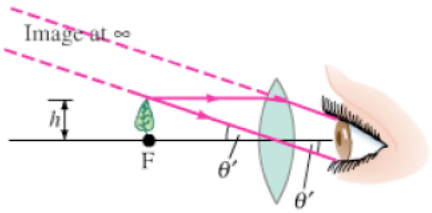


Figure 6.8: With the eye relaxed, the object is placed at the focal point, and the image is at infinity.

The magnification of a given lens can be increased a bit by moving the lens and adjusting your eye so that it focuses on the image at the eye's near point. In this case, $d_i = -N$ if your eye is very near the magnifier. Then the object distance d_o is given by:

$$\frac{1}{d_0} = \frac{1}{f} - \frac{1}{d_i} = \frac{1}{f} + \frac{1}{N}$$

We see from this equation that $d_0 = fN/(f+N) < f$, since $N/(f+N)$ must be less than 1. With $\theta' = h/d_0$ the magnification is

$$M = \frac{\theta'}{\theta} = \frac{h/d_0}{h/N} = \frac{N}{d_0} = N\left(\frac{1}{f} + \frac{1}{N}\right)$$

Or

$$M = \frac{N}{f} + 1$$ (Eye focused at near point, N; $\boldsymbol{N = 25cm}$ for normal eye)

We see that the magnification is slightly greater when the eye is focused at its near point, rather than relaxed.

Example:

A converging lens of 8 cm focal length is used as a "jeweler's loupe", which is a magnifying glass. Estimate (a) the magnification when the eye is relaxed, and (b) the magnification if the eye is focused at its near point N = 25 cm.

Solution

a). With the relaxed eye focused at infinity,

$$M = \frac{N}{f} = \frac{25cm}{8cm} \approx 3$$

b). The magnification when the eye is focused at its near point, and the lens is near the eye, is

$$M = \frac{N}{f} + 1 = \frac{25cm}{8cm} + 1 \approx 4$$

Uses of magnifying glass

Magnifying glasses have many different uses. Some people use it for fun activities such as starting fires, or use the lens to help them read. You can start a fire with a magnifying glass when the sun rays are concentrated on the lens.

Some retail stores sell reading glasses with the double convex lens. In everyday life, magnifying glasses can be used to do a variety of things. The most common use for magnifying glasses would be how scientists use them; they use magnifying glasses to study tiny germs.

6.2.2. Angular magnification (magnifying power) of a compound microscope

The overall magnification of a microscope is the product of the magnifications produced by the two lenses. The image 1is formed by the objective lens is a factor m_0 greater than the object itself. From the figure below,

$$m_0 = \frac{h_i}{h_o} = \frac{d_i}{d_o} = \frac{l - f_e}{d_0}$$

Where and are the object and image distances for the objective lens, l is the distance between the lenses, and we ignored the minus sign which only tells us that the image is inverted.

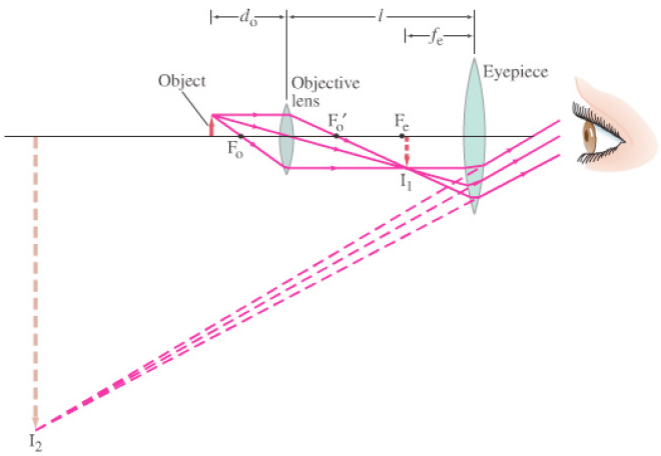


Figure 6.9: Overall magnification of a microscope

We set $d_i = l - f_e$, which is true only if the eye is relaxed, so that the image 1 is at the eyepiece focal point $\boldsymbol{F_e}$. The eyepiece acts like a simple magnifier. If we assume that the eye is relaxed, the eyepiece angular magnification $\boldsymbol{M_e}$ is

$$M_e = \frac{N}{f_e}$$

Where the near point $\boldsymbol{N=25cm}$ for the normal eye. Since the eyepiece enlarges the image formed by the objective, the overall angular magnification M is the product of the lateral magnification of the objective lens, $\boldsymbol{m_o}$, times the angular magnification, $\boldsymbol{m_e}$, of the eyepiece.

$$M = m_0 M_e = \left(\frac{N}{f_e}\right)\left(\frac{l - f_e}{d_0}\right) \approx \frac{Nl}{f_e f_0}$$

The approximation is accurate when f_e and f_o are small compared to l, so $l - f_e \approx 1$ and $d_0 \approx f_0$.This is a good approximations, which are obtained when f_0 and f_e are very small.

Example:

A compound microscope consists of a 10x eyepiece and a 50 x objective 17.0 cm apart. Determine (a) the overall magnification, (b) the focal length of each lens, and (c) the position of the object when the final image is in focus with the eye relaxed. Assume a normal eye, so $\mathrm{N} = 25\ \mathrm{cm}$.

Solution

a). The overall magnification is $(10\times)(50\times) = 500\times$

b). The eyepiece focal length is $f_e = \frac{N}{M_e} = \frac{25cm}{10} = 2.5\,cm$

$$d_0 = \frac{l - f_e}{m_0} = \frac{17cm - 2.5cm}{50} = 0.29\,cm$$

Then, from the thin lens equation for the objective with

$$d_i = l - f_e = 17\,cm - 2.5cm = 14.5\,cm$$

$$\frac{1}{f_0} = \frac{1}{d_0} + \frac{1}{d_i} = \frac{1}{0.29cm} + \frac{1}{14.5cm} = 3.52\,cm^{-1}$$

So $f_0 = \frac{1}{3.52\,cm^{-1}} = 28\,cm$

c). We just calculated $d_0 = 0.29\,cm,$, which is very close to f_0.

Application activity 6.2

1. An eye specialist prescribe spectacles having combination of convex lens of focal length 40 cm in contact with a concave lens of focal length 25 cm . Calculate the power of this lens combination
2. The near vision of an average person is 25 cm. To view an object with an angular magnification of 10, what should be the power of the microscope?
3. The focal length of the objective lens is 0.9 cm and the focal length of the ocular lens is 2.5 cm. The microscope is used by a normal eye without accommodation and the overall magnification is 90 times. Determine the distance between the object and the objective lens. N = 25 cm.

6.3. Astronomical Telescope

Activity 6.3

Task 1

You have heard in your early secondary that there are some heavenly and distant earthly bodies that cannot be seen by our naked eyes. How did the people know that there exist such bodies?

Which instrument do you think is used to see these bodies and to observe what takes place on these bodies?

Why do you think it is difficult to see distant objects using our eyes? Explain your reason

Task 2

i. Hold a convex lens of focal length 5cm close to your eye.
ii. Hold another lens of focal length 20cm at an arm's length.
iii. Use the lens combination to view distant objects.
iv. Adjust the distance of the farther lens until the image is clear (take care not to drop the lenses). What type of image do you see? Explain your demonstration

Telescopes are instruments used to view distant objects such as stars and other heavenly bodies. Distant objects are difficult to see because light from them has spread out by the time it reaches the eyes, and since our eyes are too small to gather much light.

6.3.1. Refracting telescopes

A refractor or a refracting telescope uses a combination of lenses to form the images of distant objects. Refractors have been criticized for their chromatic and spherical aberration.

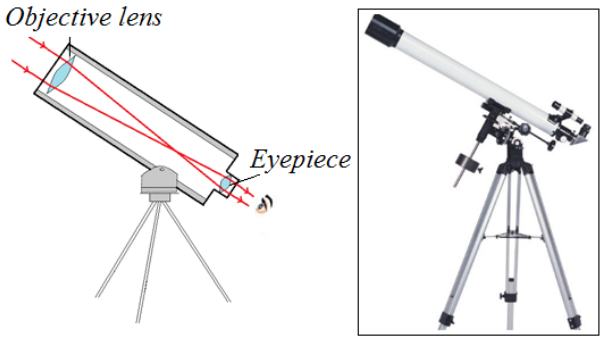


Fig 6.10: Telescopes

The above lens combination is a refracting telescope. It is called a refracting telescope because it forms an image of the object by refracting light. Therefore, Refracting telescopes use lenses and they form images by refraction of light.

The telescope made in the above activity is called an astronomical telescope. It consists of two convex lenses, the objective lens of long focal length and an eyepiece lens of short focal length.

When the eyes are relaxed, the image is at infinity and the telescope is in normal adjustment. Therefore, an astronomical telescope is in normal adjustment when the final image is formed at infinity.

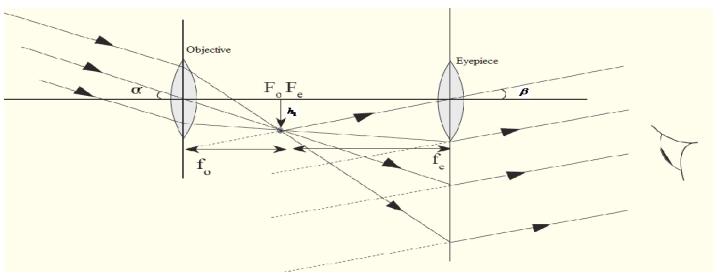


Figure 6.11: An astronomical telescope in normal adjustment

The rays of light coming from a distant object form a parallel beam of light. This parallel beam is focused by the objective lens and it forms a real, diminished image at its principal focus F_o. The eyepiece is adjusted so that this image lies in its focal plane. This image acts as the object for the eyepiece and the eyepiece produces the image at infinity.

Note that in normal adjustment, the eye is relaxed or un accommodated when viewing the image. In this case, the eye has minimum strain.

6.3.2. Magnifying power or angular magnification of an astronomical telescope

The magnifying power of a telescope is the ratio of the angle subtended by the image to the eye when the telescope is used to the angle subtended at the unaided eye by the object.

Since the telescope length is very small compared with the distance of the object from either lens, the angle subtended at the unaided eye by the object is the same as that subtended at the objective by the object.

In **normal adjustment**, the magnifying power (angular magnification) of an astronomical telescope is given by:

$$M = \frac{\text{angle subtended at the eye by the final image at infinity}}{\text{angle subtended at the objective by the object}}$$

Let β = angle subtended by the final image at the eye and α = angle subtended by the object at the objective.

Hence, $M = \frac{\beta}{\alpha}$

Supposing that the eye is close to the eye piece and h_1 is the image of image I_1 formed by the objective, then $tan\ \alpha \approx \alpha$ and $tan\ \beta \approx \beta$

So $\tan \alpha = \frac{h_1}{f_e}, and \quad \tan \beta = \frac{h_1}{f_e}$

It follows that: $M = \frac{\frac{h_1}{f_e}}{\frac{h_1}{f_o}} = \frac{h_1}{f_e} X \frac{f_o}{h_1}$

Therefore, $M = \frac{f_o}{f_e}$

Note that:

i. From the above expression, M is high when eyepiece focal length f_e is short and the objective focal length f_0 is long. This explains the fact why the objective lens of long focal length and the eyepiece lens of short focal

length are used during the construction of the astronomical telescope.

ii. For a telescope in normal adjustment, the separation of the objective and the eye piece is $f_0 + fe$

Application activity 6.3

1. An astronomical telescope, in normal adjustment position, has magnifying power 5. The distance between the objective and the eye-piece is 120 cm. Calculate the focal lengths of the objective and of the eye-piece.
2. An astronomical telescope has an objective lens of focal length 120cm and an eye piece of focal length 5 cm. If the telescope is in normal adjustment, what is;
   i. The angular magnification (magnifying power)
   ii. The distance of separation between the two lenses?

6.4. Human eye as single lens system

Activity 6.4

Task 1

i. Go outside class and view the trees around.

ii. Are the trees of the same height?

Notice that some trees at a distance look shorter than the nearby trees when it is not the case? Why do you think it is so? Discuss and write down in your notebook about your observation

Task 2

i. Hold a book at an arm's length and move it closer to find the nearest distance that you can focus the words clearly without straining your eyes.

ii. Approximate the distance between your eyes and the book.

iii. What does this distance represent?

iv. Look at the trees around your school.

v. Now, try to look at objects far from the school.

vi. Are you able to focus the distant objects?

vii. Measure this distance from the object to your eye.

Write down your observation in the notebook.

6.4.1. Structure of human eye

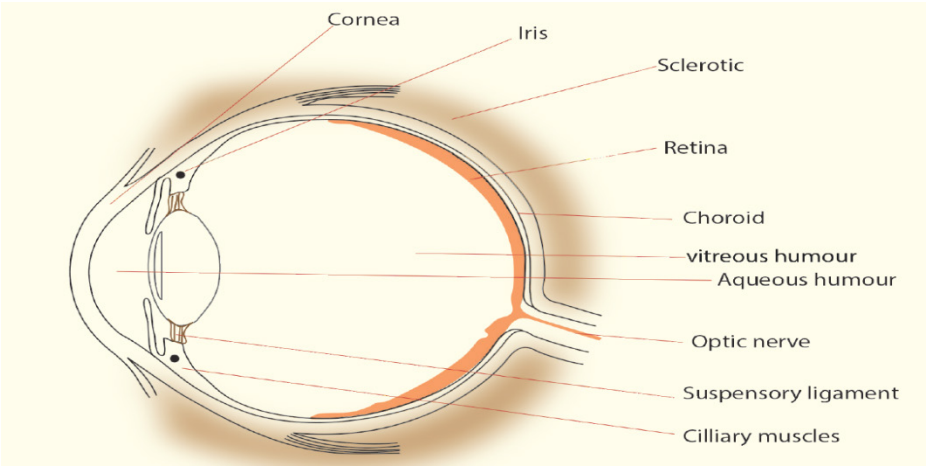


Fig 6.12: Anatomy of an eye

Functions of the parts of the eye

The cornea: It is made out of a fairly dense, jelly like material which provides protection for the eye, and seals in the aqueous humour. It also provides most of the power of the eye (59 Dioptres), having about46 Dioptres. So it provides most of the bending of light rays.

The aqueous humour: This is a watery liquid that helps to keep the cornea in a rounded shape, similar to that of a lens.

The iris: This controls the amount of light entering the eye. The amount of light that enters the eye is one of the factors determining how focused an image is on the retina. The brighter the light the eye is exposed to, the smaller the iris' opening will be. The brighter the light the eye is exposed to, the smaller the iris' opening will be. The iris is the coloured part of the eye as seen from the outside. The iris opening or a gap through which light passes is called a pupil.

The lens: This is used to focus an image on the retina. It controls the bending of light rays by change of its shape, a process called accommodation, which is done by the ciliary muscles.

The ciliary muscles: These control the thickness of the lens during focusing. By contracting or squeezing the lens, they make it thicker and vice versa. Because the power of the lens is directly related to its thickness, the ciliary muscles change the power of the lens by their movement.

The retina: This is the light sensitive part of the eye and it is where images are formed. It contains millions of tiny cells which are sensitive to light. The cells send signals along the optic nerve to the brain. So the retina is the screen of the eye and the image is formed by successive refraction at the surfaces between air, the cornea, the aqueous humour, the lens and vitreous humour. The retina is black, which prevents any light rays that hit it from reflections and thereby changing the image.

The vitreous humour: This is a jerry like substance that helps the eye to keep its round shape. It is very close in optical density to the lens material.

The yellow spot: This is a small area on the retina where the sharpest image, that is, the finest detail can be seen.

The optic nerve: This is the nerve that transmits images received by the retina to the brain for interpretation. The part of the eye where the optic nerve joins the retina is called the blind spot because no images can be observed at at this point.

6.4.2. Visual angle and accommodation of the eye

a. Visual angle

The height of an object depends on the angle of elevation of its top from the eye. The larger the angle, the taller the objects. This angle is called the visual angle. The visual angle is the angle subtended at the eye by an object.

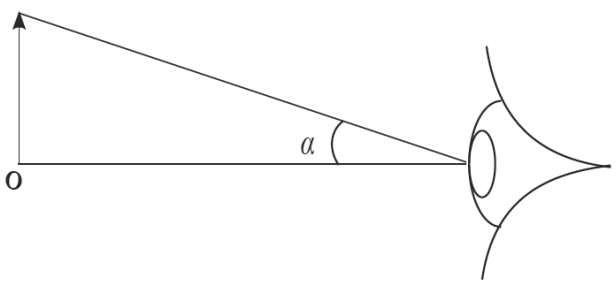


Figure 6.13: Visual Angle

Example:

Let us observe the flame of a candle: its two extremities A and B are seen by an eye at a certain angle. Expressed in radians, this angle has a measure:

$$M = \frac{\beta}{\alpha}$$

This angle decreases when the distance D increases and increases when the distance D decreases. It also increases when the length AB increases and decreases when AB decreases. We call it visual angle of the object.

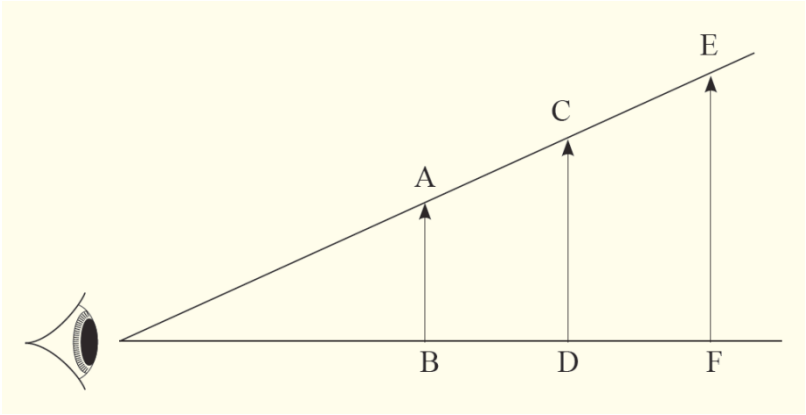


Figure 6.14: Visual Angle of trees

Objects that subtend the same angle at the eye appear to be of the same size as viewed by the eye.

The apparent size of an object depends on the size of its image on the retina. For example, the two objects above; AB and CD appear to have same size because they subtend the same angle θ at the eye. This explains why trees in a forest appear to have the same height.

It is defined as the ratio of the apparent size of the final image i.e angle subtended by the image at the position of eye to the apparent size of the object i.e angle subtended by the object at the eye.

We have seen that we can use other instruments apart from the eye to aid vision. So, angular magnification or magnifying power of an optical instrument can also be defined as the ratio of the angle subtended at the eye by the image when the optical instrument is used to the angle subtended by the object at the unaided eye (when the instrument is not used).

If β is the angle subtended at the eye by the image and α is the angle subtended by the object at some distance by unaided eye, then the angular magnification $M = \frac{\beta}{\alpha}$

b. Accommodation of the eye

Accommodation of the eye is the ability of the eye to see near and distant objects. The eye is capable of focusing objects at different distances by automatic adjustment of the thickness of the eye lens which is done by the ciliary muscles.

To focus a distant object, the eye lens is made thinner, so less powerful, and the rays from the object are brought to focus on the retina by the eye lens. In this case, the ciliary muscles are relaxed and pull the lens.

For nearer objects, the eye lens must be made thicker and hence more powerful so that the rays from the near object can be brought to a focus on the retina. In this case, the ciliary muscles tighten and squeeze the lens.

The near point of the eye is the nearest point that can be focused by the un aided eye. It is a closest distance that the 'normal' human eye can observe clearly; without any strain to the eye. It is called the least distance of distinct vision. The near point of a normal eye is 25 cm.

The distance from a distant object to the eye is the far point of the eye. The far point of the eye is infinity. The far point is the farthest point that can be focused by the eye.

The distance of 25 cm from the eye is called distance of most distinct vision or least distance for distinct vision. The range of accommodation of the normal eye is thus from 25 cm to infinity. This range is based upon the average human eye which has an age of 40 years.

Young persons have a much wider range but the average 70 year – old has a reduced range.

People with normal vision can focus both near and distant objects.

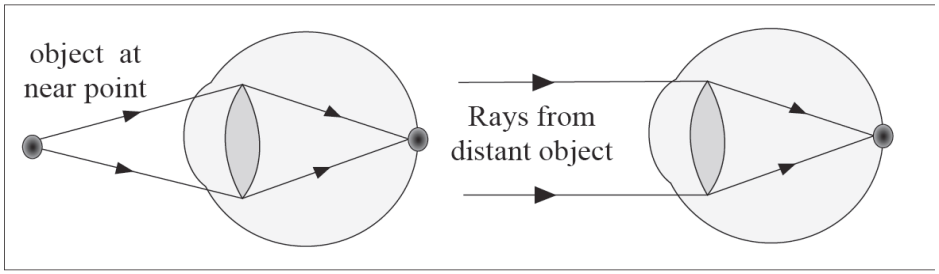


Figure 6.15: Near and far points of a normal eye

Formation of an image by the eye

Light enters the eye through the transparent cornea, passes through the lens and is focused on the retina. The retina is sensitive to light and sends messages to the brain for interpretation. Although the image is inverted, the brain interprets it correctly.

Application activity 6.4

1. Do you think that the angle of vision is the same for all people? Explain your reason.
2. What is the angle of vision for a normal healthy individual? Explain your answer.
3. When we increase the distance of an object from the eye, what happens to the image distance in the eye?
4. What do you understand by the term “cataract”? How is it happened in human eye? Explain your answer
5. Make a research on other parts of human eye and provide their functions for what you would find out.
6. How do you call Photo-receptors that detect color? Explain your answer.
7. Which part of the eye that regulates the amount of light entering the eye and gives the eye its color.

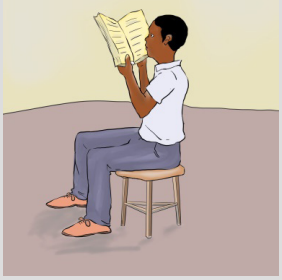
6.5. Defects of vision and their correction

Activity 6.5

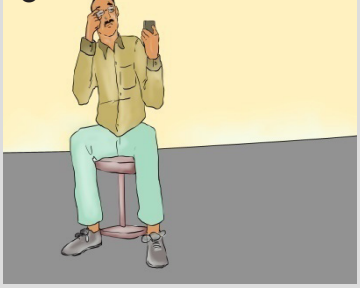
Observe the pictures below and answer the related questions.

A

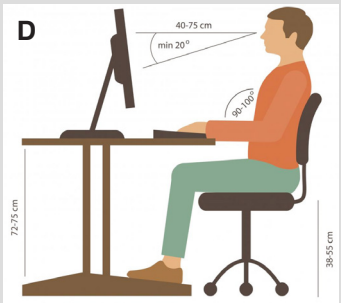
B



C



D



1. Do you think these people in images A,B, C and D are able to read in an effective way? Explain your answer to support your decision with clear evidence.
2. Based on your observation, suggest the causes for the inconveniences that leads to read in different positions?
3. Do you think these people can be able to read in a good strategic position? Suggest possible corrections to the identified defects?

People put on eye glasses for different reasons. Some people wear them in order to read a text, some put them on to see near objects if their eyes cannot be able to do so while others put them on so as to focus distant objects; others wear them for fan like sun goggles

1. Short-sightedness (myopia)

People with normal vision can focus clearly near and distant objects. Those who only focus near objects are said to be short-sighted, meaning that they see nearer.

Short-sightedness is the defect whereby a person can see near objects clearly but cannot focus distant objects. His far point is nearer than infinity. This is because the eyeball is too long or the lens is too strong so that rays of light from a distance object are focused in front of the retina.

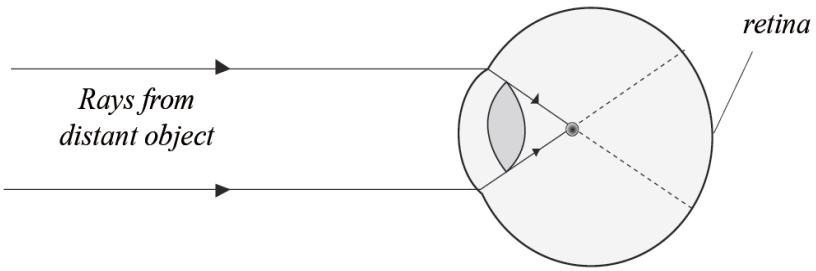


Figure 6.16: Short-sightedness

The rays are focused in front of the retina because the focal length of the eye lens is too short for the length of the eye ball. This defect can be corrected by wearing a concave (diverging) spectacle lens. The rays of light from a distant object are diverged so that they appear to come from a point near, and so they are focused by the eye.

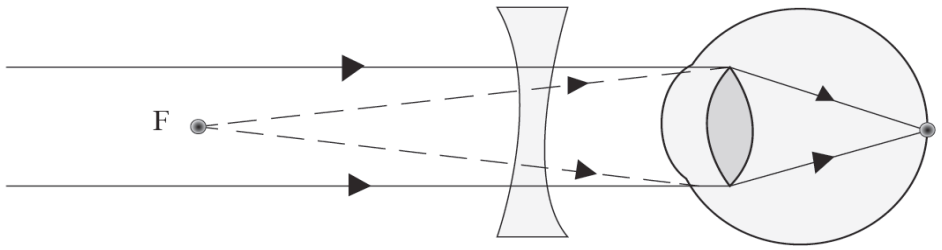


Figure 6.17: Correction of short sight

Rays from object at infinity appear to come from a near point F and converge to the retina.

2. Long-sightedness (hypermetropia)

This is where a person is able to see distant objects clearly but cannot focus near objects.

This is because either his eye ball is too short or the eye lens is too weak (thin) so that rays of light from a close object are focused behind the retina. This eye's near point is further than 25 cm.

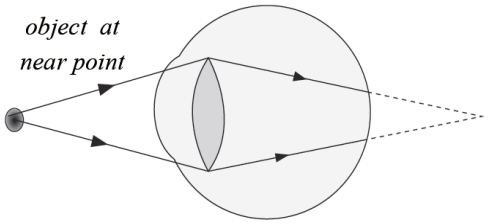


Figure 6.18: Long-sight

The image of the near object is focused behind the retina because the focal length of the eye lens is too long for the length of the eye ball.

This defect can be corrected by wearing a convex lens spectacle. The rays of light from a near object are converged so that the rays appear to come from a point far, and so are focused by the eye.

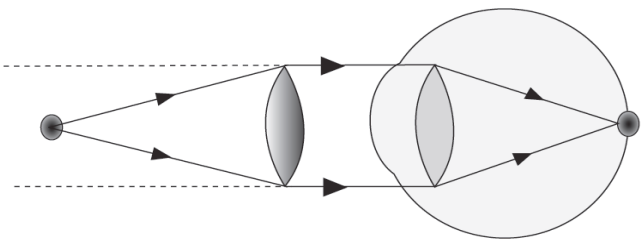


Figure 6.19: Correction of long sight

3. Presbyopia

When people grow older, their eye lens become stiff and it becomes hard for the ciliary muscles to adjust it. Such people have a defect called Presbyopia.

Presbyopia is the stiffening of the eye lens such that it is less capable of being adjusted by the ciliary muscles. This means that the eye lens becomes less flexible and loses its power (ability) to accommodate for objects at different distances.

This defect is corrected by wearing bifocals spectacles whose lenses have a top part for looking at distant objects and a bottom part for close ones. These bifocal spectacles have a diverging top part to correct for distant vision and converging lower part for reading.

4. Astigmatism

This is the defect that occurs if the curvature of the cornea varies in different directions so that rays in different planes from an object are focused in different positions by the eye and the image is distorted. A person suffering from astigmatism sees one set of lines more sharply than others. This defect is corrected by wearing corrected lenses. These help to bend the incoming rays to correct for irregular refraction.

Some of the most common signs of astigmatism are headaches, squinting, blurry vision, eye strain, and difficulty driving at night.

Application activity 6.5

1. A person is unable to see objects distinctly placed within 50cm from his eyes.
   a) Name the defect of vision the person is suffering from and list its two possible causes.
   b) Draw a ray diagram to show the defect in the above case.
   c) Mention the type of lens used by him for the correction of the defect
   d) Draw a labeled diagram for the correction of the defect in the above case.
2. a. What are the causes of astigmatism?
   b. How is astigmatism be corrected?
   c. How do I know if I have astigmatism?
3. a. what causes presbyopia?
   b. What are the common symptoms of presbyopia?
4. Why is a normal eye not able to see clearly the objects placed closer than 25cm?

Skills lab 6

Materials

- Glass microscope slides
- Plastic cover slips
- Paper towels or tissues
- Salt solution (6g salt in 100 ml of water or approx. 2 teaspoons salt in a cup of water)
- Hydrocharitaceae plant
- Water

Methods

1. Pick off an entire healthy looking hydrocharitaceae leaf, with fingers or small scissors and place it on the microscope slide.
2. Add a drop of water (hypotonic solution) and a coverslip and observe the chloroplasts (green structures) and the cell walls.
3. Add a drop of salt solution (hypertonic solution) to the side of the coverslip and observe the cell shrinking (optional).

When the salt solution is added, the salt ions outside the cell membrane cause the water molecules to leave the cell through the cell membrane causing it to shrink into a blob in the centre of the cell wall. The movement of water molecules is called osmosis.

Hydrocharitaceae plant is a species that works well for this activity, because it has thin, straight leaves. However, most other species are satisfactory.

End unit assessment 6

1. With the aid of a ray diagram, describe how a convex lens is used as a magnifying glass.
2. (a) When is a compound microscope said to be in normal use?

   (b) Derive an expression for the magnifying power of a compound microscope in normal use.

   (c) Explain why the lenses that make up a compound microscope are of short focal lengths.
3. (a) When is a telescope said to be in normal adjustment.

   (b) What is meant by the eye ring as applied to optical instruments.

   (c) What are the differences between microscope and telescopes?
4. The objective of an astronomical telescope in a normal adjustment has a diameter of 15cm and a focal length of 400cm. The eye piece has a focal length of 2.5cm. Find the magnifying power of the telescope.
5. How far must a 50mm focal-length camera lens be moved from its infinity setting to sharply focus an object 3m away?
6. a. The far point of the defective eye is 1m. What lens is needed to correct this defect?

   b. With this lens, at what distance from the eye is its near point, if the near point is 25cm without the lens?
7. The near point, N, of a defective eye is 30cm from the eye. If the normal near point is 25 from the eye, find the focal length and power of the lens needed to correct this defect.
8. A person needs a lens of power –5.5 dioptres for correcting his distant vision. For correcting his near vision he needs a lens of power +1.5 dioptre. What is the focal length of the lens required for correcting (i) distant vision, and (ii) near vision?

UNIT 7:

PROPAGATION OF MECHANICAL WAVES

Key Unit competence: Evaluate the propagation of mechanical waves

Introductory Activity 7

Questions:

a). Look at the illustration given above. What do you see in the illustration above?

b). Can you suggest what causes the water in the illustration above to move in such way? Explain to support your idea?

c). Suggest what makes the water to propagate from one place to another?

d). Is it possible to hear the water moving in such way? How do we call what is heard by someone during this movement of water? Explain your answer to support your arguments.

e). Have you ever heard of discussions about Thomas Young? If yes, how do Thomas young talk about the propagation of light waves? Justify your arguments with possible illustrations.

7.1. Introduction to waves

Activity 7.1

- Pull water into a bucket and wait it to be at rest. Drop a small stone into the water. How is the motion of water disturbed? Explain your observation
- Fix one end of a rope on a support (fixed nail or a tree or another support) and hold another end in your hands. Move the rope up and down and observer the motion of the disturbance. Does the rope oscillate? If yes, does it oscillate perpendicular or parallel to the direction of the motion of the disturbance? Suggest the type of wave is this.
- Now fix one end of an elastic spring and hold another end. Move the spring back and forth and observe the behavior of the disturbance created. Does it oscillate perpendicular or parallel to the direction of the motion of the disturbance? Suggest the type of wave is this.

In physics, a wave is a disturbance that occurs in a material medium and in such process, energy is transferred from one place to another. When studying waves, it's important to remember that they transfer energy, not matter.

There are lots of waves all around us in everyday life. Sound is a type of wave that moves through matter and then vibrates our eardrums and we hear. Light is a special kind of wave that is made up of photons that helps us to see. You can drop a rock into a pond and see wave formation in the water. We even use waves (microwaves) to cook our food really fast. Application of this concept is extensively used in telecommunication and music.

7.1.1. The concept of waves

Waves can be defined as a disturbance in a medium that transfers energy from one place to another, although the medium itself does not travel.

The term wave is often intuitively understood as referring to a transport of spatial disturbances that are generally not accompanied by a motion of the medium occupying this space as a whole. In a wave, the energy of a vibration is moving away from the source in the form of a disturbance within the surrounding medium.

Other properties, however, although usually described in terms of origin, may be generalized to all waves. For such reasons, wave theory represents a particular branch of physics that is concerned with the properties of wave processes independently of their physical origin.

7.1.2. Types of waves

There are **three types of waves**

- Transverse waves
- Longitudinal waves
- Surface wave

a). Transverse waves

A *transverse wave* is one in which the direction of the oscillation (vibration) is perpendicular to the travel of the wave

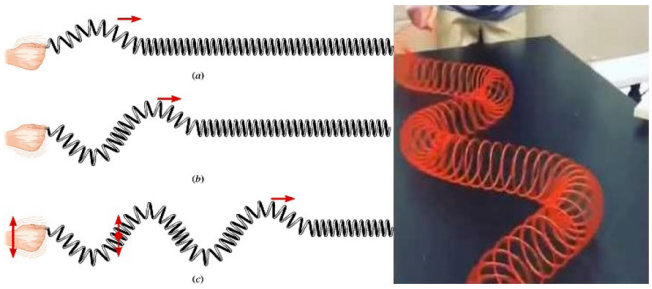


Figure 7.1: Transverse waves

Examples: wave on plucked strings, water wave, electromagnetic wave (light wave, radio and TV waves microwaves)

The particles do not move along with the wave; they simply oscillate up and down about their individual equilibrium positions as the wave passes by.

b). Longitudinal waves

A ***longitudinal wave*** is one in which the direction of the vibration is the same as the direction of travel of the wave.

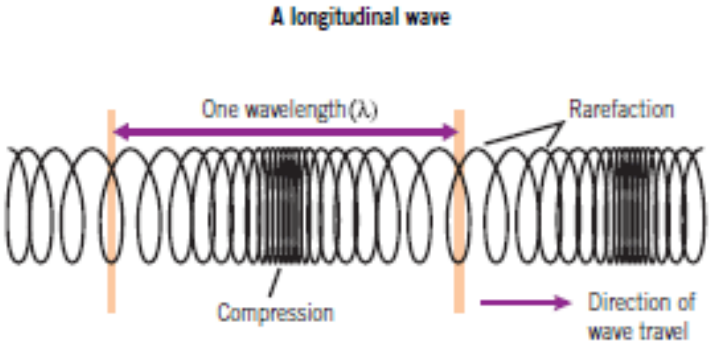


Figure 7.2: Longitudinal wave

Examples of longitudinal waves include sound waves, compression waves in gas and solids.

c). Surface waves

Waves that are deep in a lake or ocean are longitudinal; at the surface of the water, however, the particles move in a direction that is both parallel and perpendicular to the direction of wave motion. Each of the waves is a **surface wave**, which has characteristics of both transverse and longitudinal waves. The energy of water waves usually comes from distant storms, whose energy initially came from the heating of Earth by solar energy. This energy, in turn, was carried to Earth by transverse electromagnetic waves from the Sun.

7.1.3. Characteristics of waves

The wave is characterized by the *speed*, *amplitude*, *wave length, frequency*, *period* , *phase* and *wave fronts*.

A wave allows the energy to be transferred from one point to another some distance away without any particle of the medium travelling between the points.

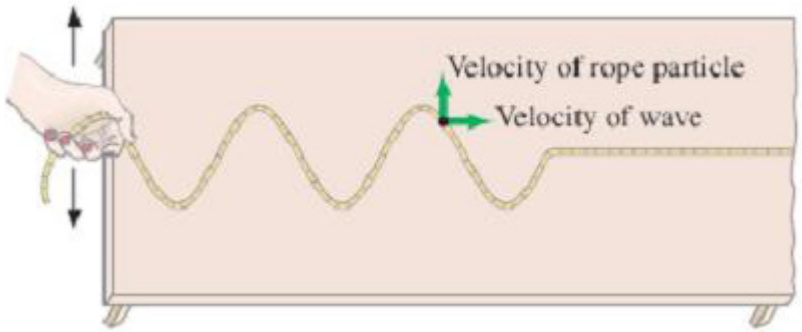


Figure 7.3: wave through a rope

In the figure 7.3, the wave travels to the right along the cord but particles of the cord oscillates back and forth then produce a continuous or periodic wave because it has its source a disturbance that is continuous and oscillating. Some of the important quantities used to describe a periodic sinusoidal wave are shown in the figure below

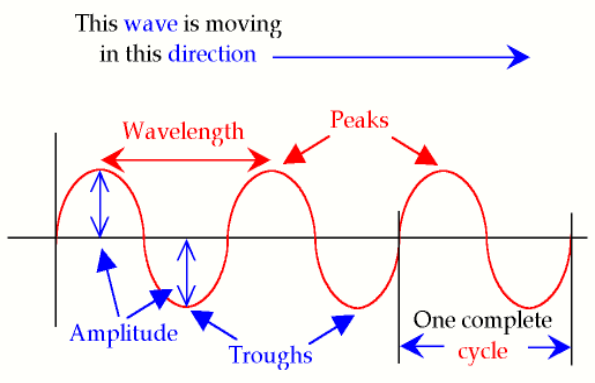


Figure 7.4: Periodic sinusoidal wave

- The high point on a wave are called **crest**s or **peaks**
- The low point on a wave are called **troughs**

The **amplitude** is the maximum height of a crest or the depth of a trough relative to the normal (or the equilibrium) level

The **wavelength** is the distance between two successive crests the wave length is also equal to the distance between any two successive identical points on the wave

The **frequency** is the number of the crests or complete cycles that pass a given point per unit time

The **period** T is the time elapsed between two successive crests or trough passing by the same point in the space

A period is also defined as the time taken by a wave to travel one wave length or to undergo one complete oscillation

The **wave velocity** v is the distance moved by any point on the wave in one second since distance moved in periodic time T is λ then

$$v = \frac{\lambda}{T} = \lambda f, \text{since} \quad f = \frac{1}{T}$$

That is, wave speed = wavelength × frequency. This is the relationship between wavelength, frequency and velocity.

Phase difference (phase angle) is the angular difference between two points on the wave or between two waves. Phase difference is a whole number.

The *wave number*, also called the propagation number k, is the spatial frequency of a wave, either in cycles per unit distance or radians per unit distance. It can be envisaged as the number of waves that exist over a

specified distance (analogous to frequency being the number of cycles or radians per unit time). Its unit is per metre (m^{-1}). Mathematically;

$$k = \frac{2\pi}{\lambda}$$

Wavefrontis a line or surface in the path of the wave motion on which the disturbance at every point have the same phase. This can also be defined as the surface which touches all the wavelets from the secondary sources of waves. Consider the Huygens construction principle for the new wavefront.

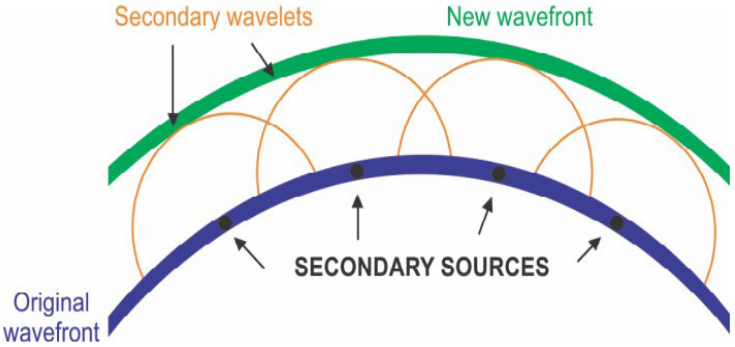


Fig. 7.5: Formation of the wave-front

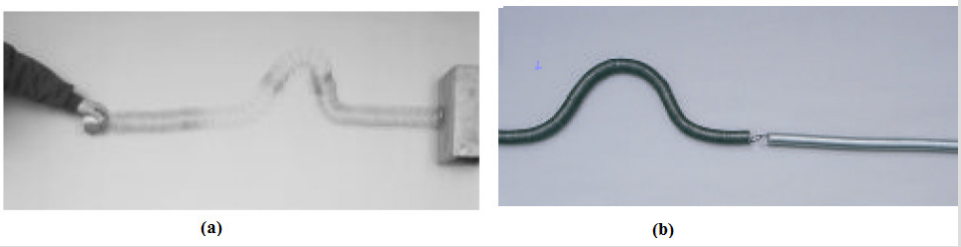
Phase: Any two points on a wave that are one or more whole wavelengths apart are in phase. Particles in the medium are said to be in phase with one another when they have the same displacement from equilibrium and the same velocity. Particles in the medium with opposite displacements and velocities are 180° out of phase. A crest and a trough, for example, are 180° out of phase with each other. Two particles in a wave can be anywhere from 0° to 180° out of phase with one another.

Application activity 7.1

1. In a long line of people waiting to buy tickets, the first person leaves and a pulse of motion occurs as people step forward to fill the gap. As each person steps forward, the gap moves through the line. Is the propagation of this gap (a) transverse (b) longitudinal?
2. Consider the “wave” at a baseball game: people stand up and shout as the wave arrives at their location, and the resultant pulse moves around the stadium. Is this wave (a) transverse (b) longitudinal?
3. Use a wave diagram to represent the crests and troughs of a wave. Then describe a crest and trough in your own words.

7.2. Properties of waves: reflection, refraction, diffraction

Activity 7.2



(a) (b)

Task 1

The figure above shows springs. (a) a pulse(disturbance) approaches a rigid wall; (b) springs with different media connected end to end. Observe the figure and answer the following questions

a). When the pulse reaches at the rigid wall (figure (a)), how does it behave? What is this phenomenon called?

b). In figure (b), what happens when the pulse reaches the boundary of the springs? Explain your observation. What is this phenomenon called?

Task 2

Pull water in a platic wash basin and wait until rest. Drop a stone in thewater so that moves in circular path. Put your two fingers with small separation in disturbed water and observe the behavior of the water disturbance throughout your fingers. Explain your observation. Suggest the name of this phenomenon.

Now drop two stones in water simultaneously. What happens when the disturbances(waves) meet or overlap? What is this phenomenon called?

7.2.1. Reflection of wave

1. When the pulse propagates along the string at the end of the string the pulse undergoes reflection i.e its velocity changes the direction (a wave pulse is a temporary disturbance formed in medium by quick and down motion of particles of the medium)
   - If the two ends of the string are free the reflected pulse has the same form as the incident pulse and the displacement does not change its sign.

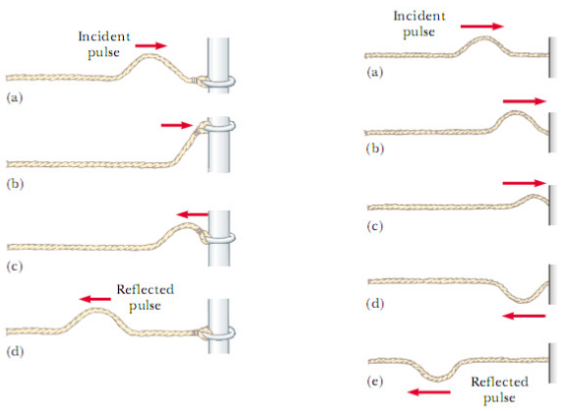


Figure 7.6: Reflection of wave at fixed boundary

- If one end is fixed the reflected pulse has the same form as the incident pulse but with change of sign of the displacement

2. When a plane waves (eg water waves in ripple tank) meet a plane barrier at an angle Q both incident and reflected wave front are straight and have equal spacing.

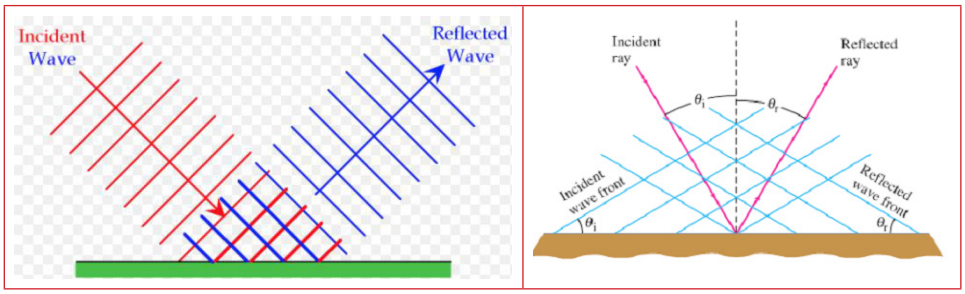


Figure 7.7: reflection of plane wave on plan barrier

When plane wave meet a concave barrier the reflected wave front are circular converging towards and passing thoughts the focal point F.

3. When plane wave meet a concave barrier the reflected wave front are circular converging towards and passing thoughts the focal point F.

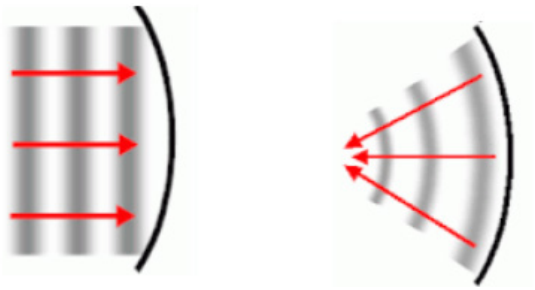


Figure 7.8: Reflection of plane wave on a concave surface

The incident are reflected waves have the same speed and wavelength

4. When circular waves meet a plane barrierthe reflected waves are also circular wave and have the samespacing and frequency.

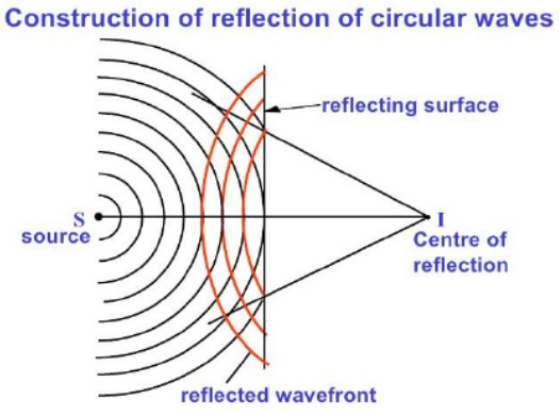


Figure 7.9: reflection of circular wave on a plane barrier.

7.2.2. Refraction of waves

When any wave strikes a boundary some of the energy is reflected and some is transmitted or absorbed. When a two or three dimensional wave travelling in one medium crosses a boundary into a medium where its speed is different the transmitted way may move in different direction than the incident wave this phenomena is known as refraction.

Consider a parallel beam of waves (for example light waves) incident on a refracting surface between two media such that its direction of travel makes angle θ_1 with the normal to the refracting surface.

Consider side A of the wave-front AB has reached the surface before B. If the ray from the other side B of the beam consequently travels to C at time t,

$BC = C_1 t$

Assuming: $C_1 \rightarrow$ speed of light in medium 1

$C_2 \rightarrow$ speed of light in medium 2

$t \rightarrow$ time taken for the ray to move from B to C

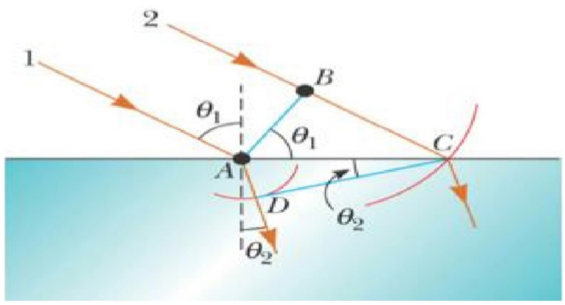


Figure 7.10: Refraction of waves

At the same time, wavelets from A travel distance AD in medium 2. Here, a refracted wavefront CD is formed by many wavelets in the beam. Fig. 5.6 above illustrates this description.

$$AD = C_2 t$$

Considering triangle ABC and ADC

$$\sin\theta_1 = \frac{BC}{AC} = \frac{C_1 t}{AC}$$

$$\sin\theta_2 = \frac{AD}{AC} = \frac{C_2 t}{AC}$$

Dividing these Equations gives:

$$\frac{\sin\theta_1}{\sin\theta_2} = \frac{C_1}{C_2}$$

By Snell's law:

$$\frac{\sin\theta_1}{\sin\theta_2} = \frac{C_1}{C_2} = Constant.$$

This Equation confirms Snell's law meaning that waves behave like normal light during reflection.

7.2.3. Diffraction of waves

This is a phenomenon in which waves from one source meet an obstacle and spread around it. Diffraction is normally observed when these waves pass through narrow slits. There are two types of diffraction: **Fresnel's**

diffraction and Fraunhofer diffraction.

7.2.3.1. Fresnel's diffraction

This is a type of diffraction in which either the source of waves or screen on which diffraction is observed or both are at finite distances from the obstacle that cause diffraction. Below are different cases to explain this diffraction.

Case 1: The source and the screen placed at finite distances.

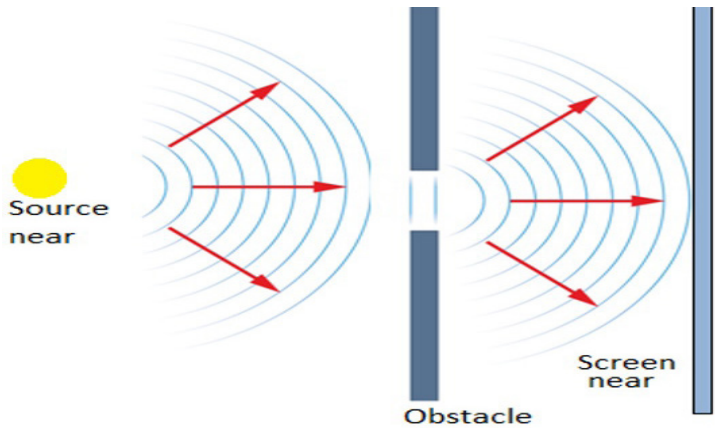


Figure 7.11 : Fresnel's diffraction case 1

Case 2: The source is placed at infinite distance from obstacle and the screen is near.

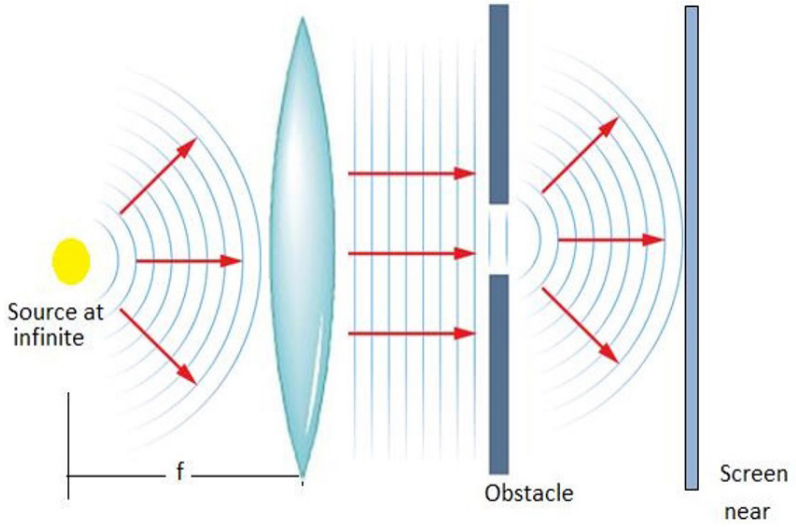


Figure 7.12: Fresnel's diffraction case 2

Case 3: The screen is placed at infinite distance from obstacle and the source is near.

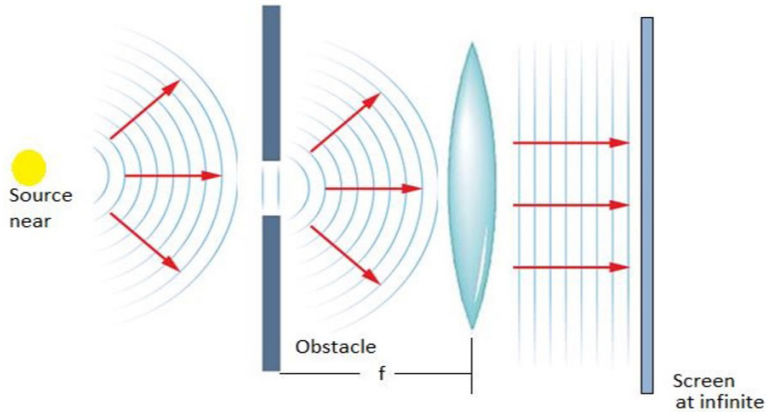


Figure 7.13: Fresnel's diffraction case 3

7.2.3.2. Fraunhofer Diffraction

This is a type of diffraction in which the source of waves and the screen on which diffraction is observed are effectively at infinite distances from the obstacle. This phenomenon is practically complicated but theoretically understood. To obtain waves to or from infinite source in laboratory, biconvex lenses are used.

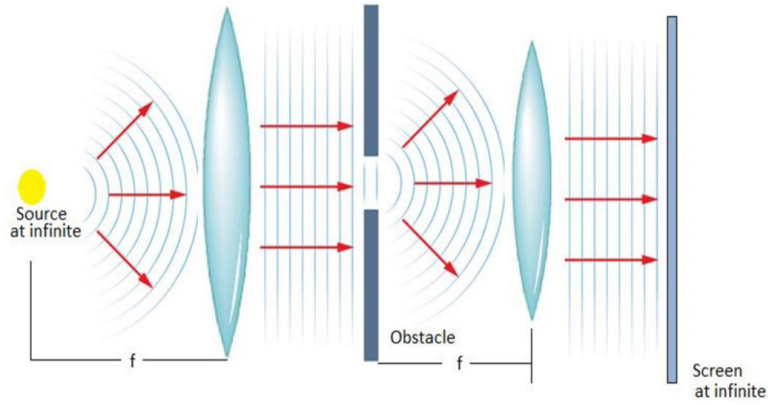


Figure 7.14: Fraunhofer diffraction

7.2.4. Interference of waves

Two or more waves traveling in the same medium travel independently and can pass through each other. In regions where they overlap, we only observe a single disturbance. We observe interference. When two or more waves interfere, the resulting displacement is equal to the vector sum of the individual displacements. If two waves with equal amplitudes overlap in phase, i.e. if crest meets crest and trough meets trough, then we observe a resultant wave with twice the amplitude. We have constructive interference. If the two overlapping waves, however, are completely out of phase, i.e. if crest meets trough, then the two waves cancel

each other out completely. We have **destructive interference**. The resulting effect is called interference pattern or the system of **fringes.**

Condition for interference is that the waves to interfere must be of the coherent sources

Coherent sources are sources of the same frequency (or wavelength) and the same amplitude and these sources have a constant phase difference.

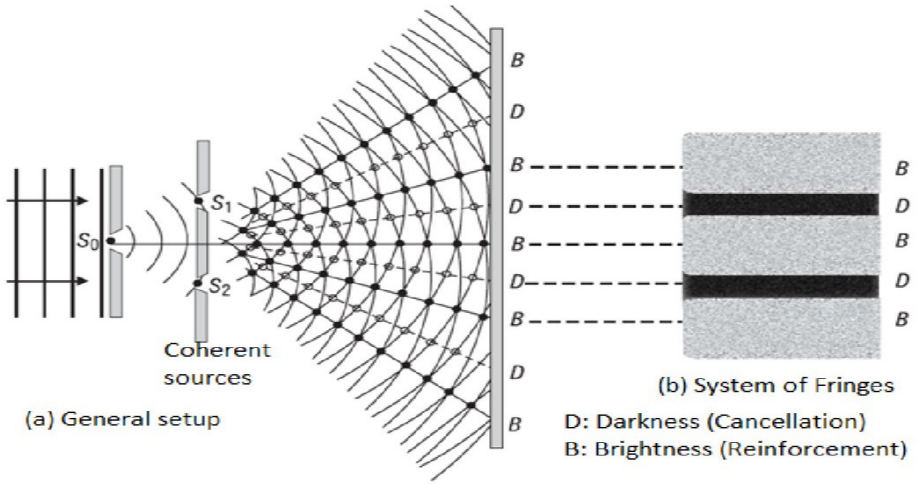


Figure 7.15: interference of waves from coherent sources.

Application activity 7.2

1. When a wave reaches the boundary of a new medium, what happens to it?
2. When a wave crosses a boundary between a thin and a thick rope, as shown in Figure below, its wavelength and speed change, but its frequency does not. Explain why the frequency is constant.

3. How does a spring pulse reflected from a rigid wall differ from the incident pulse?
4. Describe interference. Is interference a property of only some types of waves or all types of waves?

7.3. Progressive waves and its mathematical treatment

Activity 7.3

Conduct an internet research on progressive waves, its wave function and where they are applied in everyday life and report.

7.3.1.Progressive waves

A progressive wave is also called a travelling wave which consists of a disturbance moving from one point to another. As a result, energy is transferred between points.

Progressive mechanical waves can be categorized according to the direction of the effect of the disturbance relative to the direction of travel. Progressive waves are classified as longitudinal and transverse waves already discussed above.

7.3.2. Equation of a progressive wave

An equation can be performed to represent displacement y of a vibrating particle in a medium in which a wave passes. Suppose a wave moves from left to right and that a particle at the origin moves with displacement given by equation.

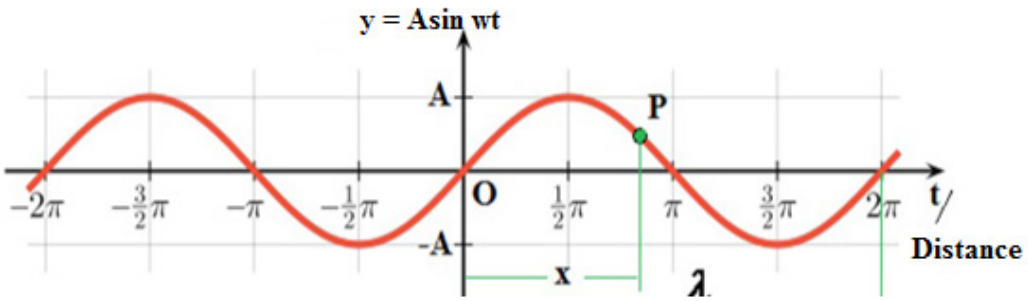


Figure 7.16: Curve of Transverse wave

A particle at P will be out of phase from the particle at O, so, its displacement is given by; $y_p = A\sin(\omega t - \Phi)$

But the phase angle $\Phi = \dfrac{2\pi x}{\lambda}$

$$y_p = A\sin\left(\omega t - \frac{2\pi x}{\lambda}\right)$$

Various forms of progressive wave function are listed below:

$$y_p = A\sin 2\pi\left(ft - \frac{x}{\lambda}\right)$$

$$y_p = A\sin\frac{2\pi}{\lambda}(vt - x)$$

$$y_p = A\sin\omega\left(t - \frac{x}{v}\right)$$

$$y_p = A\sin\left(\omega t - \frac{2\pi x}{\lambda}\right) = A\sin(\omega t - kx)$$

This is the same as $y_p = A\sin(kx - \omega t)$

Where by $y = displacement$, $A = amplitude$, $\omega = angular\ frequency$, $f = frequency$, $k = wave\ number$, $T = time\ period$, $\lambda = wave\ length$, $v = wave\ velocity$ $t = instantaneous\ time$, and $x = position\ of\ particle\ from\ origin$.

Note: These Equations represent plane progressive waves. The negative sign indicates that the wave is moving from left to right and all points on the right of O will lag behind O.

A wave travelling in opposite direction; from right to left arrives at P before O. So, vibrations at P will read that at O and then displacement is given by;

$$y_p = A\sin 2\pi\left(ft + \frac{x}{\lambda}\right)$$

$$y_p = A\sin\frac{2\pi}{\lambda}(vt + x)$$

$$y_p = A\sin\omega\left(t + \frac{x}{v}\right)$$

$$y_p = A\sin\left(\omega t + \frac{2\pi x}{\lambda}\right) = A\sin(\omega t + kx)$$

This is the same as

$$y_p = A\sin(kx + \omega t)$$

In general, the equation of progressive wave is

$$y_p = A\sin(kx \mp \omega t)$$

The equation of progressive wave can also be expressed as

$$y_p = A\cos\left(kx \mp \omega t\right)$$

Example

A certain transverse wave is described by

$$y(x,t) = (6.50mm)\cos 2\pi\left(\frac{x}{28.0\,cm} - \frac{t}{0.0360\,s}\right)$$

Determine the wave's (a) amplitude; (b) wavelength; (c) frequency; (d) speed of propagation; (e) direction of propagation.

Solution

a). Amplitude $A = 6.50mm$

b). Wavelength $\lambda = 28.0\,cm$

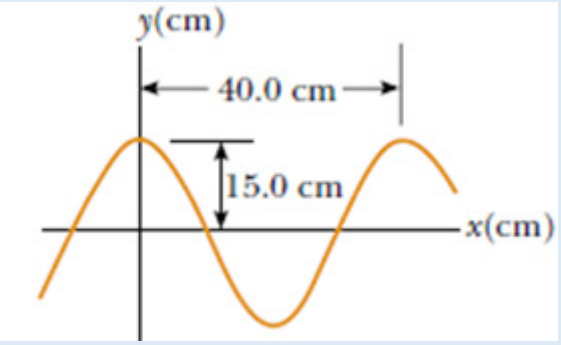
c). Frequency $f = \frac{1}{T} = \frac{1}{0.0360s} = 27.8Hz$

d). Speed of propagation $v = \lambda f = (0.28m)(27.8Hz) = 7.78\,m/s$

e). Since there is a minus sign in front of the t/T term, the wave is traveling in the + x-direction.

Application activity 7.3

A sinusoidal wave traveling in the positive x direction has amplitude of 15.0 cm, a wavelength of 40.0 cm, and a frequency of 8.00 Hz. The vertical position of an element of the medium at t = 0 and x = 0 is also 15.0 cm as shown in fig…...



a). Find the wave number k, period T, angular frequency ', and speed v of the wave.

b). Determine the phase constant Φ, and write a general expression for the wave function

7.4. Young's double slit experiment and its mathematical treatment

Activity 7.4

Experiment about using Fingers as Slits

Look at a light, such as a street lamp or incandescent bulb, through the narrow gap between two fingers held close together. What type of pattern do you see? How does it change when you allow the fingers to move a little farther apart? Is it more distinct for a monochromatic source, such as the yellow light from a sodium vapor lamp, than for an incandescent bulb?

In Young's double-slit experiment, when monochromatic light passing through two narrow slits illuminates a distant screen, a characteristic pattern of bright and dark fringes is observed. This interference pattern is caused by the superposition of overlapping light waves originating from the two slits.

In this experiment, a source S of monochromatic waves is used to illuminate two narrow and parallel slits $S_1$and S_2that are apparent sources of light. The arrangement is shown in Fig.7.2.

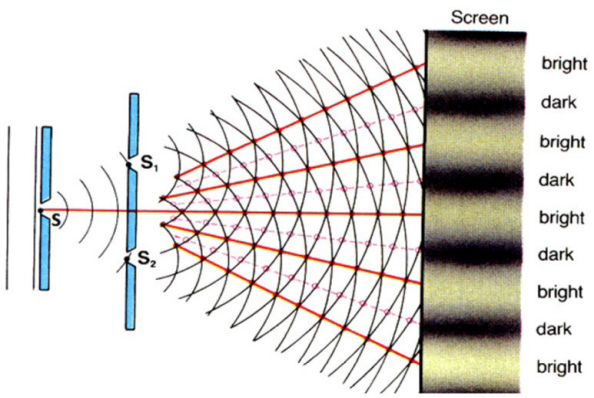
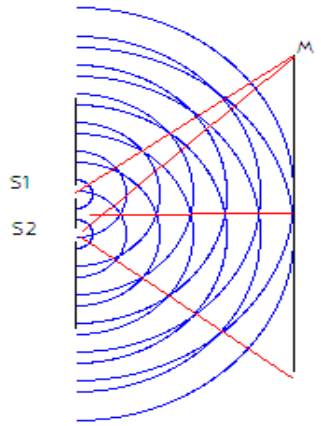


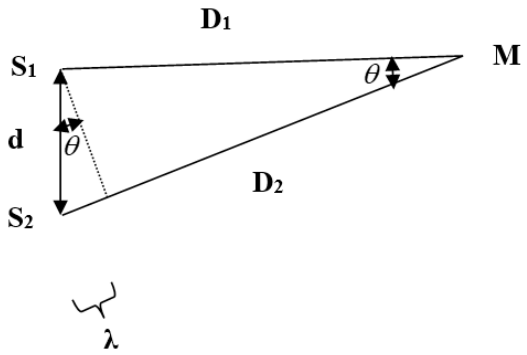
Fig. 7.17: Young's double slit experiment

From the above experiment, a narrow slit S diffracts light falling on it and so illuminates $S_1$and S_2. Diffraction also takes place at $S_1$and S_2 and interference takes place in the region where light from S_1 overlaps the light from S_2.

Light from $S_1$and S_2 is from a monochromatic source; hence $S_1$and S_2 act as coherent sources (they are also monochromatic).

Let us emphasize on the point M which is encountered by the waves from the source S_1 and the source S_2. The scheme is considered as below:





$$\sin\theta = \frac{\lambda}{d} = \frac{\lambda}{s_1 s_2}$$

D_1 is the distance between the source S_1 and the point M, and D_2 the distance between the source S_2 and the point M. d is the distance separating two sources.

The path difference is then $D_2 - D_1 = d\sin\theta$

Conditions for interference

Constructive interference occurs at points where the path difference is an integral number of wavelength, $n\lambda$, where $n = 0, \pm1, \pm2, \pm3, \ldots$

So the bright regions (fringes) on the screen occur at angles θ for which

$$d\sin\theta = n\lambda \quad (n = 0, \pm1, \pm2, \pm3, \ldots)$$

Similarly, destructive interference occurs, forming dark regions on the screen, at point for which the path difference is the half integral number of the wavelengths $\left(n + \frac{1}{2}\right)\lambda$

$$d\sin\theta = \left(n + \frac{1}{2}\right)\lambda, \quad (n = 0, \pm1, \pm2, \pm3, \ldots)$$

Thus the pattern on the screen of figure above is a succession of bright and dark bands, or interference fringes, parallel to the slits S_1 and S_2.the center of the pattern is a bright band corresponding to $n = 0$; this point on the screen is equidistant from the two slits.

We can derive an expression for the positions of the centers of the bright bands on the screen. Let y_n be the distance from the center of the pattern ($\theta = 0$) to the center of the n^{th} bright band. Let θ_n be the corresponding value of θ; then

$$y_n = R \tan \theta_n$$

In experiments such as this, the distance y_n are often much smaller than the distance R from the slits to the screen. Hence θ_n is very small, $\tan \theta_n$ is very nearly equal to $\sin \theta_n$, and

$$y_n = R \sin \theta_n$$

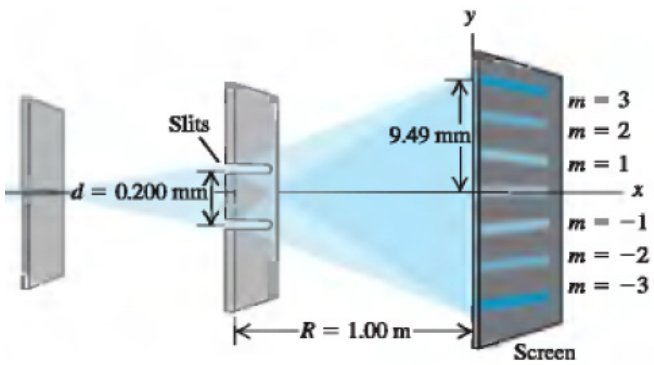
We find that for small angles only

$$y_n = R \frac{n\lambda}{d}$$

We can measure R and d, as well as the positions y_n for the bright fringes, so this experiment provides a direct measurement of the wavelength λ. Young's experiment was in fact the first direct measurement of the wavelengths of light.

Example

In a two-slit interference experiment, the slits are 0.200 mm apart, and the screen is at a distance of 1.00 m. the third bright fringe (not counting the central bright fringe straaight ahead from the slits) is found to be displaced 9.49 mm from the central fringe (figure). Find the wavelength of the light used.



Solution

This problem aks to determine the wavelength from the dimensions $d=0.200mm$ (slit separation), $R = 1.00m$ (distance from slits to screen), and $y_n = 9.49mm$ (distance of the third bright fringe from the center of the pattern).

$$y_n = R\frac{n\lambda}{d} \rightarrow \lambda = \frac{y_n d}{nR}$$

$$\lambda = \frac{\left(9.49\times10^{-3}m\right)\left(0.200\times10^{-3}m\right)}{3\left(1.00m\right)} = 633\times10^{-9}\,m = 633\,nm$$

Application activity 7.4

1. Young's double slit experiment breaks a single light beam into two sources. Would the same pattern be obtained for two independent sources of light, such as the headlights of a distant car? Explain your answer with possible sketches.
2. Suppose you use the same double slit to perform Young's double slit experiment in air and then repeat the experiment in water. Do the angles to the same parts of the interference pattern get larger or smaller? Does the color of the light change? Explain to justify your arguments.
3. Is it possible to create a situation in which there is only destructive interference? Explain.

7.5. Stationary waves

Activity 7.5

- Attach one end of a rope or coiled spring to a fixed point, such as a doorknob, and then start to vibrate the other end, what happens to the wave leaves your hand when it reaches at the fixed end? How is the reflected wave?
- Suppose you adjust the motion of your hand so that the period of vibration equals the time needed for the wave to make one round-trip from your hand to the door and back. What happens to the displacement given by your hand to the rope and displacement of the reflected wave?.
- Compare the oscillation of one segment and the motion of your hand .

- Do you observe points on the rope that do not move and others that move at maximum height? Suggest their names. How do you call such kind of wave?

7.5.1. Meaning of stationary waves

A stationary wave (or a standing wave) is a wave which results when two waves travelling in opposite directions and having the same speed; frequency and approximately equal amplitudes are superposed. A standing wave is shown in Fig. 7.18.

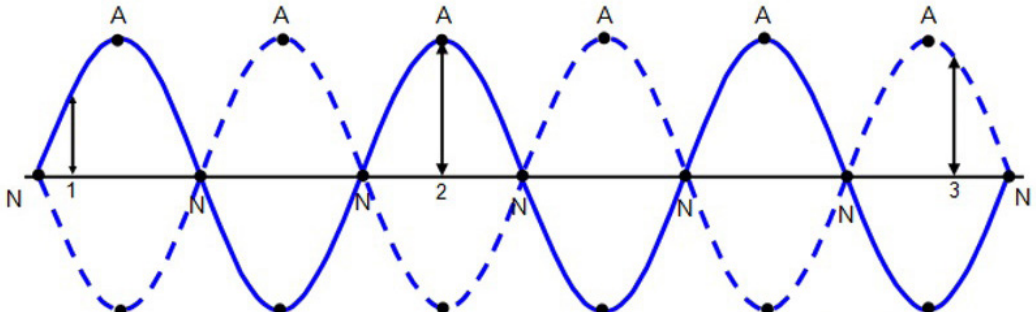


Figure 7.18: Nodes and antinodes of a standing wave

7.5.2. Meaning of nodes and antinodes

Here, A denotes the antinodes and N denotes the nodes. One characteristic of every standing wave pattern is that there are points along the medium that appear to be standing still. These points, sometimes described as points of no displacement, are referred to as **nodes**. There are other points along the medium that undergo vibrations between a large positive and large negative displacement.

These are the points that undergo the maximum displacement during each vibrational cycle of the standing wave. In a sense, these points are the opposite of nodes, and so they are called **antinodes**.

A standing wave pattern always consists of an alternating pattern of nodes and antinodes. The nodes and antinodes are labeled on the diagram.

When a standing wave pattern is established in a medium, the nodes and the antinodes are always located at the same position along the medium; they are *standing still*. It is this characteristic that has earned the pattern the name *standing wave*.

7.5.3. Mathematical treatment of standing waves

Consider the displacement $y_1 = a\sin(\omega t - \Phi)$ of a progressive sinusoidal wave at time t and at a distance x from the origin and moving to right.

Consider also the displacement $y_2$of an identical wave travelling in opposite direction given by

$$y_2 = a\sin(\omega t + \Phi)$$

If these waves are superposed, the resultant displacement y is given by;

$$y = y_1 + y_2 = a\sin(\omega t - \Phi) + a\sin(\omega t + \Phi)$$
$$y = 2a\cos\Phi\sin\omega t = 2a\,\text{coskx}\sin\omega t, \text{sin}\,ce\ \Phi = \text{kx}$$

The only variable part of equation is . This means that the amplitude of the resultant displacement is given by equation

$$A = 2a\,\text{coskx}$$

Case1: Position of nodes

A node is defined as the point of zero amplitude. This means

$$A = 2a\,\text{coskx} = 0$$
$$\text{coskx} = 0 \rightarrow kx = \cos^{-1} 0$$
$$kx = \frac{\pi}{2}, \frac{3\pi}{2}, \frac{5\pi}{2}, \frac{7\pi}{2}, ...$$

But $k = \frac{2\pi}{\lambda}$

Therefore

$$x = \frac{\lambda}{4}, 3\frac{\lambda}{4}, 5\frac{\lambda}{4}, 7\frac{\lambda}{4}, ...n\frac{\lambda}{4}$$

Where $\text{n} = 1,3,5,7,9, \ldots$

Where x is the horizontal displacement of waves.

Nodes are obtained when the horizontal displacement of waves are odd quarter values of wavelength.

Case 2: Position of antinodes

Antinodes are points of maximum displacements. So, antinodes are obtained when the amplitude is maximum. This occurs when;

$$\text{coskx} = 1 \rightarrow kx = \cos^{-1} 1$$
$$kx = 0, \pi, 2\pi, 3\pi, 4\pi, 5\pi....$$

Thus, the possible position of antinodes are:

$$x = 0, \frac{\lambda}{2}, 2\frac{\lambda}{2}, 3\frac{\lambda}{2}, 4\frac{\lambda}{2}, 5\frac{\lambda}{2}, 6\frac{\lambda}{2}, ...$$

That is $x = n\frac{\lambda}{2}$

where n = 0, 1, 2, 3, 4, 5, 6, ...

This means that antinodes are obtained when horizontal displacement x is a multiple of $\frac{\lambda}{2}$.

Note: The separation of adjacent nodes and adjacent antinodes are $\frac{\lambda}{2}$.

Separation of adjacent nodes = Separation of adjacent antinodes

Therefore, Separation of adjacent nodes = $\frac{\lambda}{2}$

7.5.4. Harmonics and overtone in vibrating string

When a string is fixed at one end and the other end is moved up and down, a transverse wave is formed. The simplest mode of vibration is the one in which both ends are nodes. Let us use as the wavelength, l as the length of the string, vas the speed of the wave and f as the frequency of the wave.

Fig. 7.19: Fundamental mode of vibrating string

$$l = \frac{\lambda}{2} \rightarrow \lambda = 2l$$

$$f_1 = \frac{v}{\lambda} = \frac{v}{2l}$$

The figure above shows the simplest mode of vibration which is the first harmonic (fundamental mode) and its frequency is called the fundamental frequency. Higher order frequencies are called overtones. The second harmonic is the first overtone and has the mode shown by Fig.7.18.

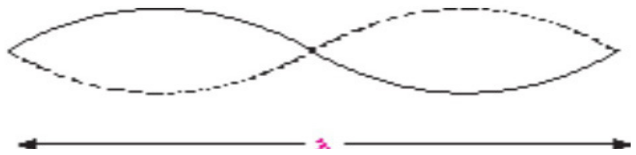


Fig.7.20: First overtone of a vibrating string

It is seen that

$$l = \lambda$$

$$f_2 = \frac{v}{l} = 2f_1$$

The third harmonic, which is the second overtone, is shown in Fig.7.19.

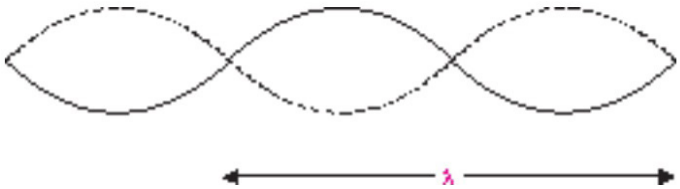


Fig.7.19: Second overtone

If l is the length of string used, then $l = 3\lambda/2 \rightarrow \lambda = 2l/3$

Therefore,

$$f_3 = \frac{3v}{2l} = 3f_1$$

From the above derived expressions of harmonic frequencies, we note that generally;

$$f_n = \frac{3v}{2l} = 3f_1$$

where n is the number of harmonics.

The frequency n^{th} harmonic is given by $f_n = v/\lambda_n$ where is the velocity of progressive wave that produces the standing wave. If l is the length of the string, T is the tension in the string and m is the mass on the scale pan, then the velocity is given by:

$$v = \sqrt{\frac{T}{\mu}}$$

where μ is the mass per unit length, i.e.μ = m/l

Therefore, $v = \sqrt{\frac{Tl}{m}}$

Thus $f_n = \left(\frac{n}{2l}\right)\sqrt{\frac{Tl}{m}}$

Example

A guitar string 30 cm long resonates in its fundamental to a frequency of 256 Hz. What is the tension in the string if 80 cm of the string have a mass of 0.75g? What are the first and second overtones of this string?

Solution

First we shall find the speed and then the tension. We know that the string vibrates in one segment when f = 256 Hz. Therefore,

$$l = \frac{\lambda}{2} \rightarrow \lambda = 2l = (0.30m)(2) = 0.60\,m$$
$$v = \lambda f = (0.60m)(256Hz) = 154\,m/s$$

The mass per unit length of the string is

$$= m/l = \frac{0.75\times10^{-3}kg}{0.8m} = 9.4\times10^{-4}kg/m$$
$$v = \sqrt{\frac{Tl}{m}} \rightarrow T = v^2\mu = (154\,m/s)^2\left(9.4\times10^{-4}kg/m\right)$$
$$T = 22\,N$$

First overtone(2nd harmonics):

$$f_2 = 2f_1 = 2(256Hz) = 512Hz$$

Second overtone (3rd harmonics):

$$f_2 = 3f_1 = 3(256Hz) = 768Hz$$

Application activity 7.5

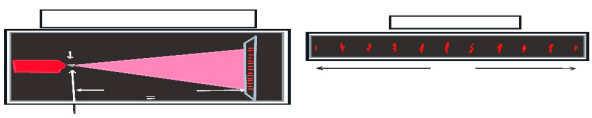
1. Why is a pulse on a string considered to be transverse?
2. How would you create a longitudinal wave in a stretched spring? Would it be possible to create a transverse wave in a spring?
3. By what factor would you have to multiply the tension in a stretched string in order to double the wave speed?
4. When traveling on a taut string, does a pulse always invert upon reflection? Explain.
5. Does the vertical speed of a segment of a horizontal taut string, through which a wave is traveling, depend on the wave speed?
6. If you shake one end of a taut rope steadily three times each second, what would be the period of the sinusoidal wave set up in the rope?

Skills lab 7

In the laboratory, discuss how you can perform Fresnel's diffraction and Fraunhofer diffraction. Draw a report of your discussion.

End unit assessment 7

1. The diagram below depicts the results of Young's Experiment. The appropriate measurements are listed on the diagram. Use these measurements to determine the wavelength of light in nanometers. (GIVEN: 1 meter = 10^9 nanometers)



2. A student uses a laser and a double-slit apparatus to project a two-point source light interference pattern onto a whiteboard located 5.87 meters away. The distance measured between the central bright band and the fourth bright band is 8.21 cm. The slits are separated by a distance of 0.150 mm. What would be the measured wavelength of light?

3. The analysis of any two-point source interference pattern and a successful determination of wavelength demands an ability to sort through the measured information and equating the values with the symbols in Young's equation. Apply your understanding by interpreting the following statements and identifying the values of y, d, m and L. Finally, perform some conversions of the given information such that all information share the same unit.

   a) Two slits separated by 0.250 mm produces an interference pattern in which the fifth dark band is located 12.8 cm from the central antinode when the screen is placed a distance of 8.2 meters away.

      y = ____ d = ____ m = ____ L = ____

      This question simply asks to equate the stated information with the variables of Young's equation and to perform conversions such that all information is in the same unit.

   b) An interference pattern is produced when light is incident upon two slits that are 50.0 micrometers apart. The perpendicular distance from the midpoint between the slits to the screen is 7.65 m. The distance between the two third-order antinodes on opposite sides of the pattern is 32.9 cm.

      y = ____ d = ____ m = ____ L = ____

   c) The fourth nodal line on an interference pattern is 8.4 cm from the first antinodal line when the screen is placed 235 cm from the slits. The slits are separated by 0.25 mm.

      y = ____ d = ____ m = ____ L = ____

   d) Two sources separated by 0.500 mm produce an interference pattern 525 cm away. The fifth and the second antinodal line on the same side of the pattern are separated by 98 mm.

      y = ____ d = ____ m = ____ L = ____

e) Two slits that are 0.200 mm apart produce an interference pattern on a screen such that the central maximum and the 10th bright band are distanced by an amount equal to one-tenth the distance from the slits to the screen.

y = ____ d = ____ m = ____ L = ____

f) The fifth antinodal line and the second nodal line on the opposite side of an interference pattern are separated by a distance of 32.1 cm when the slits are 6.5 m from the screen. The slits are separated by 25.0 micrometers.

y = ____ d = ____ m = ____ L = ____

g) If two slits 0.100 mm apart are separated from a screen by a distance of 300 mm, then the first-order minimum will be 1 cm from the central maximum.

y = ____ d = ____ m = ____ L = ____

h) Consecutive bright bands on an interference pattern are 3.5 cm apart when the slide containing the slits is 10.0 m from the screen. The slit separation distance is 0.050 mm.

y = ____ d = ____ m = ____ L = ____

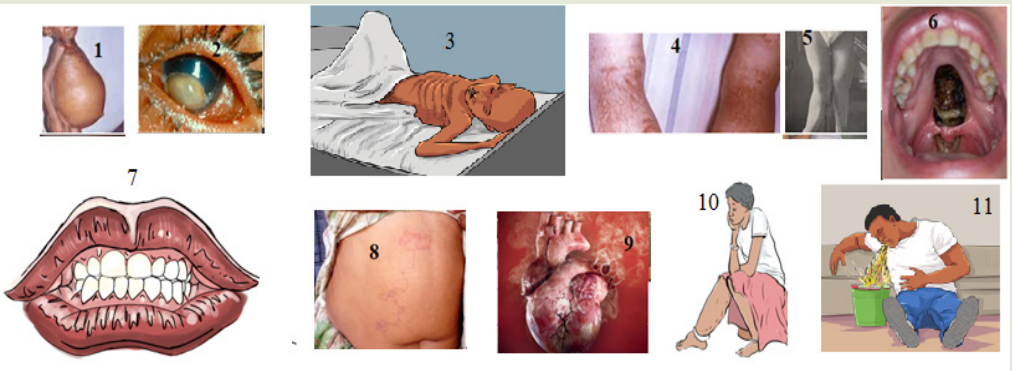
4. A vibrating source generates a sinusoidal wave on a string under constant tension. If the power delivered to the string is doubled, by what factor does the amplitude change? Does the wave speed change under these circumstances?
5. 8. Consider a wave traveling on a taut rope. What is the difference, if any, between the speed of the wave and the speed of a small segment of the rope?
6. If a long rope is hung from a ceiling and waves are sent up the rope from its lower end, they do not ascend with constant speed. Explain.
7. When a particular wire is vibrating with a frequency of 4.00 Hz, a transverse wave of wavelength 60.0 cm is produced. Determine the speed of waves along the wire.
8. A wave is described by y = (2.00 cm) sin(kx - ωt), where k = 2.11 rad/m, ω = 3.62 rad/s, x is in meters, and t is in seconds. Determine the amplitude, wavelength, frequency, and speed of the wave.

UNIT 8:

COMMON DISEASES AND HYGIENE

Key Unit competence: Implement ways of preventing and controlling common diseases and hygiene related issues

Introductory Activity 8



The figures above indicate different situations

a). According to what you always observe in community you live in, what desease observed on each image (No. 1-11)?

b). What do you think is the cause of each case on the figure.

c). What can you do in community for limiting those situations.

8.1. Common diseases

Activity 8.1

1. The government of Rwanda encourages people to visit hospitals and test their blood. A person has done that test and found being HIV positive.

   a). Which disease is that person suffering from?

   b). What is its causative agent?

   c). How is it transmitted from one people to another?

2. Basing on the ways of transmission, suggest other diseases that can be transmitted as the disease stated in 1a.
3. How can you prevent those diseases?

Infectious diseases are caused by microorganisms known as pathogens which may include viruses, bacteria, fungi and protozoa. Those diseases are called communicable diseases as they can be transmitted from one person to another. They include cholera, malaria, typhoid, HIV and AIDS... Malaria is one of the most dangerous infectious diseases, endemic in Latin America, Africa and South-East Asia. Some infectious diseases can also be from animals to humans.

Some technical terms used when discussing about infectious diseases are:

- **Etiology:** The study of the cause of disease.
- **Epidemiology:** The study of all the factors that contribute to the appearance of a particular disease.
- **Causative agent:** The organism which causes the disease.
- **Vector:** An organism which carries the causative agent of the disease from one person to another or from infected animal to human.
- **Incubation period:** The period of time between the original infection and the appearance of signs and symptoms.
- **Infective period:** The time during which a person is capable of passing the disease on to another person.
- **Carrier:** The person who has been infected but develop no signs and symptom, the carrier can pass the disease on to another person.
- **Prevention:** Measures taken to prevent diseases.
- **Treatment:** Measures taken to cure diseases.
- **Antibody:** Is a protein produced by the body's immune system when it detects harmful substances called antigen.
- **Antigen:** Is any substance that causes your immune system to produce antibodies against it.
- **Host:** A host can be anything or living organism in/on which pathogens can survive.
- **Hygiene:** Practices that help to maintain health and prevent the spread of diseases.
- **Immunity:** Is the ability of the body to resist to infections.

Some groups of communicable diseases

- **Bacterial diseases:** These are diseases caused by bacteria. They include cholera, typhoid, tetanus, tuberculosis, etc.
- **Viral diseases:** These are diseases caused by viruses. They include AIDS, polio, measles, Ebola, etc.
- **Protozoan diseases:** these are diseases caused by protozoa. They include malaria, sleeping sickness, trichomoniasis, etc.
- **Fungal diseases:** These are diseases caused by fungi. They include candidiasis, athlete's foot, ring worms, etc.
- **Worm diseases:** these are diseases caused by worms. They include elephantiasis, bilharzias, etc.
- **Sexually transmitted diseases:** These are diseases transmitted through sexual contact. They include HIV-AIDS, syphilis, gonorrhea, etc. c.

Transmission of infectious diseases

Pathogens can spread when you have direct contact with an infected person. For example, if you have contact with the person's blood, body fluids or open wounds. Pathogens can also be spread through contaminated food, water or air. Infected animals can spread pathogens to people.

The following conditions lead to the spread of an infectious disease:

- A pathogen which causes the disease.
- A source which is an infected organism.

Mode of transmission a pathogen must be able to enter the body of the new host to cause an infection. Infectious diseases follow a pattern of development from the time of infection.

8.1.1. Cholera

Cholera is a good example of a waterborne disease. It is endemic in parts of Asia, particularly India. The organism which causes cholera is a comma shaped motile bacterium called *Vibrio cholerae*.

a. Transmission and symptoms of cholera

The main source of infection is water contaminated by feces with Vibrios. It is estimated that only about one infected person in 50 develops the disease, the rest being carriers. Drinking contaminated water, or washing food or utensils in it, is the most common means of transmission. Direct contamination of

food with feces as a result of poor hygiene is also possible, house flies being the main vector in this last case.

Figure 8.1: People using and drinking contaminated water

b. Signs and symptoms of cholera

Vibrio cholerae multiply in the intestine, releasing a powerful toxin which results in violent inflammation of the intestine and production of the watery diarrhea.

The main sign of the disease is severe diarrhoea due to irritation of the bowel by toxins from the vibrios. The liquid of the feces is so profuse and cloudy like “rice water”.

Abdominal pain and vomiting are also common. Dehydration is rapid and quickly results in death unless rehydration treatment is given.

Figure 8.2: Abdominal pain and vomiting signs cholera

Fever is absent; in fact, the skin feels deathly cold and often damp.

Table 8.1: The features of cholera

Pathogen	Vibrio cholerae
Methods of transmission	Food-borne, water-borne
Global distribution	Asia, Africa, Latin America
Incubation period	Two hours to five days
Site of action of pathogen	Wall of small intestine
Clinical features	Severe diarrhoea ('rice water'), loss of water and salts, dehydration, weakness
Method of diagnosis	Microscopical analysis of faeces

c. Treatment of cholera

The prime cause of death from cholera is dehydration i.e. loss of water with its minerals salts. For that it is obligatory to rehydrate with oral serum which contain mineral salts and sugar,

The fluid loosed may be replaced by administration of a drip food into a vein.

Various antibiotics, such as tetracyclines and chloramphenical, are used to treat cholera. Chloramphenical is effective against tetracycline-resistant vibrios.

d. Prevention of cholera

- Use clean drinking water.
- Proper treatment of sewage and sanitation.
- High standards of public and personal hygiene, particularly in relation to food (such as washing hands after defecation).

Figure 8.3: A person washing hands with a soap after living toilet

- Health education.
- Vaccination is recommended for people visiting areas where cholera is endemic and for those living in such areas. But this vaccine lasts few months.
- Isolation of patients and hygienic disposal of feces and vomit from patients.

e. Failure to eradicate cholera

- Vaccination is not very effective.
- It is a waterborne disease i.e. transmitted through contaminated water.
- Poor sanitation condition in camps.

8.1.2. Tuberculosis

a. Causal agent of tuberculosis

Tuberculosis is caused by bacterium called *Mycobacterium tuberculosis*, first discovered by Robert Koch in 1882. It is sometimes referred to as the tubercle bacillus, bacilli being rod-shaped bacteria. The common form is pulmonary T.B which infects the lungs, although other organs may be affected.

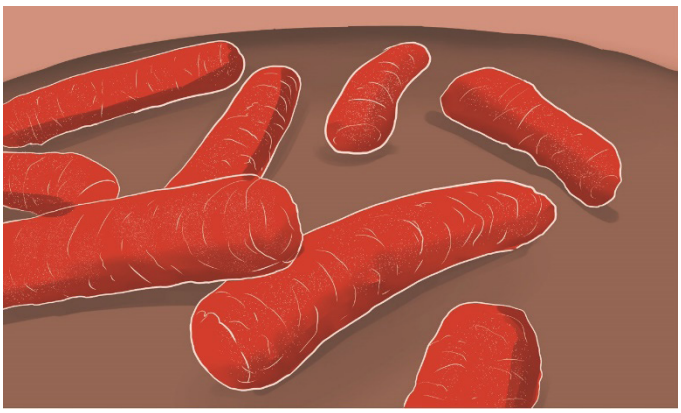


Figure 8.4: Mycobacterium tuberculosis

Two strains of the bacterium may cause the disease, the human and the bovine forms. The latter can be present in cattle and can enter the milk of cows. It is very resistant and can remain alive for long time in milk products as well as in dust.

Table 8.2: The features of TB

Pathogen	Mycobacterium tuberculosis; Mycobacterium bovis
Methods of transmission	Airborne droplets (M. tuberculosis); via undercooked meat and unpasteurised milk (M.bovis)
Global distribution	Worldwide
Incubation period	Few weeks or up to several years
Site of action of pathogen	Primary infection in lungs; secondary infections in lymph nodes, bones and gut
Clinical features	Racking cough, coughing blood, chest pain, shortness of breath, fever, sweating, weight loss
Methods of diagnosis	Microscopic examination of sputum for bacteria, chest X-ray

b. Transmission of tuberculosis

Tuberculosis is mainly airborne disease. The infection is done through the droplets from the patient.

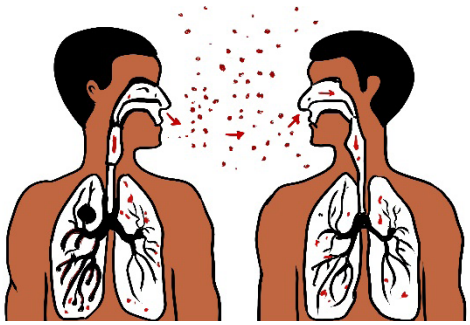


Figure 8.5: Patient transmitting TB germ to another through coughing

It is much less infectious than the common cold and requires prolonged contact between people, poor ventilation and overcrowded living conditions. Other factors include poverty, bad housing, malnutrition, age, smoking and AIDS. In addition, TB is an opportunistic infection, striking many people with a depressed immunity.

c. Signs and symptoms of tuberculosis

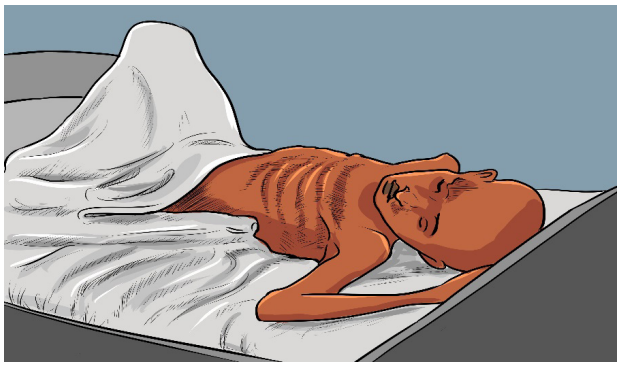


Figure 8.6: A patient on bed suffering from TB

The disease frequently shows itself by vague symptoms such as: loss of appetite; loss of weight; excessive sweating; coughing, appearance of blood in the sputum, pains on the chest, shortness of breath (case of lung tuberculosis).

d. Treatment and prevention of tuberculosis

The development of an effective vaccine against the disease result of the work of Abert Calmette and Camileguérin (BCG). A cure for people already affected by T.B did not come until 1843 when the antibiotic streptomycin was discovered. The number of cases started to fall more rapidly after this and continued to decline aided by introduction of further antibiotics such as rifampicin, isoniazid and others.

e. Failure to eradicate tuberculosis

- Patients can carry pathogen and infection without showing symptoms. Therefore they are difficult to identify and isolate / long period of incubation
- Germs of tuberculosis can survive longer in the house dust
- The disease is related to poverty where many people share the same room and have malnutrition.
- The disease is associated with AIDS that reduced the body immunity
- Long period of medication (6-8 moths), hence patients give up when not yet fully healed. The pathogens then form endospores that resists to medicines.
- The disease is also spread through milk from infected animals. Hence it is difficult to vaccinate.
- Tuberculosis is an airborne disease i.e. spread in air.

8.1.3. Malaria

a. Causal agent of Malaria

Human malaria is caused by infections from four species of plasmodium: ***Plasmodium falciparum***, ***P. vivax***, ***P. ovale***, and ***P. malariae***, each responsible for a different form of the disease.

Malaria is characterized by chills, fever, and, in the most severe cases, coma leading to death.

The ***parasite***, Plasmodium is transmitted by the bite of female mosquitoes (the ***vector***) belonging to the genus ***Anopheles.***

The life cycle of Plasmodium involves both sexual reproduction within the host mosquito, and asexual reproduction within the human being.

It is the reproductive activity of the parasite within the human bloodstream that produces the characteristic recurrent attacks of the disease.

b. Symptoms

Malaria is a very serious disease characterized by severe chills, fever, sweating, fatigue and great thirst. Malaria is caused by a protozoan of the genus Plasmodium.

Victims die of anemia, kidney failure or brain damage. The genus Plasmodium infects humans, and all have life cycles that involve the female anopheles mosquito.

c. Life cycle of Plasmodium

It is the reproductive activity of the parasite within the human bloodstream that produces the characteristic recurrent attacks of the disease.

The infection begins when a mosquito vector injects parasite particles (infectious stages) called sporozoites, present in the mosquito's saliva, into the bloodstream when it feeds on a human blood.

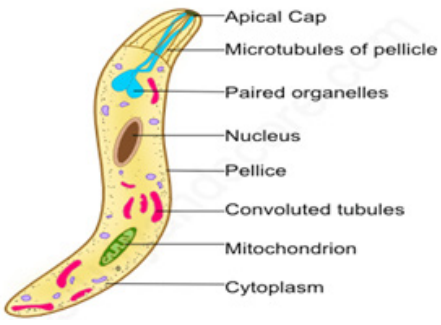


Figure 8.7: Structure of Sporozoite

These enter liver cells where they multiply by asexual reproduction for about 7 to 14 days (the incubation period of the disease) before producing daughter cells called merozoite, which invade red blood cells. The parasites multiply in the red cells, again by asexual reproduction, to produce between 8 and 16 merozoites every 48 or 72 hours, depending on the species of Plasmodium. These merozoites are released by the bursting of the infected red blood cells and the cycle is repeated.

The bursting of the red blood cells and the release of toxic substances cause the characteristic fever of malaria. After a number of such cycles, sexual stages, male and female gametocytes, are produced, and these are taken up by a feeding mosquito, in which the *Plasmodium* life cycle is completed by sexual reproduction, resulting in new sporozoites.

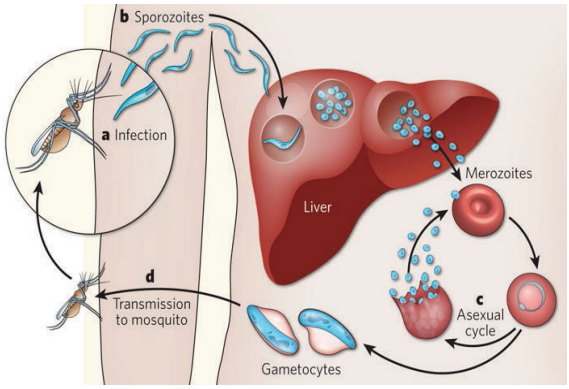


Figure 8.8: Life cycle of Plasmodium

d. Prevention of malaria

- Drainage of stagnant water: The larval stages of the mosquito live in stagnant water, so drainage removes breeding sites. This has had some success.
- Destruction of the breeding sites of the mosquito: The larvae and pupae of mosquitoes obtain their oxygen by means of small tubes which are pushed through the water surface film. Thus any method of blocking these tubes will result in the death of the intermediate life stages of the mosquito (petrol, oil….)
- Destruction of the adult mosquitoes: This is aimed at killing the mosquitoes that enter houses. Thus, the indoor surfaces are sprayed with a persistent insecticide.
- Use of mosquito nets.

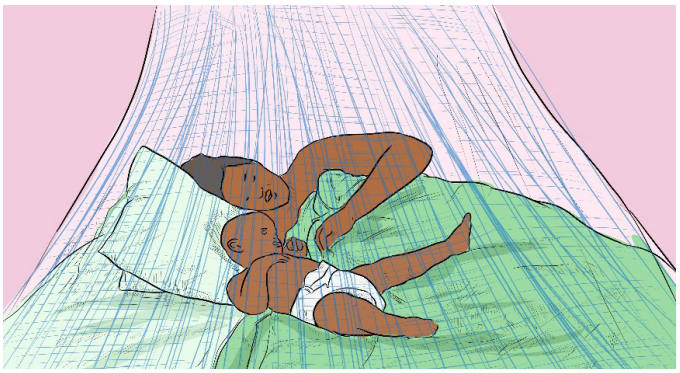


Figure 8.9: Sleeping under mosquito nets.

e. Occurrence of malaria

The disease now occurs in tropical and subtropical regions of the world, and its distribution is limited by conditions that are inimical to the development of the mosquito vector, such as temperature and altitude.

Malaria is endemic in tropics because:

- Tropical climate provides the best breeding and living conditions for the Anopheles mosquito which transmits malaria
- The Anopheles cycle requires areas of stagnant water and these are common within tropics
- In the tropical areas there is presence of bushes or abundant vegetation which makes suitable habitat for mosquitoes

Plasmodium needs temperature in excess of 20□C for it to complete its cycle within the mosquito.

Table 8.3: The features of malaria

Pathogen	Plasmodium falciparum, P. vivax, P. malariae
Methods of transmission	Insect vector: female anopheles mosquito
Global distribution	Throughout the tropics and subtropics (endemic in 106 countries)
Incubation period	From a week to a year
Site of action of pathogen	Liver, red blood cells, brain
Clinical features	Fever, anemia, nausea, headache, muscle pain, shivering, sweating, enlarged spleen.
Method of diagnosis	Microscopic examination of blood; dip stick test for malaria antigens in blood.

f. Treatment of malaria

The common drugs used in treatment of malaria include the following: Chloroquine, Atovaquone-proguanil (Malarone), Artemether-lumefantrine (Coartem), Mefloquine, Quinine, Doxycycline (used in combination with quinine)

g. Failure to eradicate malaria

- There is no effective vaccine against malaria
- The pathogens are transmitted by mosquitoes which are not eradicated.
- The plasmodium have become resistant to different anti-malarial drugs
- Ignorance of some people toward the disease and how it is spread.

8.1.4. HIV/AIDS and other sexual transmission diseases (STD)

Human Immunodeficiency Virus (HIV) which causes AIDS (Acquired Immune Deficiency Syndrome), is an acquired viral infection that destroys important white blood cells and weakens the immune system. People with HIV/AIDS can become seriously ill with infections that most people can fight off.

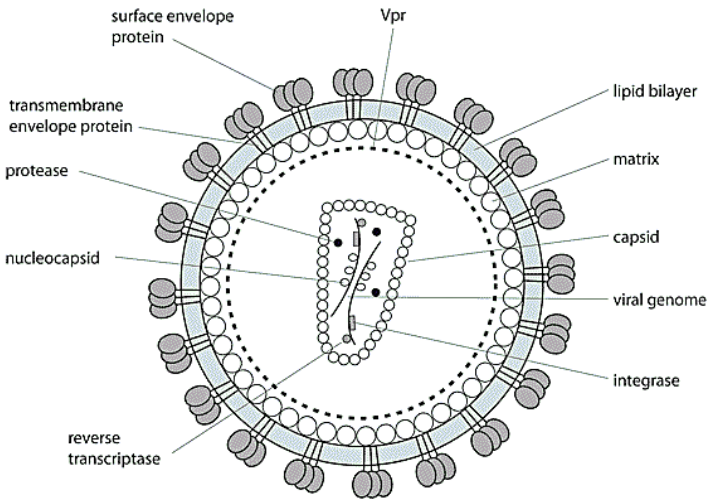


Figure 8.10: Structure of HIV

This is an RNA virus. The virus replicates inside the T4 lymphocytes or helper T cells. Thus these cells can no longer help or induce other T cells, called killer cells, to fight invaders. The body's immune system breaks down leaving the patient exposed to a variety of diseases called opportunistic infections. AIDS is not a disease; it is a collection of these opportunistic diseases associated with immunodeficiency caused by HIV infection. It is

important to realize, however, that the infection with the HIV virus does not necessarily result in AIDS. As with other diseases, some people remain symptomless and are therefore called carriers.

Transmission, signs and symptoms

The HIV virus can only survive in body fluids and is transmitted by blood or semen. In 90% of cases the transmission is achieved by sexual contact.

People can contract the disease as follows:

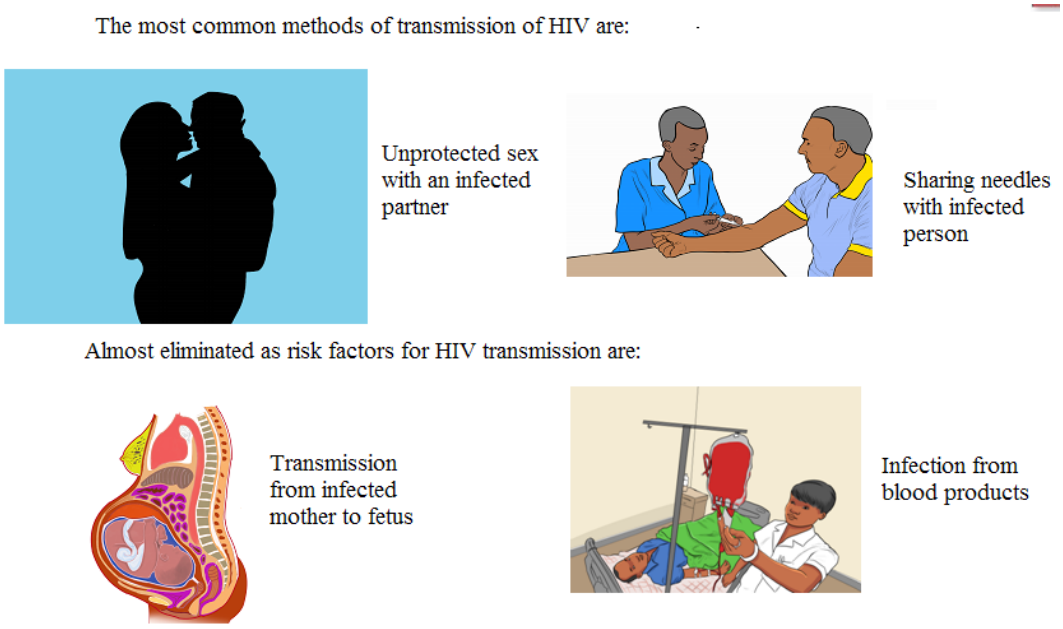


Figure 8.11: Modes of HIV/AIDS transmission

- **Intimate sexual contact:** The most frequent mode of transmission of HIV is through sexual contact with an infected person.
- **Infected blood entering bloodstream:** By means of unsterilized needles and syringes. Unfortunately the disease can be contracted after being given blood or blood products already infected with HIV. Close contact between infected and non infected people through cuts and open wounds has also been known to pass on the virus.
- **From mother to baby:** An infected pregnant woman can pass on the virus to her baby through the placenta, at birth or through breast milk during suckling. The chances of infection being transmitted from the mother to her baby are currently estimated to be 25-50%.

Test for the disease

A blood test is used to tell whether or not a person has been infected by the HIV virus. Under normal circumstances the immune system reacts to infection by producing antibodies and when the HIV virus enters the body, anti-HIV antibodies are produced. The blood of the person being tested is added to HIV proteins which have been commercially prepared. If there are anti-HIV antibodies in the blood sample they will bind to the viral proteins and the person is described as HIV positive. However, if the test proves negative that person may still be infected. This is because it takes up to three months or longer for HIV antibodies to be produced after infection.

Prevention of the disease

There are many precautions which can be followed in trying to prevent the disease:

- Abstinence
- Fidelity or restriction to one sex partner and the absence of promiscuity will also clearly reduce the risk of infection.
- Use of a sheath or condomduring intercourse can prevent the virus from infecting through blood or semen.
- A reduction in the spread of HIV can be brought about by the use of clean needles and syringes by drugs addicts.
- The blood donated should be tested for the presence of antibodies to HIV which indicates whether or not the donor is infected. Blood containing these antibodies is not used.
- Educating the people about the disease.
- Taking antiretroviral drugs during pregnancy and delivery.
- To avoid breastfeeding and to administer antiretroviral drugs to the newborn.

Table 8.4: The features of HIV/AIDS

Pathogen	Human immune deficiency virus
Methods of transmission	In semen and vaginal fluids during sexual intercourse, infected blood or blood products, contaminated hypodermic syringes, mother to fetus across the placenta, at birth, mother to infant in breast milk
Global distribution	Worldwide, especially in sub-Saharan Africa and South-East Asia
Incubation period	Initial incubation a few weeks, but up to ten years or more before symptoms of AIDS may develop
Site of action of pathogen	T helper lymphocytes, macrophages, brain cells
Clinical features	HIV infection-flu-like symptoms and then symptomless AIDS-opportunistic infections including pneumonia, TB and cancers; weight loss, diarrhoea, fever, sweating, dementia.
Method of diagnosis	Testing blood, saliva or urine for the presence of antibodies produced against HIV.

Other sexual transmission diseases (STDs)

Sexually transmitted diseases (STDs), also referred to as sexually transmitted infections (STIs) and venereal diseases (VDs), are illnesses that have a significant probability of transmission between humans by means of unprotected sexual intercourse, including vaginal intercourse, anal sex and oral sex. Some STIs can also be contracted by using drug needles after their use by an infected person, as well as through any incident involving the contact of a wound with contaminated blood or through childbirth or breastfeeding.

Sexually transmitted infections have been well known for hundreds of years, and venereology is the branch of medicine that studies these diseases. While in the past, these illnesses have mostly been referred to as STDs or VD, the term **sexually transmitted infections (STIs)** has been preferred by many up-to-date medical sources, as it has a broader range of meaning; a person may be *infected*, and may potentially infect others, without having a **disease.**

For example, there are 19 million new cases of sexually transmitted infections every year in the United States, and, in 2005, the World Health Organization estimated that 448 million people aged 15-49 were being infected a year with curable STIs (such as syphilis, gonorrhea and chlamydia).

Until the 1990s, STIs were commonly known as **venereal diseases,** the word venereal being derived from the Latin word venereus, and meaning relating to sexual intercourse or desire, ultimately derived from Venus, the Roman goddess of love. **Social disease** was a phrase used as a euphemism.

While many people with STDs show no signs or symptoms of their infection, when there are signs of STDs they are most likely to be in the genital area. The genital area in women includes the vulva (the area around the vagina including the lips), vagina (the opening where menstrual blood comes out), buttocks, urethra (the opening above the vagina where urine comes out) and anus (the opening where a bowel movement comes out). The genital area in men includes the penis, scrotum ("balls"), urethra, and anus.

What are the symptoms of STDs?

Sometimes, there are no symptoms of STDs. If symptoms are present, they may include one or more of the following:

- Bumps, sores, or warts near the mouth, anus, penis, or vagina.
- Swelling or redness near the penis or vagina.
- Skin rash.
- Painful urination.
- Weight loss, loose stools, night sweats.
- Aches, pains, fever, and chills.
- Yellowing of the skin (jaundice).
- Discharge from the penis or vagina. (Vaginal discharge may have an odor.)
- Bleeding from the vagina other than during a monthly period.
- Painful sex.
- Severe itching near the penis or vagina.

Examples of these diseases are chlamydia, gonorrhoea, syphillis and, HIV and AIDS.

1. **Chlamydia**

Chlamydia is caused by the bacterium ***Chlamydia trachomatis***.

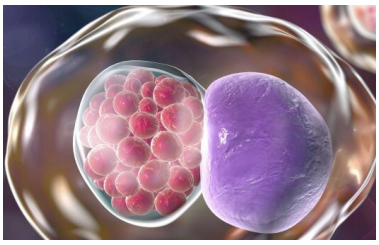


Figure 8.12: Chlamydia trachomatis

The disease is spread by oral, vaginal or anal sex, and also through touch, for example, touching the eyes with a contaminated hand, may lead to conjunctivitis. Chlamydia can also be passed to the infant during birth.

It causes inflammation of the cervix in women, urethra and rectum in both men and women. Occasionally, other parts of the body like eyelids and throat may be affected. Any sexually active person is at risk of contracting the disease. However, it is more common in young people.

The disease is known as a 'silent' infection because it is mainly asymptomatic, thus the symptoms can be mild or be confused with gonorrhea.

Signs and symptoms of chlamydia

In males

- Pain when passing out urine.
- White discharge from the penis.
- The testicles may be painful or swollen.
- Swelling of skin around the anus.

In females

- Painful and frequent urination.
- Smelly yellowish and abnormal vaginal discharge.
- Pain in the lower abdomen.
- Swollen skin in the vagina or around the anus.

Treatment

Chlamydia is easily treated using antibiotics usually Azithromycin or doxyclyne.

2. Gonorrhoea

Gonorrhoea is transmitted through sexual contact with the penis, vagina, mouth or anus of an infected partner. Gonorrhoea can also be spread from mother to baby during childbirth. Gonorrhoea is caused by a bacterium called *Neisseria gonorrhoeae*.

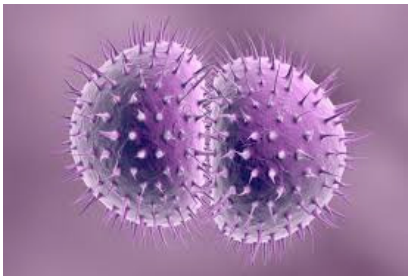


Figure 8.13: ***Neisseria gonorrhoeae***

The bacteria attaches on the epithelial cells of the vagina or male urethra. This results in inflammation and discharge of pus. If left untreated, the infection spreads to the other reproductive parts and may eventually block the passages resulting to infertility.

Signs and symptoms of gonorrhea

Some men with gonorrhea may have no symptoms at all. However, men who do have symptoms may have:

- A burning sensation when urinating.
- A white, yellow, or green discharge from the penis.
- Painful or swollen testicles (although this is less common).

Most women with gonorrhea do not have any symptoms. Even when a woman has symptoms, they are often mild and can be mistaken for a bladder or vaginal infection.

Women with gonorrhea are at risk of developing serious complications from the infection, even if they do not have any symptoms.

Symptoms in women can include

- Painful or burning sensation when urinating.
- Increased vaginal discharge.
- Vaginal bleeding between periods.

Treatment

Gonorrhea can be treated using antibiotics like penicillin.

3. Syphilis

Syphilis is transmitted from person to person by direct contact with a syphilitic sore, known as a chancre. Chancres occur mainly on the external genitals, vagina, anus or in the rectum. Chancres also can occur on the lips and in the mouth.

Transmission of syphilis occurs during vaginal, anal or oral sex.

Syphilis is caused by a bacteria called ***Treponema pallidum.***

Figure 8.14: Treponema pallidum

The bacterial infection progresses through several stages:

- In the primary stage, small hard painless sores develop at the site of infection usually the penis and the vagina.
- The disease enters secondary stage several weeks later characterized by rashes on the skin and mild fever. These symptoms subside after a few weeks followed by a latent asymptomatic period.
- In the tertiary stage, lesions develop and cause extensive tissue damage that may lead to paralysis, insanity, blindness and eventually death.

Treatment

Antibiotics like penicillin, erythromycin or tetracycline are used to treat syphilis although some strains can be resistant to certain antibiotics.

The following are general ways of reducing STDs and HIV infection:

a). Abstinence is the only sure way to prevent STDs.

b). Being faithful to one trusted partner.

c). Using condoms every time when engaging in sexual intercourse. Condoms are not 100% effective at preventing disease or pregnancy. However, they are extremely effective if used properly.

d). Reduce the number of sexual partners.

e). Avoid sharing towels or underclothing.

f). Get a vaccination for hepatitis B.

g). Get tested for HIV.

h). Avoiding alcohol consumption and abuse of drugs. Individuals who are drunk or on drugs often fail to have safe sex.

Table 8.5: Summary of sexual transmitted infections

STI	SYMPTOMS
Chlamydia caused by a bacterium called ***Chlamydia trachomatis***	Most women have no symptoms. Women with symptoms may have: • Abnormal vaginal discharge • Burning when urinating • Bleeding between periods Infections that are not treated, even if there are no symptoms, can lead to: • Lower abdominal pain
	• Nausea • Fever • Pain during sex
Genital herpes caused by a DNA virus	Some people may have no symptoms. During an "outbreak," the symptoms are clear: • Small red bumps, blisters, or open sores where the virus entered the body, such as on the penis, vagina, or mouth • Vaginal discharge • Fever • Headache • Muscle aches • Pain when urinating • Itching, burning, or swollen glands in genital area • Pain in legs, buttocks, or genital area Symptoms may go away and then come back. Sores heal after 2 to 4 weeks.

Gonorrhoea caused by a bacterium called ***Neisseria gonorrhoea***	Symptoms are often mild, but most women have no symptoms. If symptoms are present, they most often appear within 10 days of becoming infected. Symptoms are: • Pain or burning when urinating • Yellowish and sometimes bloody vaginal discharge • Bleeding between periods • Pain during sex • Heavy bleeding during periods Infection that occurs in the throat, eye, or anus also might have symptoms in these parts of the body.
AIDS caused by a retrovirus called HIV	Some women may have no symptoms for 10 years or more. About half of people with HIV get flu-like symptoms about 3 to 6 weeks after becoming infected. Symptoms people can have for months or even years before the onset of AIDS include: • Fevers and night sweats • Feeling very tired • Quick weight loss • Headache • Enlarged lymph nodes • Diarrhoea, vomiting, and upset stomach • Mouth, genital, or anal sores • Dry cough • Rash or flaky skin • Short-term memory loss
Syphilis caused by a bacterium called *Treponema pallidum*	Syphilis progresses in stages. Symptoms of the primary stage are: • A single, painless sore appearing 10 to 90 days after infection. It can appear in the genital area, mouth, or other parts of the body. The sore goes away on its own. If the infection is not treated, it moves to the secondary stage. This stage starts 3 to 6 weeks after the sore appears. Symptoms of the secondary stage are: • Skin rash with rough, red or reddish-brown spots on the hands and feet that usually does not itch and clears on its own • Fever • Sore throat and swollen glands • Patchy hair loss • Headaches and muscle aches • Weight loss • Tiredness

Trichomoniasis (sometimes called "trich") caused by a protozoan called *Trichomonas vaginalis*	Many women do not have symptoms. Symptoms usually appear 5 to 28 days after exposure and can include: • Yellow, green, or gray vaginal discharge (often foamy) with a strong odor • Discomfort during sex and when urinating • Itching or discomfort in the genital area • Lower abdominal pain (rarely)
Candidiasis caused by a fungus called *Candida albicans*	Severe irritation, extreme itching in the vaginal area, soreness and redness in the vaginal area, painful intercourse in females. In men the symptoms include: red rash on penis, itching or burning on the tip of the penis.
Chancroid is caused by a bacterium.	Symptoms may include genital sores, vaginal discharge, a burning feeling when urinating, and swollen lymph nodes in the groin. It can be spread by vaginal or anal sex or skin-to-skin contact with sores. Chancroid can be treated with antibiotics.

8.1.5. Deficiency diseases

Deficiency diseases are diseases that are caused by a dietary deficiency of specific nutrients, especially a vitamin or mineral, possibly stemming from insufficient intake, digestion, absorption, or utilization of a nutrient. The examples of common deficiency diseases are Kwashiorkor, Marasmus, and Vitamin deficiencies .

1. Kwashiorkor

Kwashiorkor is a form of severe protein malnutrition characterized by edema, and an enlarged liver with fatty infiltrates.

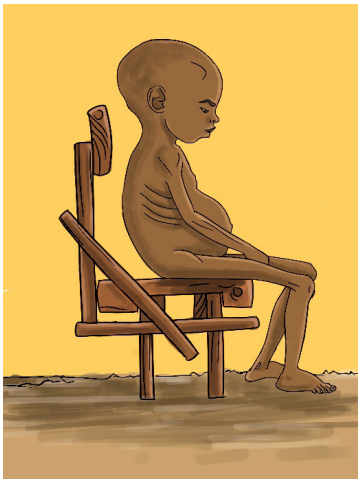


Figure 8.15: Child suffering from kwashiorkor

Signs and symptoms of kwashiorkor

The following are well defined symptoms and signs:

- Pitting edema(swelling of the ankles and feet).
- Distended abdomen,
- An enlarged liver with fatty infiltrates,
- Thinning of hair,
- Loss of teeth,
- Skin depigmentation and dermatitis.

Children suffering from kwashiorkor often develop irritability and anorexia.

Prevention

Kwashiorkor can be prevented by making sure you eat enough calories and protein –rich foods.

Treatment of the disease

Generally, the disease can be treated by adding protein to the diet; however, it can have a long-term impact on a child's physical and mental development, and in severe cases may lead to death.

The person should be given small but frequent rations in every two to four hours. During the first week, he or she must be given a diet high in sugar and enriched in protein as well as essential elements: sweet milk with mineral salts and vitamins. The diet may include lactases so that children who have developed lactose intolerance can ingest dairy products and antibiotics to compensate for immunodeficiency.

After two to three weeks, the milk is replaced by boiled cereals fortified with minerals and vitamins until the person's mass is at least 80% of normal weight

2. Marasmus

Marasmus is a form of severe malnutrition. It usually occurs in children. It typically occurs in developing countries. Marasmus can be life-threatening, but you can get treatment for it.

Causes of marasmus

Nutrient deficiency is the main cause of marasmus. It occurs in children that do not eat enough protein, calories, carbohydrates, and other important nutrients. This is usually due to poverty and a scarcity of food.

Symptoms of marasmus

In children with marasmus, the following are symptoms :

- Chronic diarrhea
- Respiratory infections
- Intellectual disability
- Being underweight
- Stunted growth
- Subcutaneous fat is the layer of fat just under the skin.
- Dry skin and brittle hair
- Children may look older.
- Children are short-tempered and irritable

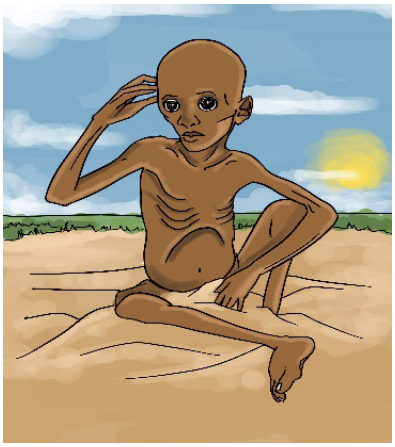


Figure 8.16: Child suffering from marasmus

3. Other deficiency diseases

a. Goitre

The goiter is the swelling in the neck caused by lack of an adequate iodine in the diet the pregnant mother

Prevention and treatment

- Eat food rich in iodine
- Go to doctor for treatment and the surgery is done to remove the swelling

b. Anaemia

This disease is due to lack of an adequate Iron mineral in the diet. The iron is essential for formation of haemoglobin pigment of blood contained by red blood cells. If haemoglobin is not formed due to lack of iron, the oxygen will

not be transported into the body, hence the person with anaemia becomes tired easily and has little energy to do a given work due to inadequate oxygen in the body.

Signs and symptoms

- Shortness of breath
- Filling weak and tired quickly
- Filling one has no energy to work
- Inside the eyelids and tongue looks pale
- Short attention span

Prevention and treatment

- Regular taking lots of leafy vegetables to provide iron
- Eat food with plenty of calcium and vitamin D

8.1.6. Vitamin deficiencies

1. Scurvy

Figure 8.17: Scurvy

Scurvy happens when there is a lack of vitamin C, or ascorbic acid. The deficiency leads to symptoms of weakness, anemia, gum disease, and skin problems.

This is because vitamin C is needed for making collagen, an important component in connective tissues. Connective tissues are essential for structure and support in the body, including the structure of blood vessels. A lack of vitamin C will also affect the immune system, absorption of iron, metabolism of cholesterol and other functions.

Symptoms

Symptoms of vitamin C deficiency can start to appear after 8 to 12 weeks. Early signs include a loss of appetite, weight loss, fatigue, irritability, and

lethargy.

Within 1 to 3months, there may be signs of:

- Anemia
- Myalgia, or pain, including bone pain
- Swelling, or edema
- Petechiae, or small red spots resulting from bleeding under the skin
- Corkscrew hairs
- Gum disease and loss of teeth
- Poor wound healing
- Shortness of breath
- Mood changes, and depression

Treatment

Treatment involves administering vitamin C supplements by mouth or by injection.

The recommended dosage is:

- 1 to 2 grams (g) per day for 2 to 3 days
- 500 milligrams (mg) for the next 7 days
- 100 mg for 1 to 3 months

Within 24 hours, patients can expect to see an improvement in fatigue, lethargy, pain, anorexia, and confusion. Bruising, bleeding, and weakness start to resolve within 1 to 2 weeks.

After 3 months, a complete recovery is possible. Long-term effects are unlikely, except in the case of severe dental damage.

Prevention

Scurvy can be prevented by consuming enough vitamin C, preferably in the diet, but sometimes as a supplement.

During pregnancy, women should consume 85 mg of vitamin C, rising to 120 mg while breastfeeding.

Smokers need 35 mg more than non-smokers every day.

2. Rickets

Figure 8.18: Rickets

Rickets is a skeletal disorder that's caused by a lack of vitamin D, calcium, or phosphate which are important nutrients for the development of strong, healthy bones. The sufferer of rickets may have weak and soft bones, stunted growth, and, in severe cases, skeletal deformities.

Symptoms of rickets

Symptoms of rickets include:

- Pain or tenderness in the bones of the arms, legs, pelvis, or spine
- Stunted growth and short stature
- Bone fractures
- Muscle cramps
- Teeth deformities, such as: delayed tooth formation, holes in the enamel, abscesses, defects in the tooth structure, an increased number of cavities
- Skeletal deformities, including: an oddly shaped skull, bowlegs, or legs that bow out, bumps in the ribcage, a protruding breastbone, a curved spine, pelvic deformities

Rickets treatment

Treatment for rickets focuses on replacing the missing vitamin or mineral in the body. This will eliminate most of the symptoms associated with rickets. The exposure to sunlight, if possible is recommended. To consume food products high in vitamin D, such as fish, liver, milk, and eggs is also necessary.

If skeletal deformities are present, the child may need braces to position their bones correctly as they grow. In severe cases, the child may need corrective surgery.

For hereditary rickets, a combination of phosphate supplements and high levels of a special form of vitamin D are required to treat the disease.

Prevention

- The best way to prevent rickets is to eat a diet that includes adequate amounts of calcium, phosphorous, and vitamin D. People with kidney disorders should have their calcium and phosphate levels monitored on a regular basis by their doctors.
- Rickets can also be prevented with moderate sun exposure. We need to expose the hands and face to sunlight a few times a week during the spring and summer months to prevent rickets.

3. Beriberi

Beriberi is a disease caused by a vitamin B-1 deficiency, also known as thiamine deficiency. There are two types of the disease: wet beriberi and dry beriberi.

Wet beriberi affects the heart and circulatory system. In extreme cases, wet beriberi can cause heart failure.

Dry beriberi damages the nerves and can lead to decreased muscle strength and eventually, muscle paralysis. Beriberi can be life-threatening if it isn't treated.

Today, beriberi mostly occurs in people with an alcohol use disorder. Still, the disease can be seen in women who have extreme nausea and vomiting in pregnancy, in people with AIDS, and after bariatric surgery.

Symptoms of beriberi

The symptoms of beriberi vary depending on the type.

a). Wet beriberi symptoms include:

- Shortness of breath during physical activity
- Waking up short of breath
- Rapid heart rate
- Swollen lower legs

b). Dry beriberi symptoms include:

- Decreased muscle function, particularly in the lower legs
- Tingling or loss of feeling in the feet and hands
- Pain
- Mental confusion
- Difficulty speaking
- Vomiting

- Involuntary eye movement
- Paralysis

Beriberi treatment

Beriberi is easily treated with thiamine supplements. Doctor may prescribe a thiamine shot or pill. For severe cases, a healthcare professional will administer intravenous thiamine.

Your progress will be monitored with follow-up blood tests to see how well your body is absorbing the vitamin.

Prevention

- To prevent beriberi, eat a healthy, balanced diet that includes foods rich in thiamine. These include: beans and legumes, seeds, meat, fish, whole grains, nuts, dairy, certain vegetables, such as asparagus, acorn squash, Brussels sprouts, spinach, and beet greens and breakfast cereals that are enriched with thiamine
- Cooking or processing any of the foods listed above decreases their thiamine content.
- Always be sure to purchase infant formula from a reliable source.
- Limiting alcohol consumption will reduce your risk of developing beriberi.

4. Pellagra

Pellagra is a disease caused by low levels of niacin, also known as vitamin B-3. It's marked by dementia, diarrhea, and dermatitis, also known as "the three Ds". If left untreated, pellagra can be fatal.

Symptoms

The main symptoms of pellagra are dermatitis, dementia, and diarrhea. Dermatitis related to pellagra usually causes a rash on the face, lips, feet, or hands. In some people, dermatitis forms around the neck.

Additional dermatitis symptoms include:

- Red, flaky skin.
- Areas of discoloration, ranging from red to brown
- Thick, crusty, scaly, or cracked skin
- Itchy, burning patches of skin

As the disease progresses, possible dementia symptoms include: apathy, depression, confusion, irritability, or mood changes, headaches, restlessness or anxiety, disorientation or delusions

Other possible pellagra symptoms include: sores on the lips, tongue, or gums, decreased appetite, trouble eating and drinking, nausea and vomiting

Causes

There are two types of pellagra, known as primary pellagra and secondary pellagra.

Primary pellagra is caused by diets low in niacin or tryptophan. Tryptophan can be converted to niacin in the body, so not getting enough can cause niacin deficiency.

Secondary pellagra occurs when your body can't absorb niacin.

There are things that can prevent from absorbing niacin like:alcoholism, eating disorders, certain medications, including anti-convulsants and immunosuppressive drugs, gastrointestinal diseases, such as ulcerative colitis, cirrhosis of the liver, carcinoid tumors, Hartnup disease

Treatment

Primary pellagra can be treated by dietary changes and a niacin or nicotinamide supplement.

Nicotinamide is another form of vitamin B3. If left untreated, primary pellagra usually causes death after four or five years.

Treating secondary pellagra usually focuses on treating the underlying cause. However, some cases of secondary pellagra also respond well to taking niacin or nicotinamide either orally or intravenously.

5. Night blindness

This is a disease caused by lack of Vitamin A in the diet. The main sources of vitamin A are green leafy vegetables, liver, egg yolk and milk products.

Figure 8.19: Sources of Vitamin A

Symptoms

- Inability of seeing during night

Prevention

- Provide diet with a plenty of vitamin A

8.1.7. Worm diseases

In our country, even in other African country many people suffer from worm diseases. There are many worms that can infest people due to poor sanitation. Worms causing infection in people are parasites that live and breed mostly in the intestine. Infection is caused by worms such as roundworms, hookworms and tapeworms.

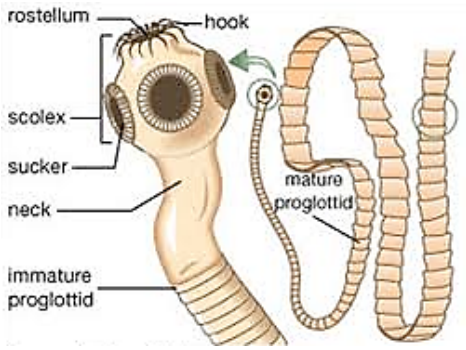


Figure 8.20: Tapeworm structure

1. Ascariasis

Ascariasis is an infection of the small intestine and it is the most common roundworm infection. Ascariasis is most common in places without modern sanitation. People get the parasite through unsafe food and water. The infection usually causes no symptoms, but a high number of roundworms (heavier infestations) can lead to problems in the lungs or intestines.

Causes an ascariasis infection

You can become infected with ascariasis after accidentally ingesting the eggs of the ***Ascaris lumbricoides*** roundworm. The eggs can be found in soil contaminated by human feces or uncooked food contaminated by soil that contains roundworm eggs.

Figure 8.21: Structure of Ascaris lumbricoides worm

Children often become infected when they put their hands in their mouths after playing in contaminated soil, even it can also be passed directly from person to person.

Symptoms of ascariasis?

People with ascariasis often have no symptoms. Symptoms become more noticeable when the roundworm infestation grows.

Roundworms in your lungs can cause:

- Coughing or gagging
- Wheezing or shortness of breath
- Aspiration pneumonia (rarely)
- Blood in mucus
- Chest discomfort
- Fever

Roundworms in your intestines can cause:

- Nausea
- Vomiting
- Irregular stools or diarrhea
- intestinal blockage, which causes severe pain and vomiting
- Loss of appetite
- visible worms in the stool
- Abdominal discomfort or pain
- Weight loss
- Growth impairment in children due to malabsorption.

Life cycle of the roundworm

After ingestion, the ***A. lumbricoides*** roundworm reproduces inside your intestine. The worm goes through several stages:

- Swallowed eggs first hatch in the intestine.
- The larvae then move through the bloodstream to your lungs.
- After maturing, the roundworms leave your lungs and travel to your throat.
- You'll either cough up or swallow the roundworms in your throat. The worms that are swallowed will travel back to your intestine.
- Once they're back in your intestine, the worms will mate and lay more eggs.
- The cycle continues. Some eggs are excreted through your feces. Other eggs hatch and return to the lungs.

Environmental risk factors for ascariasis include:

- Lack of modern hygiene and sanitation infrastructure.
- Use of human feces for fertilizer.
- Living in or visiting a tropical or subtropical climate.
- Exposure to an environment where dirt might be ingested.

You can limit your exposure to roundworms by avoiding unsafe food and water. Keeping your immediate environment clean also helps. This includes laundering clothing exposed to unsanitary conditions and cleaning cooking surfaces well.

You should make sure to take precautions if you're visiting a remote area. It's important to:

- Always wash your hands with soap and water before eating or preparing food.
- Boil or filter your water.
- Inspect food preparation facilities.
- Avoid unclean common areas for bathing.
- Peel or cook unwashed vegetables and fruit in regions that lack sanitation infrastructure or that use human feces for fertilizer.

Complications of ascariasis

Dangerous complications, including:

- Intestinal blockage. Intestinal blockage occurs when a mass of worms blocks your intestines, causing severe pain and vomiting. Intestinal blockage is considered a medical emergency and requires treatment right away.

- Duct blockage. Duct blockage occurs when the worms block the small passageways to your liver or pancreas.
- Nutritional deficiency. Infections that lead to loss of appetite and poor absorption of nutrients put children at risk of not getting enough nutrients, which can affect their growth

Children are more likely to have gastrointestinal complications because the smaller size of their intestines increases their chances of having an intestinal blockage.

How ascariasis is treated

Doctors usually treat roundworm with antiparasitic drugs. Medications most commonly used include:

- Albendazole (Albenza)
- Ivermectin (Stromectol)
- Mebendazole (Vermox)

If you have an advanced case, you may need other treatment. Your doctor may recommend surgery to control a larger infestation. You'll need surgery if the roundworms are completely blocking your intestines.

The best way to avoid ascariasis is by:

- Practicing good hygiene. That means always wash your hands with soap and water before eating or handling food, and after using the bathroom. Teach your children to do the same.
- Dining only at reputable places.
- Drinking only bottled water and avoiding raw fruits and vegetables unless you're able to wash and peel them yourself when you're in places without modern sanitation.

2. Schistosomiasis or Bilharzia

This disease is also known as snail fever and bilharzia. It is a disease caused by parasitic flat worms called *Schistosoma mansonni*.The urinary tract or the intestines may be infected.

Symptoms include abdominal pain, diarrhea, bloody stool, or blood in the urine. Those who have been infected for a long time may experience liver damage, kidney failure, infertility, or bladder cancer. In children, it may cause poor growth and learning difficulty

Modes of transmission

Infected individuals release *Schistosoma* eggs into water via their fecal material or urine. After larvaehatch from these eggs, the larvae infect a very specific type of freshwater snail(Vector of the disease). The *Schistosoma* larvae undergo the next phase of their life cycles in these snails, spending their time reproducing and developing. Once this step has been completed, the parasite leaves the snail and enters the water column. The parasite can live in the water for only 48 hours without a human host. Once a host has been found, the worm enters its blood vessels. For several weeks, the worm remains in the vessels, continuing its development into its adult phase. When maturity is reached, mating occurs and eggs are produced. Eggs enter the bladder/intestine and are excreted through urine and feces and the process repeats. If the eggs do not get excreted, they can become engrained in the body tissues and cause a variety of problems such as immune reactions and organ damage.

Humans encounter larvae of the *Schistosoma* parasite when they enter contaminated water while bathing, playing, swimming, washing, fishing, or walking through the water.

Modes of prevention

- Avoiding drinking or coming into contact with contaminated water in areas where schistosomiasis is common.
- Increasing access to clean water and sanitation
- Snail control
- Health education.

Treatment

There are two drugs available,praziquantel and oxamniquine for the treatment of schistosomiasis. They are considered equivalent in relation to efficacy against *S. mansoni* and safety. Because of praziquantel's lower cost per treatment, and oxaminiquine's lack of efficacy against the urogenital form of the disease caused by *S. haematobium*, in general praziquantel is considered the first option for treatment. Schistosomiasis is treatable by taking by mouth a single dose of the drug praziquantel annually.

Figure 8.22: Praziquantel for Schistosomiasis treatment

3. **Elephantiasis or lymphatic filariasis**

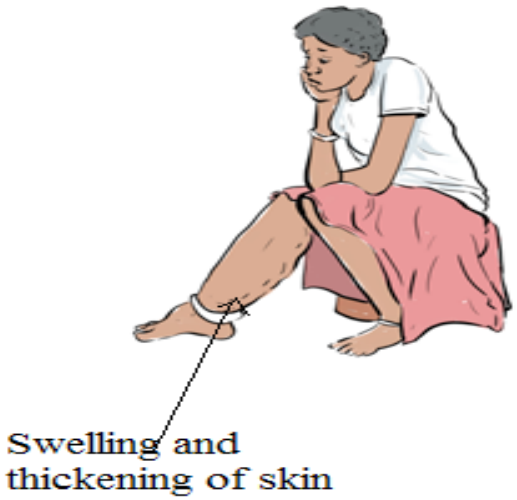


Figure 8.23: Victime of elephantiasis lymphatic filariasis disease

Elephantiasis, also known as Lymphatic filariasis, is a human disease caused by parasitic worms known as filarial worms.Most cases of the disease have no symptoms. Some people, however, develop a syndrome called elephantiasis, which is marked by severe swelling in the arms, legs,breasts or genitals. The skin may become thicker as well, and the condition may become painful.

The worms are spread by the bites of infected mosquitoes. Three types of worms are known to cause the disease: Wuchereria bancrofti,Brugia malayi and Bruggia timori with Wuchereria bancrofti being the most common. These worms damage the lymphatic system. The disease is diagnosed by microscopic examination of blood collected during the night. The blood is typically examined as a smear after being stained with Giesma stain.

Figure 8.24: Wuchereria bancrofti

Table 8.5: The features of Elephantiasis

Symptoms	None, severe swelling of the arms, legs, or genitals
Causes	Filarial worms
Vector	mosquitos
Prevention	Bed nets, mass deworming
Diagnostic method	Microscopic examination of blood
Treatment	Albendazole with ivermectin or diethylcarbamazine

4. Ankilostomiasis

Figure 8.25: Victim of ankilostomiasis and drugs for treatment

Ankilostomiasis is a hookworm disease caused by infection with Ancylostoma hookworms. Ankilostomiasis is caused when hookworms, present in large numbers, produce an iron deficiency anemia by sucking blood from the host's intestinal walls.

Signs and symptoms

Depending on the organism, the signs and symptoms vary. Ancylostoma duodenale and Necator americanus can enter the blood stream while Ancylostoma braziliensis cannot.

Transmission

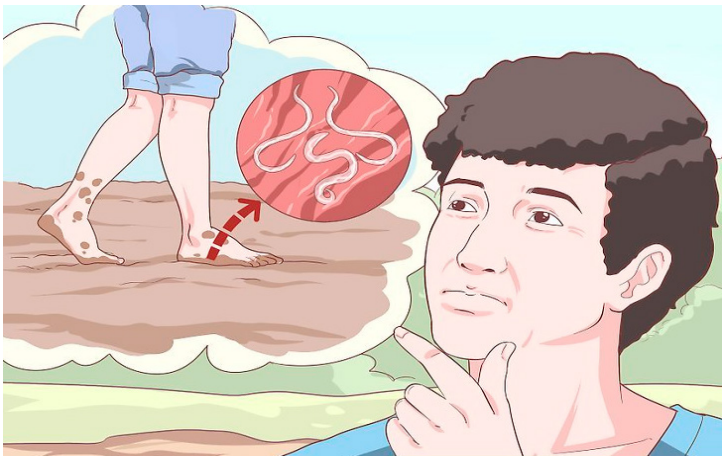


Figure 8.26: One way of hookworm infection

The infection is usually contracted by people walking barefoot over contaminated soil. In penetrating the skin, the larvae may cause an allergic reaction. It is due to the itchy patch at the site of entry that the early infection gets its nickname “ground itch”. Once larvae have broken through the skin, they enter the bloodstream and are carried to the lungs (however, unlike ascarids, hookworms do not usually cause pneumonia). The larvae migrate from the lungs up the windpipe to be swallowed and carried back down to the intestine. If humans come into contact with larvae of the dog hookworm or the cat hookworm, or of certain other hookworms that do not infect humans, the larvae may penetrate the skin. Sometimes, the larvae are unable to complete their migratory cycle in humans. Instead, the larvae migrate just below the skin producing snake-like markings.

Prevention

- Control of this parasite should be directed against reducing the level of environmental contamination.
- Treatment of heavily infected individuals is one way to reduce the source of contamination.
- improve access to sanitation, e.g. toilets
- Wearing shoes when you walk outdoors, especially in areas that might have feces in the soil
- Drinking safe water
- Properly cleaning and cooking food
- Practicing proper handwashing

Treatment

The drug of choice for the treatment of hookworm disease ismebendazole which is effective against both species, and in addition, will remove the intestinal worm Ascaris also, if present. The drug is very efficient, requiring only a single dose and is inexpensive.

An infection of *N. americanus* parasites can be treated by using benzimidazoles,albendazole and mebendazole.

Application activity 8.1

1. List the ways in which cholera is transmitted from person to person.
2. Explain why there is such a high risk of cholera following natural disasters such as earthquakes, hurricanes, typhoons and floods.
3. Explain why there is a high death rate from TB in countries with a high proportion of the population who are HIV-positive.
4. TB is an opportunistic infection. Why?
5. Describe how malaria is transmitted.
6. Describe the biological factors that make malaria a difficult disease to control.
7. Describe the precautions that people can take to avoid catching malaria.
8. Which of the following diseases is transmitted by an insect vector?
   a) Cholera
   b) HIV/AIDS
   c) Malaria
   d) TB
9. Tuberculosis is a serious infectious disease.
   a) Name the causative agent of tuberculosis.
   b) Explain how tuberculosis is transmitted.
   c) State the regions of the world with the highest number of cases of tuberculosis.
   d) Suggest reasons for the high number of tuberculosis in some parts of the world.
10. The table shows the number of cases of cholera and deaths from the disease for the five countries with the greatest outbreaks as reported to the WHO in 2010.

Country	Region	Total number of cases	Number of deaths	Case fatality rate %
Haiti	Caribbean	179 379	3 990	
Cameroon	West Africa	10 759	657	6.10
Nigeria		44 455	1 712	3.85
Democratic Republic of Congo		13 884	182	1.31
Papua New Guinea	Australasia	8 997	95	1.06
Total	All regions of the world	317 534	7 543	2.38

a). Describe how cholera is transmitted.

b). With reference to the table:

i. Calculate the case fatality rate for Haiti in 2010.

ii. Suggest why the case of fatality rate varies between countries.

iii. Explain why it is important that the WHO collects data on outbreaks of cholera.

c). Explain why there are no epidemics of cholera in highly economically developed countries such as Australia, Malaysia and USA.

8.2. Hygiene practices and their importance

Activity 8.2

A student had an exam and then he waked up early in the morning without practicing the morning hygiene and went to school. When he arrived at school, the teacher sent him back to do the morning hygiene. Discuss the possible hygiene practices that the student was supposed to do.

It is important to understand family health issues, because poor personal and environmental hygiene may cause hygiene related diseases. It is therefore important to acquire knowledge and skills on personal and environmental hygiene in order to avoid such diseases.

8.2.1. Personal hygiene

Personal hygiene and sanitation is very important because:

- It insures proper growth and development of children.
- It helps to prevent diseases especially hygiene related diseases.

- It prevents bad smell it helps to keep the environment clean, tidy and beautiful.
- It makes the environment appealing and attractive.

The personal hygiene may be maintained through the hygiene practices below:

- Washing the body regularly with clean water and soap.
- Wearing clean clothes.
- Living in clean environment with adequate fresh air.
- Eating adequate balanced diet. Young children should be fed between 5 to 6 times per day. Their diet should be rich in proteins.
- Having regular exercises.

To ensure personal hygiene, the important aspects below are considered:

Care of body

To prevent skin diseases and bad smell one needs:

- Washing the body regularly with clean water and soap.
- Avoid sharing clothes, socks and towels.
- Wearing clean under wears.
- Stay in clean environment.
- Wear clean clothes.

Care of the feet

- Washing the feet regularly with water and soap.
- Keeping the feet dry to avoid fungi or foot rot and bad smell.
- Wearing clean socks.
- Avoid sharing socks.
- Keep nails short and clean.
- Airing your feet daily.
- Apply some oil like Vaseline to keep the feet smooth.

Care of hands

- Wash hands regularly with clean water and soap.
- Keep nails short and clean.
- Wash hand before touching and preparing the food
- Wash hand before eating.

- Wash hand after visiting the toilet.
- Apply some oil like Vaseline to keep the hands smooth.

Care of hair

- Wash your head regularly with clean water and soap.
- Keep your hair well combed.
- Oil your hair to prevent dandruff.

Care of teeth

- Brush your teeth daily with a tooth brush, clean water and tooth paste.
- Avoid eating too much sweets, biscuits and sugary foods, as they can cause tooth decay and dental cavities.
- Eating foods that help build strong teeth like milk, eggs, vegetables, fruits and fish.
- To visit dentist regularly for checkups.
- To avoid putting sharp dirty objects in the mouth.
- Pregnant mothers should feed on dies rich in Calcium which is used to make teeth.

Care for eyes, ears and nose

These sensory organs are very important. We must care for them well. Injury may result in permanent disability such as: blindness, deafness and lack ability to detect smells.

Care of the eyes

- Wash eyes with clean water.
- Protect your eyes from foreign objects and dusts.
- Protect your eyes from too much or too low light. You always read under adequate light.

Care for ears

- Wash ears with clean water and soap.
- Never insert hard objects in the ear, this may damage your ear and cause deafness.
- Avoid very loud noise.
- Clean ears with cotton around a matchstick or use ear buds and use it gently.

Care for the nose

- Never insert hard objects in the nose.
- Keep the nose clean by blowing it regularly.
- Avoid being hit in the nose.

8.2.2. Environmental hygiene

Our surrounding environment including water, air and soil should be protected and kept clean. There should be different hygiene practices to prevent soil pollution, water pollution and air pollution.

Application activity 8.2

1. What are possible hygiene activities are required to care for;
    a) Eyes
    b) Nose
    c) Feet
2. Explain general human hygiene practices.

8.3. Human immune system: Immunity, structure and role of antibodies

Activity 8.3

The human body is always exposed to environment full of pathogens and it remains healthy. How is this possible?

Immunity is a process by which the body of a living organism defends itself against pathogens. The immune system is a protective system that is made of a series of defenses that fight against diseases by: recognizing, attacking, destroying and remembering each type of pathogen that enters the body. It does this by producing specialized cells which inactivate pathogens.

The immune system includes two general categories of defense mechanisms against infections:*nonspecific defense and specific defense.*

8.3.1. Nonspecific defenses

They include physical barriers (skin) and chemical barriers (mucus, tears and sweat) which prevent foreign pathogens from entering the body. Nonspecific defenses occur into two lines of defense:

a) First line of defense: Its function is to keep pathogen out of the body. This role is carried out by: the skin, sweat, tears and mucus. The most important nonspecific defense is the skin. The skin epidermis consists of layers of dead cells called keratinocytes (which have been keratinized: their cytoplasm is replaced by a protein called keratin). The keratinized layers of dead cells act as an effective barrier to pathogens and cannot normally be penetrated by bacteria or viruses.

Likewise, the mucous that line the digestive,respiratory, nose and genitals block the entry of harmful microbes. Beyond theirrole as a physical barrier, the skin and mucous membranes counter pathogens with chemicaldefenses. For example, secretions from sebaceous and sweat glands give the skin an acidic PH ranging from3 to 5, which kills bacteria. Many secretions of the body including: tears, mucus and saliva contain lysozyme, ananti-microbial enzyme that breaks the walls of bacteria. Also, stomach acid and digestive enzymes destroy many pathogens which are swallowed with the food.

b) The second line of defense: Microbes which have managed to penetrate the first line of defense face the second line of defense which depends mainly on phagocytosis. Phagocyte function is intimately associated with an effective inflammatory response and also with certain antimicrobial proteins.

i. Phagocytosis:

When pathogens are detected in the body, the immune system produces millions of white blood cells which fight the infection. Many of this white blood cells are phagocytes, which engulf and destroy bacteria.

The word Phagocyte comes from Greek where: *phag* means: to eat and *kutos* means cell. Thus phagocytes are cells that engulf and destroy pathogens. There are two main types of phagocytes:

- Neutrophils are the common phagocytes produced in bone marrow and carried in blood to be released in tissue fluid. They will be released in large number as a result of infection.

  The phagocytic cells called neutrophils constitute about 60 % to 70% of all white blood cells (leucocytes). The neutrophils enter the infected tissue, engulfing and destroying microbes there.

- Macrophages are large cells manufactured in bone morrow and travel in blood as monocytes. They tend to be settle in lymph nodes, liver and in spleen.

How phagocytes work: When pathogens enter the body, phagocytes move towards them, they engulf and destroy them. Once the phagocyte is bound to the pathogen, it will envelop it by folding its membrane inwards.

The pathogen is trapped inside the vacuole called phagosome. Lysosomes fuse with the phagosome and release enzymes called lysins into it. These enzymes digest pathogens and the end products are harmless nutrients that can be absorbed into the cytoplasm.

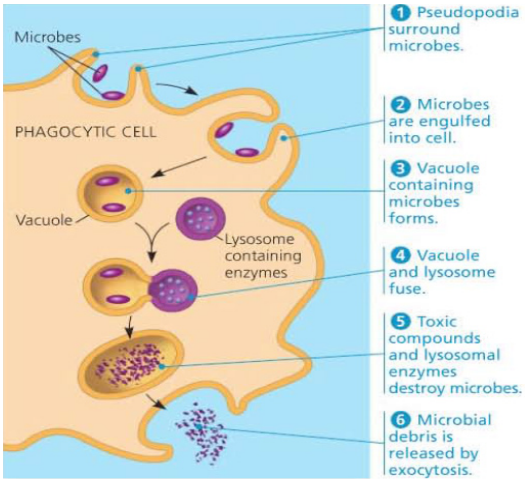


Figure 8.27: Phagocytosis of microbes

ii. The inflammatory response

One result of phagocytosis is inflammation or inflammatory response that is characterized by: redness, hotness, swollen and painful at the site of local infection. An inflammatory response is a nonspecific defense reaction to tissue damage caused by injury or infection.

The tissues injured by bacteria or physical damage (like a cut) release chemicals such as histamine and prostaglandins. These chemicals work as signals which induce an increased heart rate and increased blood capillary permeability (so that white blood cells get to the site of infection faster).

Infected cells also release chemicals which attract phagocytes. When phagocytes and macrophages arrive at the site of injury, they engulf and destroy pathogens and the tissues heal.

The inflamed tissues contains many bacteria and phagocytes, many of which die and form the pus. The increased body temperature slows or stops the growth of pathogens. The fever and increased number of leucocytes are two indications that the body is working hard at fighting infections.

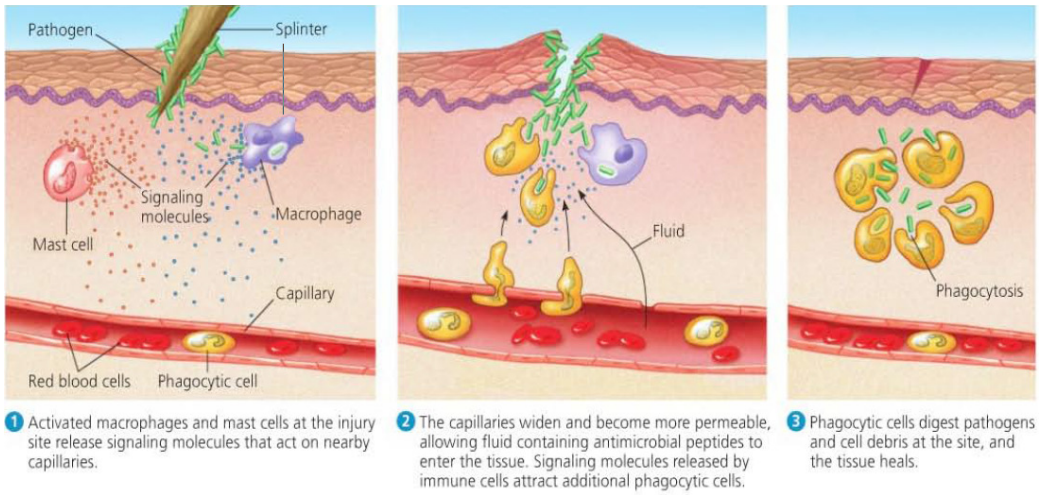


Figure 8.28. Inflammatory response at the injury site

iii. Anti-microbial proteins

Two groups of proteins: Interferons and the complement system provide short-term, nonspecific defense to viral and bacterial infections. Interferons are proteins secreted by virus-infected cells which inhibit the synthesis of viral proteins in affected cells and help to block viral replication. They diffuse to neighboring cells and stimulate them to produce antiviral proteins, which prevent viruses from multiplying within them.

Interferons also activate natural killer cells and macrophages, which destroy infected host cells before they release more viruses. Interferons can also promote destruction of cancer cells. The complement system consists of about 30 proteins in blood plasma that function together to fight infections. These proteins circulate in an inactive state and are activated by substances on the surface of many microbes. Activation results in a cascade of biochemical reactions leading to lysis (bursting) of invading cells. The complement system also functions in inflammation.

8.3.2. Specific defense/immunity: third line of defense

If pathogens are able to get past the body's non-specific defenses, the immune system reacts with a series of specific defenses that attack the particular disease-causing agent. These defenses are called immune response. A substance that triggers (causes) this response is known as an antigen. Viruses, bacteria and other pathogens may serve as antigens. The cells of the immune system that recognize specific antigens are two types of lymphocytes:

a) B lymphocytes (B cells): Which originate in the bone marrow and migrate to the lymph nodes where they proliferate into plasma cells which produce antibodies. B cells provide immunity against antigens and pathogens in the body fluid. B lymphocytes give rise to the humoral immunity or antibody-mediated immunity.

b) T lymphocytes (T cells): Originate in bone marrow and migrate into the thymus. They don't produce antibodies, but they themselves provide defense against antigens and pathogens inside living cells. T lymphocytes give rise to the cell-mediated immunity.

Humoral immunity

When pathogens invade the body, their antigens are recognized by few B cells of the body. These B cells grow and divide rapidly, producing large number of plasma cells and memory cells. Plasma cells produce and release antibodies. Antibodies are proteins that recognize and bind to antigens. They are carried in the bloodstream to attack the pathogen at the site of infection. As antibodies overcome pathogens, the plasma cells die out and stop producing antibodies.

Once the body has been exposed to a pathogen, millions of memory B cells remain capable of producing antibodies specific to that pathogen. These memory B cells, greatly reduce the chance that the disease could develop a second time. If the same antigen enters the body a second time, a second response occurs. The memory B cells divide rapidly, forming new plasma cells. Plasma cells produce the specific antibodies to destroy the pathogen as faster as possible.

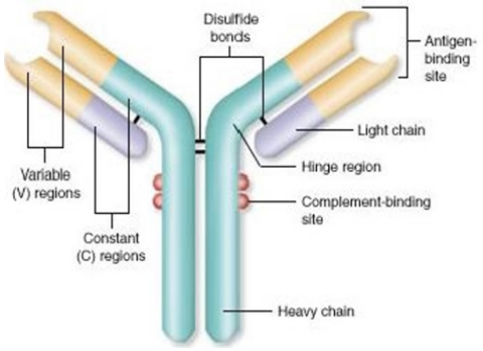


Figure 8.29: Structure of an antibody

An antibody is shaped like the letter Y, and has two identical antigen-binding sites. The shape of the binding site allows the antibody to recognize a specific antigen with a complementary shape. The different shapes give antibodies the ability to recognize a large variety of antigens. A healthy adult cane produce about 100 million different types of antibodies.

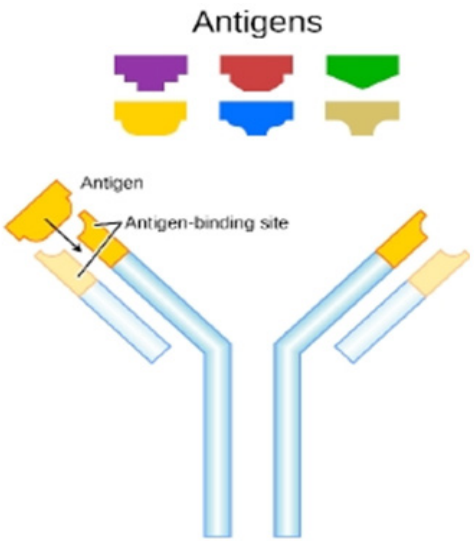


Figure 8.30: Antigen-biding site of an antibody and corresponding antigen

Cell-mediated immunity

It is the body's primary defense against its own cells when they have become cancerous or infected by viruses. The cell-mediated immunity is also important in fighting infection caused by fungi and protists. When viruses and other pathogens get inside living cells, antibodies alone cannot destroy them. During cell-mediated immunity, T cells divided and differentiated into: killer T cells (cytotoxic T cells), helper T cells, suppressor T cells and memory T cells.

Killer T cells track down and destroy bacteria, fungi, protozoa and foreign tissues containing antigens. Helper T cells produce memory T cells. The memory T cells, like memory B cells, will cause a secondary response, if the same antigen enters the body again. As pathogens are brought under control, suppressor T cells release substances that shut down the killer T cells.

Types of immunity

Immunity can be active or passive. In active immunity, the body makes its own antibodies after exposure to an antigen, whereas in passive immunity, the body acquires antibodies produced by other person or other animals. Either active immunity or passive immunity can be naturally or artificially. Thus, there are four types of immunity:

a. Natural active immunity

In natural active immunity, the body makes its own antibodies, in response to an antigen.

b. Artificial active immunity

In artificial active immunity, the body makes its own antibodies, as a result of vaccination against diseases. Vaccination (immunization) is the production of immunity as a result of injection of weakened pathogens which cannot cause diseases to the body, but which can induce production of specific antibodies to those pathogens.

Figure 8.31: The child being immunized

c. Natural passive immunity

Natural passive immunity is a temporary immunity resulting from acquiring antibodies produced by another individual. The only natural passive immunity is when a foetus acquires antibodies from the mother, either through the placenta before birth, or through the breast feeding after birth. This immunity protects the child against most infectious diseases for the first few months of its life, or longer, if the infant is breast-fed.

d. Artificial passive immunity

It is a temporary immunity that results from the injection of the serum containing antibodies from another organism. In this passive form of immunity, a recipient is not induced to produce his or her own antibodies, but is supplied with them, from an outside source. For example: the horse serum well prepared, can then be used to protect humans against diseases.

The trouble with the protection induced by this immunity is that it is short-lived. Nowadays, the passive immunity is normally used only in emergencies, when is too rate for active immunization to work quickly enough.

Application activity 8.3

I. Multiple Choice Questions

1. Humoral immunity is carried out by the
   a). B lymphocytes
   b). T lymphocytes
   c). Phagocytes
   d). T lymphocytes, phagocytes and NK cells
2. Plasma cells represent
   a). B lymphocytes which are actively secreting antibodies.
   b). T lymphocytes which are actively secreting cytokines
   c). Monocytes which have entered tissues
   d). CTL which are secreting perforins.
3. Live attenuated vaccine types have the disadvantage over the inactivated vaccine types as
   a). They require booster shots
   b). They do not confer life-long immunity
   c). They may mutate to virulent form
   d). They do not stimulate the immune system strongly.
4. Immune response is
   a). The defense mechanism of our body.
   b). Reaction of the cells and fluids of the body.
   c). A substance that destroys or inhibits the growth of other microorganisms
   d). None of the above

II. Long Answer Type Questions

1. With an illustrative diagram, state the origin and describe the mode of action of phagocytes.

8.4. Common addictive substances and their effects

Activity 8.4

Carefully, analyze the substances 1-7 in figure below and answer to the following questions

1. Among those substances which one you use in daily life?
2. Which do you think are dangerous to life?
3. Suggest one effect of each of the dangerous substances.

Tobacco, Alcohol and drugs addiction is the physical and psychological need to continue using these substance, despite its harmful or dangerous effects.The signs and symptoms of drug addiction drug vary according to the individual and the substances he or she uses.

8.4.1. Tobacco smoking and its effects

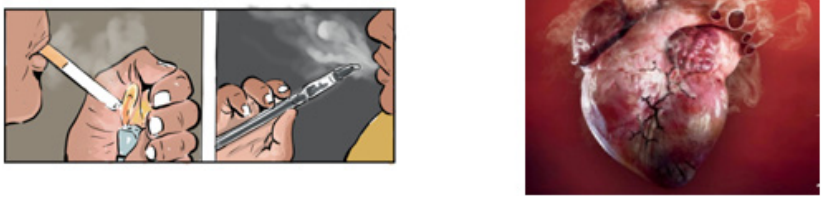


Figure 8.32: Tobacco smoking and heart attack

Smoking harms nearly every organ in your body. Among heavy smokers, two in three will die from a disease caused by smoking. Tobacco smoke is made up of thousands of chemicals and many of them are very harmful. Around 70 of them cause cancer. Cigarette smoke contains over 4000 different chemicals. Many of them are harmful. The most harmful substances in cigarette smoke include: nicotine, carbon monoxide and tar.

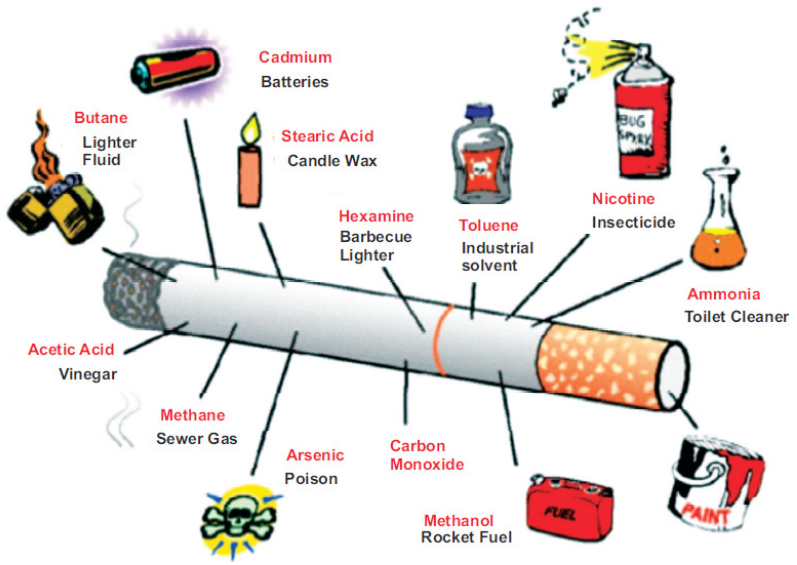


Figure 8.33: Cigarette ingredients.

8.4.2. Effects of Tar and carcinogens in tobacco smoke on the gas exchange system

Tar is a combination of severe chemicals lying on the lining of the airways and alveoli. This increases the diffusion distance for Oxygen entering the blood and CO_2 leaving the blood. The lumen of airways gets smaller and this restricts the flow of air to the alveoli. The tar paralyses or destroys the cilia on the surface the airways, so they are unable to move the mucus away. Bacteria and viruses trapped in mucus are not removed. They can multiply in the mucus and may block the bronchioles and affect breathing process. A combination of bacteria, viruses and mucus in airways and alveoli causes the lungs to be more susceptible to infections. Smokers are more likely to be attacked by influenza and pneumonia.

Tar contains also the carcinogen compounds which cause cancer. When the tar lies on the surfaces of airways, carcinogens enter the cells of lung tissues. They enter the nucleus of these cells and have a direct effects on their genetic material. Any change on genetic material is called mutation, if mutations affect the genes that control cell division, then uncontrolled cell division take place. This is cancer. Lung cancer often takes 20-30 years to develop, and a cancer may grow for many years before it is discovered.

Figure 8.34: Healthy (Left) and tar coated lung (Right)

8.4.3. Effects of Nicotine and carbon monoxide on the cardiovascular system

a). Carbon monoxide (CO)

It is fatal in large amount. This poisonous gas is also found in car exhaust fumes. CO enters the red blood cells and combines with haemoglobin more rapidly than Oxygen, and form the carboxyhaemoglobin.

This reduces the oxygen-carrying capacity of blood, preventing your lungs, heart, and other organs from getting oxygen they need to function properly. CO can also damage the lining of arteries, and may rise the heart beat rate.

b). Nicotine

It is a poisonous alkaloid drug that is addictive. 60 mg of nicotine placed on the tongue would kill an individual within minutes. It is absorbed by the body very rapidly, reaching the brain in less than 30 seconds.

It is a highly toxic chemical and its manufacture, use and sale is controlled under the State Poisons Acts, except where it occurs in tobacco.This exception of tobacco is for political reasons, not because nicotine is deemed 'safe' in cigarettes. Nicotine, once inhaled, affects the body very quickly. It causes changes to the structure and the working of the brain, which lead to nicotine addiction. Nicotine also *raises heart rate, blood pressure*, releases hormones (adrenaline) affecting the central nervous system, and *constricts small blood vessels (arterioles)* under the skin. In the long term, nicotine may be a factor in *causing coronary disease*. It is believed to be involved in the development of gastrointestinal disorders and problems during pregnancy, and is linked with the development of cancers.

8.4.4. Contribution of tobacco smoking to atherosclerosis and coronary heart disease

Atherosclerosis is the deposition of fatty substances in the walls of the arteries. The CO in the cigarette smoke can damage the inner lining of arteries. This encourages the deposit of fatty substances like cholesterol on the walls of arteries. This reduces the lumen of arteries, which reduces blood flow, and causes high blood pressure.

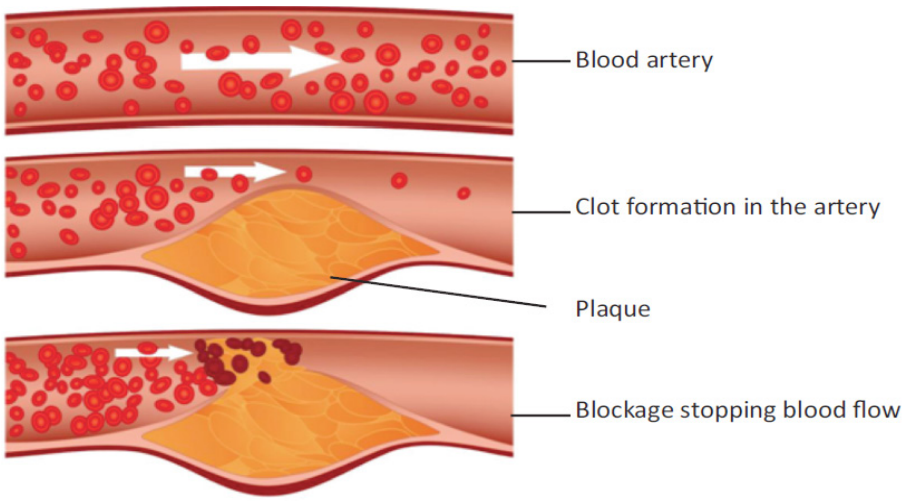


Figure 8.35: The process of blocking arteries by plaque in coronary heart disease

Coronary heart disease (CHD) is a disease of the heart caused by malfunction of the coronary artery. Coronary arteries are the arteries which carry blood to the heart.

The nicotine and carbon monoxide found in cigarette enter the lungs and diffuse into blood in blood capillaries, that conducting them into coronary arteries. This causes the atherosclerosis in coronary arteries, which reduces the blood flow in the heart muscles which receive less oxygen for respiration. This can lead to coronary heart diseases, which take three forms: In **circulatory system, they cause changes which lead to cardiovascular diseases like:**

a). **Angina:** A severe pain in the chest, which may extend down the left arm or up into the neck.

b). **Heart attack or myocardial infections:** Caused by a clot in the coronary arteries, blocking the flow of blood to the heart muscles.

c). **Heart failure:** When the heart cannot sustain its pumping action. This can be caused by the blockage of the major coronary arteries.

8.4.5. Symptoms of lung cancer and chronic obstructive pulmonary diseases (COPD)

a). Chronic obstructive pulmonary disease: Is a combination of diseases that include chronic bronchitis, emphysema and asthma. The symptoms of COPD are a combination of symptoms of chronic bronchitis, and of emphysema.

i. Chronic bronchitis: Is the inflammation of the airways. This is accompanied by damage to the cilia and the over production of mucus. The symptoms are: irritation in lungs, continual coughing, and coughing up mucus that is often filled with bacteria and white blood cells.

ii. Emphysema: Is the loss of elasticity in alveoli, which causes the alveoli to bust. The symptoms are: short breath, hard exhalation, the blood is less well oxygenated and fatigue occurs.

b). Lung cancer: The symptoms of lung cancer can be recognized by: continual coughing, shortness of breath, pain in the chest and the blood coughed up in the sputum is often the first sign of the lung cancer.

8.4.6. Evidences linking cigarette smoking to diseases and early death

a. Links of cigarette smoking to early death

- Regular smokers are three times more likely to die prematurely than a non-smoker.
- 50% of regular smokers are likely to die of a smoking-related disease
- The more cigarettes a person smokes per day, the more he is likely to die prematurely.

b. Links of cigarette smoking to lung cancer

- A smoker is 18 times more likely than a non-smoker to develop lung cancer.
- 25% of smokers die of lung cancer.
- A heavy smoker (more than 25 cigarettes a day) is 25 times more likely than a non-smoker to die of lung cancer
- The chances of developing lung cancer reduce as soon as a person stops smoking.

c. Links to other lung diseases

- COPD is rare in non-smokers
- 98% of people with emphysema are smokers.
- 20% of smokers have emphysema

8.4.7. Alcohol

Alcohol is any organic compound in which the hydroxyl functional group (-OH) is bound to a carbon. The term *alcohol* originally referred to the primary *alcohol* ethanol *(ethyl alcohol)* which is used as a drug and is the main *alcohol* present in *alcoholic* beverages.

Alcohol is produced by adding yeast on a liquid that contains sugar in absence of oxygen. This yeast ferment sugar to produce energy and alcohol.

Most popular alcohol containing drinks include fermented drinks such as beer and wine, other drinks such as vodka, whiskey, scotch, and gin are made by distillation.

Figure 8.36: Different fermented and distilled alcoholic drinks

Effects of alcohol

The effects of alcohol vary from person to person. While some people may be able to limit their drinking, others have a difficult time controlling their alcohol consumption. Effects of alcohol include: slurred speech, vision impairment, lack of coordination, extreme shifts in mood, memory lapses and slowed breathing.

The most immediate alcohol effects are on the nervous system. The small dose of alcohol slows down the rate of nervous system functions. So, alcohol is a depressant.

The pregnant women who drink regularly run the risk of *"fetal alcohol syndrome"* or damage to the developing babies due to the effect of alcohol.

People who have become addicted to alcohol suffer from a disease called alcoholism. If someone cannot work effectively without alcohol, that indicates an alcohol abuse problem.

Taking alcohol in excess leads to damage of neurons in the brain, cells in the liver. The damage of liver cells causes the live to become less able to deal with high amount of alcohol, then the formation of the scar tissue known as cirrhosis in liver occurs. Finally the drinker may die from the chronic liver failure.

Basing on the effect of alcohol on nervous system, driving can lead to sudden accident that may kill people.

Figure 8.37: Accident as bad effect of alcohol

8.4.8. Drugs

Drug is any substance that cause a change in the body. Drugs affect the body in different ways, because some are very powerful and dangerous and their possession is not allowed. Other drugs like penicillin and codeine are drugs that can be used under the supervision of doctor.

All drugs (legal and illegal ones) have the capacity of harming life if they are abused. In this section, some of the most commonly abused drugs and the way they affect the body will be considered.

Drugs affect the particular system of the body like digestive and circulatory systems.

The following are the most powerful drugs that can also affect the nervous system and their negative side effects.

Fugure 8.38: Marijuana plantation and hallucinogens (LSD and PCP)

Marijuana (Also called Hashish or Hash)

Negative effects of using marijuana

1. **Addiction**: Marijuana is physically addictive and psychologically addictive, especially as it concerns younger.
2. **Memory Loss**: People who always use marijuana developed a poorer verbal memory in middle age than people who didn't smoke.
3. **Social Anxiety Disorders:** Regular use of marijuana can lead to mental health issues such depression, anxiety and even schizophrenia.
4. **Paranoia**: Researches indicated that the use of marijuana can lead users to feel a sense of paranoia as a result of the changes in their sensory perception
5. **Heart Damage**: Marijuana can also significantly raise a person's heart rate for up to three hours. Even, people who use marijuana are more likely to have a stroke at some point in their lives than people who didn't use it.
6. Other effect are: Lung problems, low testosterone, appetite irregularities, risk of greater potency, decrease in motor Responses, poor decisions.

Hallucinogens

Hallucinogens are drugs that cause hallucinations. Users see images, hear sounds and feel sensations that seem very real but do not exist. Some hallucinogens also produce sudden and unpredictable changes in the mood of those who use them.

There are many types of hallucinogens but the two common types such as LSD(Lysergic acid diethylamide) and PCD(Phencyclidine) are the ones which are used.

LSD is a powerful hallucinogen that interferes with the normal transmission of nerve impulses in the brain. Its effects vary from person to person and all people who use LSD regularly have a bad trip. Some of LSD users have lost touch with reality after only one single dose.

PCP produces fillings of strength and great power. High dose of PCP heart attacks. The users of PCP often become extremely violent and are danger to themselves and others.

Stimulants

These are drugs that speed up the actions of nervous system. The most powerful stimulants are Amphetamines, these chemically resemble natural neurotransmitters found in the body (Compounds that pass nerve impulse from one neuron to another). Once amphetamine enters the blood stream, it floods the body and behaves like neurotransmitters. This causes the nervous system to increase its activity producing a feeling of strength and energy in the user body. Later the user suffers from fatigue and depression as the nervous system becomes unable to handle the overstimulation produced by amphetamines after their dose wears off.

Long-term use of amphetamines causes hallucinations, circulatory problems. psychological difficulties.

Depressants

These are drugs that reduce the rate of nervous system activity. These drugs are also known as downers. People become dependent on them as long as they use them. Example of downers is barbiturates. Use of the barbiturates with alcohol is fatal, because the nervous system becomes so depressed and breath stops.

The barbiturates abuse is danger because it brings about serious medical problems that must be immediately treated once the user stops supplying the drugs to the body.

Cocaine

This drugis compound extracted from leaves of coca plant. Users supply it to the body by smoking it, sniffing it or inject it into the bloodstream.

Figure 8.39: People sniffing, smoking and injecting cocaine

Cocaine stimulate the release of neurotransmitter called dopamine which is normally released by the brain if person is satisfied during lunch or dinner. So, cocaine creates feeling of pleasure and satisfaction.

Effects of using cocaine

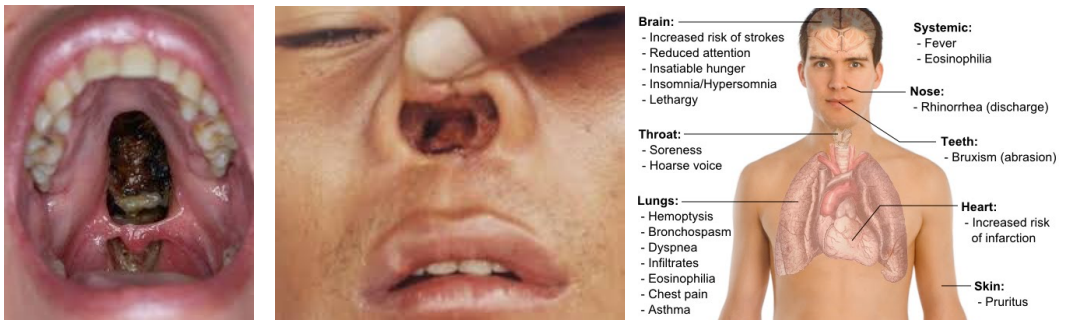


Figure 8.40: Some negative effects of using cocaine

- Heavy use of cocaine leads to lung damage similar to emphysema. Overdose stops breathing.
- Use of cocaine increases the heart rate and blood pressure leading to risk of irregular heartbeat or even heart attack.
- Cocaine causes extreme mood changes, irritability. Long-term use can lead to psychological problems.
- Cocaine causes the users to experience the sensation of bugs crawling over the skin.
- Cocaine usage reduces a desire for food, leading to body weight loss.

Opiates

This is a group of drugs produced from the opium poppy. The most common opiates are opium produced directly from the opium poppy. Other examples of opiates are morphine and heroin which are forms of opium. Apart from heroin, other opiates can be used under supervision of doctor to reduce severe pain as they are used as pain killers. If opiates are taken in large dose, they can result to death.

In addition, opiates such as heroin, are example of drugs that cause a strong physical dependence. The regular user becomes addicted and the nervous system becomes also dependent on the supply of drugs. Any attempt of withdrawal or stopping the use of drug will cause severe pain, nausea, chills and fever.

Application activity 8.4

I. Choose whether the following statements are True (T) or False (F)

1. Smoke from cigarettes can make non-smokers sick.
2. Smoking can affect a person's ability to smell and taste food.
3. Secondhand smoke kills about 3,000 non-smokers each year from lung disease.
4. It takes about ten seconds for nicotine absorbed into the bloodstream to reach the brain.
5. Smoking is a difficult habit to quit.
6. Nicotine, the chemical found in cigarettes, is an addictive drug.
7. A smoker is twice as likely to have a heart attack as a non-smoker is.
8. Cigarette brands that are heavily advertised on TV, in magazines, on billboards, and on T-shirts are the brands more teens buy.
9. One out of every ten smokers will die of a smoking-related sickness.
10. More germs get into your lungs when you smoke.

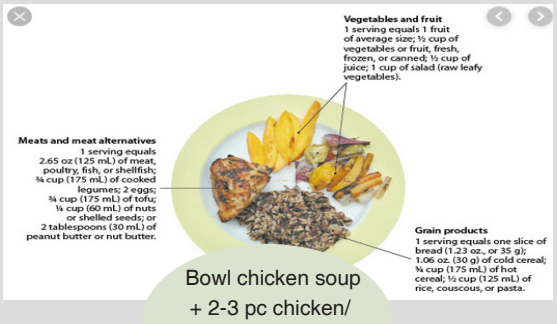
II. Suggest the components of a tobacco smoke

III. Give the effects of alcohol on human body health

8.5. Balanced diet

Activity 8.5

Carefully, discuss and find out the component of well balanced diet



A

Bowl chicken soup + 2-3 pc chicken/ Fish/1 egg + 1 cup vegetable

B

1. Observe the recipes (A & B) above then respond to related questions.
   a) What are main nutrients in A and B?
   b) Compare recipes A and B. Which one is not balanced? Why?
2. What are consequences when B is in your diet all the time?
3. Propose a good healthy family eating plan.

Eating a balanced diet means eat at least 5 portions of a variety of fruit and vegetables every day (see 5 A Day), base meals on higher fibre starchy foods like potatoes, bread, rice or pasta, have some dairy or dairy alternatives (such as soya drinks) eat some beans, pulses, fish, eggs, meat and other protein.

Food variety means eating a wide variety of foods from each of the five food groups, in the amounts recommended. Eating many different foods helps to maintain a healthy and interesting diet which provides a range of different nutrients to the body. Eating a variety of foods promotes good health and can help reduce the risk of disease.

The nutritional requirement is influenced by age,sex, growth,pregnancy and breastfeeding, illness, psychological and emotional stress, activity level and other factors like smoking and drinking. Biological factors include age, gender, growth, disease states, and genetic makeup. Among the no biological factors, socio-economic status is the most important. Poverty is one of the major socio-economic causes of variation in nutrient intake, and it also impacts nutrient requirements.

Example:

Aging is linked to a variety of changes in the body, including muscle loss, thinner skin and less stomach acid. ... Low stomach acid can affect the absorption of nutrients, such as vitamin B12, calcium, iron and magnesium. Although the recommended breakdown of carbohydrate, protein, and fat are the same for both genders, because men generally need more calories, they also require higher total intake of each of the macronutrients. Women need fewer calories than men, but in many cases, they have higher vitamin and mineral needs.

Reference Intakes

Nutritional needs vary depending on sex, size, age and activity levels so use this chart as a general guide only. The chart shows the Reference Intakes (RI) or daily amounts recommended for an average, moderately active adult to achieve a healthy, balanced diet for maintaining rather than losing or

gaining weight. The RIs for fat, saturates, sugars and salt are all maximum amounts, while those for carbs and protein are figures you should aim to meet each day. There is no RI for fibre although health experts suggest we have 30g a day.

Table 8.6: Food nutrients and reference intakes

	Reference intakes (RI)	
	Men	**Women**
Energy (kcal)	2500	2000
Protein (g)	55	50
Carbohydrates (g)	300	260
Sugar (g)	120	90
Fat (g)	95	70
Saturates (g)	30	20

Table 8.7: Food groups and their percentages in the diet

Food group	Importance	Examples	% of diet
Carbohydrates	Provide energy for growth and development.	Whole grain forms are best because they provide extra fibre. Bread, maize and cassava	33%
Fruits and vegetables	These foods provide vitamins and minerals needed for a healthy immune system. They also provide fibre. You should aim to eat five portions of these foods per day.	Bananas, spinach and tomatoes	33%
Dairy foods	These foods provide fat for energy, protein for muscle and nerve development, calcium for bone development and vitamins for fighting infections.	Milk, yoghurt and sour milk	15%
Non-dairy proteins	These foods provide protein needed for muscle and nerve development, and for tissues to repair after infections.	Meat, eggs, beans	12%

Foods high in sugar or fat	These foods should be limited to prevent excessive weight gain or problems with blood sugar.	Cakes, biscuits, fried foods, chocolate and cooldrinks	7%

A balanced diet is one that contains all nutrients required in health in appropriate proportion. A balanced diet should contain all food groups such as: body building food, energy giving food and protective food in an appropriate amount. A balanced diet helps a person in the following ways:

- Makes you strong.
- Provides better health.
- Makes you more productive.
- Ensures strong immune system.

It's not hard to include foods from the five food groups into appetizers and meals. Some suggestions include:

Figure 8.41: Balanced diet include foods from all food groups

- **Vegetables and legumes:** Raw or cooked vegetables can be used as a snack food or as a part of lunch and dinner. Salad vegetables can be used as a sandwich filling. Vegetable soup can make a healthy lunch. Stir-fries, vegetable patties and vegetable curries make nutritious evening meals. Try raw vegetables like carrot and celery sticks for a snack 'on the run.'
- **Fruit:** This is easy to carry as a snack and can be included in most meals. For example, try a banana with your breakfast cereal, an apple for morning tea and add some berries in your yoghurt for an afternoon snack. Fresh whole fruit is recommended over fruit juice and dried fruit.

Fruit juice contains less fibre than fresh fruit and both fruit juice and dried fruit, and are more concentrated sources of sugar and energy. Dried fruit can also stick to teeth, which can increase the risk of dental caries.

- **Bread, cereals, rice, pasta and noodles:** Add rice, pasta or noodles to serves of protein and vegetables for an all-round meal. There are many varieties of these to try. Where possible, try to use wholegrains in breads and cereals.
- **Lean meat, fish, poultry, eggs, nuts, legumes:** These can all provide protein. It's easy to include a mixture of protein into snacks and meals. Try adding lean meat to your sandwich or have a handful of nuts as a snack. You can also add legumes to soups or stews for an evening meal.
- **Milk, yoghurt and cheese:** Try adding yogurt to breakfast cereal with milk, or using cottage cheese as a sandwich filling. Shavings of parmesan or cheddar can be used to top steamed vegetables or a salad. Use mostly reduced fat products.

Feeding on unbalanced diet for a longtime may lead to malnutrition diseases. Malnutrition means feeding on a meal lacking some food nutrients (deficient diseases), or on a meal with all food nutrients but in unappropriated amount (over eating).

Some deficient diseases include kwashiorkor (caused by the meal lacking proteins), marasmus (caused by the meal lacking overall nutrients), and goitre (caused by the meal lacking iodine). The diseases caused by over eating include obesity, a condition in which excessive fats are deposited in the body. More malnutrition diseases are described in the tables above which describe the functions of minerals and vitamins.

8.6. Basic food service technics

a. Basic table service methods

Activity 8.6

Observe the picture below:

1. Describe the food setting order on pictures above?
2. Analyze the table setting on the picture above? Justify your answer.

They are many different approaches to serving food. An operation should use a service style that is the best to satisfy its family members. The traditional table service provides service for family members who are seated at table. The English service comparable to Rwandan style is a type of service known as “family style service”. In this service the big dish is placed in front of the host along with serving plate and family members serve themselves.

Figure 8.42: Food service technics

b. Principles for meal service

The family style meal service allows participants to eat together and to make food choices based on individual appetites and food preferences. It

promotes mealtime as a learning experience to help participants develop positive attitudes toward nutritious foods, share in group eating situations, and develop good eating habits. Family style meal service can be conducted in a variety of ways. For example, participants may help in preparing for the meal by clearing the table and setting places, sharing conversation during the meal, and cleaning up after the meal.

Family style meal service operates as follows:

- All required meal components are placed on the table at the same time.
- Participants may serve themselves from serving dishes that are on the table
- Adults supervising the meal help those participants who are not able to serve themselves.
- Participants can make choices selecting foods and in the size of the serving.
- A supervising adult actively encourage family members to serve themselves and offers the food item again later in the meal if member (s) initially refuse the food or take a very small portion. Adult should model good eating habits while supervising participants at the dining table.

c. Standards of serving food temperatures

The crucial important part of food safety in the home is to keep hot food hot and to keep cold food cold. For safety it is vitally important to keep food out of that danger zone.

The food being served should be kept at specified range and appropriate temperature (to):

Type of food	Hot food	Cold food
To	63°C < to <75°C	0°C < to <5°C

Note:

- The temperature danger zone for bacteria reproduction and growth: 5°C < to < 63°C, food is not suitable for eating.

  Bacteria do not multiply and start to die at 63°C above and do not grow and multiply at 5°C below.

Application activity 8.6

1. List the principle sources of food nutrients.
2. Justify the different functions of food nutrients in the body.

Skills lab 8

1. In your school environment, create a center for workshops about nutrition, hygiene and sanitation, so that different diseases can be eradicated in surrounding population. Help family members in the following:

   a). To install and maintain a backyard garden plantation (green vegetables garden) around their homes.

   b). To practice general body hygiene, food hygiene, clothes hygiene, home hygiene, water sanitation will be the first priority.

   c). To effectively cook foods using appropriate methods like:

   - Boiling (a method involving cooking food completely covered with water and boiling it until fully cooked)
   - Steaming (cooking food using steam from boiling water).
   - Frying (cooking food with hot fat or oil).
   - Stewing (cooking food with low quantities of water).
   - Roasting (cooking food with very hot radiant heat)
   - Baking (cooking food using a direct heat).
   - Grilling (cooking food with direct heat using a grill).

   Elaborate a report on positive changes happening in the community due to your activities. Remember to attach photos and testemonies.

2. Test for malaria

   One of the techniques to test for malaria is the microscopic observation of sample of blood. Nowadays, many medical clinics test for malaria directly by an appropriate Malaria Rapid Test kit.

   Conduct a research to suggest the procedure of testing for malaria.

End unit assessment 8

I. Choose whether the given statements are True (T) or False (F)

1. Innate immunity is present at birth.
2. Breast milk confers protection to newborn by providing IgE type of antibodies.
3. Antibodies can work by promoting phagocytosis of microbial agents.
4. Antibiotics help a patient in mounting an effective immune response.
5. Treating tuberculosis is becoming difficult because Mycobacterium tuberculosis has become resistant to a number of antibiotics.
6. High fever is a generalized allergic reaction caused by release of active mediators from mast cells.
7. Secondary immune response appears much faster because of the presence of memory cells persisting from previous infection.
8. Vaccination against snakebite is an example of passive immunization.
9. Allergies are of two types—innate and adaptive.
10. Beta-lactam antibiotics kill bacteria by blocking synthesis of their cell walls.

III. Long answers type question

1. Analyse and relate the molecular structure of antibodies to their functions.
2. Explain the following:
   a) Phagocytes
   b) Lymphocytes
   c) Immune response
   3. Differentiate between the following: Active and passive immunity
4. State why vaccination programs are able to eradicate smallpox but not measles, TB, malaria or cholera.
5. Name organism which have the following effects on health

	Name
Cause malaria	
Transmits malaria	
Cause influenza	
Causes Ancylostomiasis	
Provide penicillin	
Elephantiasis	
Transmits venereal disease	
Scurvy	
Causes body's bad smell	
Creates feeling of pleasure and satisfaction	

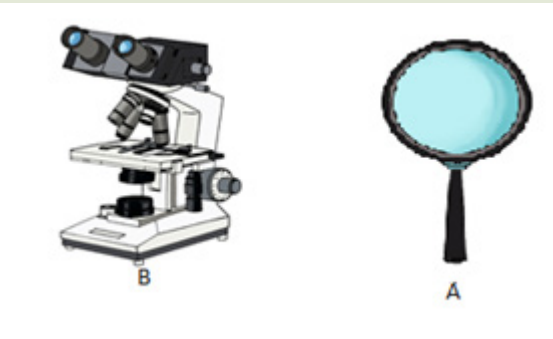
UNIT 9:

MICROSCOPE AND CELL STRUCTURE

Key Unit competence: Describe the structure and function of cells in an organism.

Introductory Activity

Observe the following images and answer the asked questions

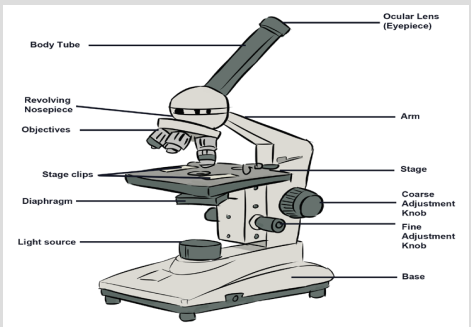


1. Name the illustrations in A) and B) above
2. What are the use of illustrated materials in a)? Where are they used?
3. Predict what you are going to learn in this unit

9.1. Cell theory and microscopes

Activity 9.1

Carefully, analyze the following diagram and answer the related questions



This is one of materials used in Biology.

a). Have you ever seen and manipulate it?

b). Observe the parts of the material, discuss and group them into the following:

i. Suporting parts

ii. Magnifying parts

iii. Adjusting parts

iv. Lighting parts

c). Point out scientific activities that require the use of this material in biology.

9.1.1. Cell theory

Cytology is the study of the structure and function of cells. A Cell is the basic unit of life surrounded by a cell membrane and containing organelles. All living organisms are made of cells and nothing less than a cell can truly be said to be alive. The word cell comes from the Latin *cellula,* meaning a small room or cubicle, and was first used by Robert Hooke in his book *Micrographia*, published in 1665. Hooke was observing slides of cork taken from the bark of an Oak tree under the compound microscope. He decided that the slides were made up of a lot of many small chambers that he called cells that range in size are from 1µm – 1mm.

Living organisms are classified into unicellular organisms made by only one cell, such as bacteria, whereas animals and plants are composed of many millions of cells built into tissues and organs. They are called multicellular organisms. In a multicellular organism, cells divide and thereafter they undergo differentiation or specialisation for specific functions.

In biology, the historic scientific *theory of cells is called cell theory*. The cell theory states that all living organisms are made up of $cells$, and $cells$ are the basic unit of structure function in all living organisms. The main principles of cell theory is that all known living organisms are made up of one or more cells, all cells come from pre-existing cells by division and cells contain the hereditary information that is passed from cell to cell during cell division.

9.1.2. Microscopes: Compound-light microscope and Electron microscopes

Microscopy is the technical field of using microscopes to view objects and areas of objects that cannot be seen with the naked eye. The microscope was created by Zacharias Janssen in the late 16th century.

Prior to the invention of the microscope, the details of objects on slides were limited. Single microscopes were similar to using a magnifying glass such as hand lens. The invention of the light microscope in the 17th century by Antony van Leeuwenhoek made cells visible for the first time and, for hundreds of years afterwards. Electron microscope invented in the 1930s enables the researchers to see very small organelles which cannot be seen by using light microscopes. The purpose of the use of a microscope is to magnify small objects such as cells or to magnify the fine details of a larger object in order to examine minute specimens that cannot be seen by the naked eye.

a. Compound Light Microscope

The optical microscope, often referred to as light microscope is a type of microscope which uses visible light and a system of lenses to magnify images of small samples.

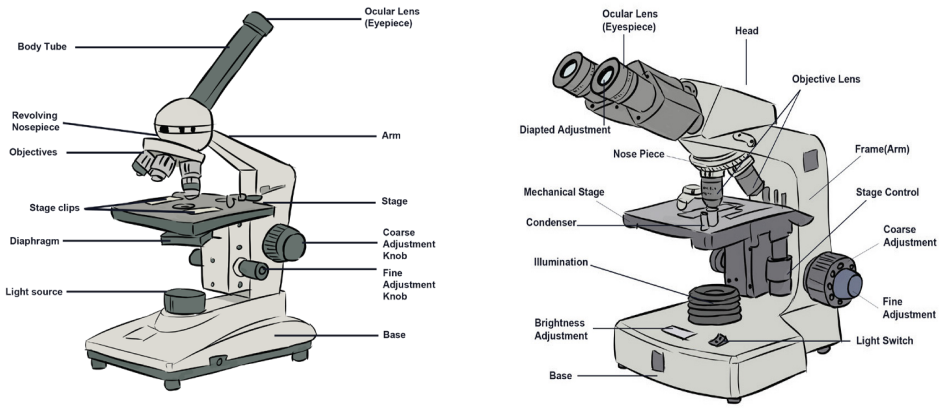


Figure 9.1: Monocular light microscope (1) and Binocular light microscope (2)

The parts of light microscope and their roles:

- Base: Supports and stabilizes the microscope on the table or any other working place.
- Light source: It is made by lamp or mirror which provides light for viewing the slide.
- Stage: Is a platform used to hold the specimen in position during observation.
- Stage clips: Are pliers used to fix and hold tightly the slide on stage.
- Arm: Supports the body tube of microscope
- Body tube: maintains the proper distance between the objective and ocular lenses

- Arm: Used for holding when carrying the microscope and it holds the body tube which bears the lenses.
- Coarse focus adjustment: Moves stage up and down a large amount for coarse focus
- Fine focus adjustment: Moves stage up and down a tiny amount for fine focus
- Objective lenses: Focuses and magnifies light coming through the slide
- Revolving nosepiece: Rotates to allow use of different power objectives
- Slide: Is a transparent pane on which a specimen is placed.
- Eye piece/ocular lens: Magnifies image produced by objective lens
- Condenser: It will gather the light from the illuminator and focus it on the specimen lying on the stage. The function of the condenser is to focus the light rays from the light source onto the specimen.
- Iris diaphragm lever: This allows the amount of light passing through the condenser to be regulated to see the object.

All parts of a microscope work together to magnify a specimen to be observed. Light from the source is focused on the specimen by the condenser lens. It then enters the objective lens, where it is magnified to produce a real image. The real image is magnified again by the ocular lens to produce a virtual image that is seen by the eye.

Advantages and limitations of the light microscope

Magnification: Most light microscopes can magnify a specimens up to a maximum of X1500.

Resolution: It is the degree at which it is possible to distinguish clearly between two objects that are very close together. It is the smallest distance apart that two separate objects can be seen clearly as two objects. The resolution for the:

- Human eye is 100µ.
- Light microscope is 200 nm.
- Electron microscope is 0.20 nm.

The maximum resolution of the light microscope is 200 nm. This means that if two objects are closer together than 200 nm, they will be seen as one object. The light microscopes are used widely in education, laboratory analysis and research. But because they do not have high resolution, they cannot give detailed information about internal cell structure.

Specimens: a wide range of specimens are viewed using the light microscope. These include: unicellular organisms, small sections from large plants and animals, and smear preparation of blood or cheek cells.

Preparation of specimens for the light microscope

A lot of biological materials are not coloured, so you cannot see details. Also, some material distorts when you try to cut it into thin sections. Preparation of slides to overcome these problems involves the following steps:

a) Staining: Coloured stains are chemicals that bind on or in the specimens. This allows the specimens to be seen. Some stains bind to specific cell structures. For example, acetic orcein stains DNA dark red, while gentian violet stains bacterial cell walls.

b) Sectioning: Specimens are embedded in wax, where thin sections are then cut without distorting the structure of the specimen. This is particularly useful for making sections of soft tissue, such as brain. Safety measures might be taken. Make sure that hands are washed with soap and warm water after theexperiment. Use a disinfectant to wipe down all surfaces where bacteria mayhave been deposited for example. Be sure that some substances and animalsmight be harmful to the life.

Measuring cells

Cells and organelles can be measured with a microscope by means of an eyepiece called graticule. This is a transparent scale, usually having 100 divisions (Figure 9.2, A). The eyepiece graticule is placed in the microscope eyepiece so that it can be seen at the same time as the object to be measured (Figure 9.2, B). At this figure (Figure 9.2, B), the cell lies between 40 and 60 on the scale, so that it measures 20 eyepiece units in diameter (60- 40 = 20).

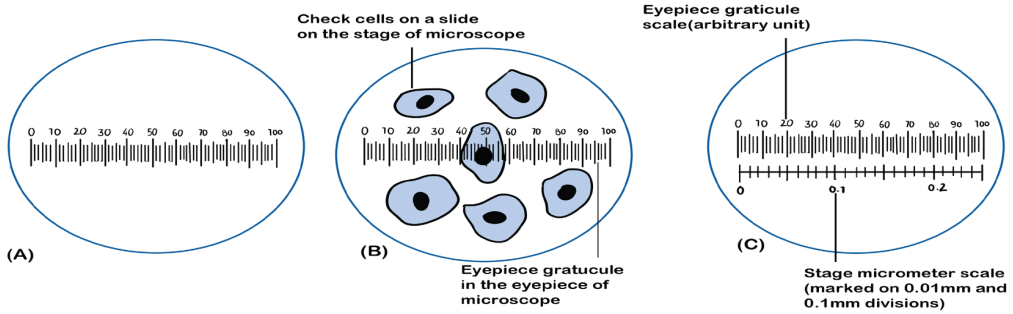


Figure 9.2: Measuring the length of cell using a graticule in a light microscope

Note that it is not possible to know the actual size of the eyepiece units until the eyepiece graticule scale is calibrated. To calibrate the eyepiece

graticule scale, a miniature transparent ruler called a stage micrometre scale is placed on the microscope stage and is brought into focus. This scale may be fixed onto a glass slide or printed on a transparent film. It commonly has subdivisions of 0.1 and 0.01 mm. The images of the two scales can then be superimposed (Figure 9.2, C). If in the eyepiece graticule, 100 units measure 0.25 mm, the value of each eyepiece unit equals $\frac{0.25}{100}$ =0.0025 mm. By converting mm to µm, the value of eyepiece equals $\frac{(0.25 \times 1000)}{100}$= 2.5 µm. The diameter of the cell shown superimposed (Figure 9.2, B) measures 20 eyepiece units. Its actual diameter equals 20 × 2.5 µm = 50 µm. This diameter is greater than that of many human cells because the cell is a flattened epithelial cell.

There is a relationship between the actual size, magnification and image size, where:

Actual size = image size/ magnification.

Magnification = image size/ actual size.

Image size = Magnification x actual size.

b. Electron microscope

An electron microscopes use a beam of accelerated electrons as a source of illumination. Electron beams have a much smaller wave length of 0.004 nm, 100 000 times shorter than light wavelength, and therefore have greater resolving powers and can produce higher effective magnifications than light microscopes. Electron microscopes are used to study the details of internal structures (the ultrastructure's) of cells. Most modern TEMs can distinguish objects as small as 0.2nm. This means that they can produce clear images magnified up to 500,000 times greater than that of the human eye.

There are two types of electron microscopes:

- Transmission electron microscope (TEM).
- Scanning electron microscope (SEM)

i. Transmission electron microscope (TEM)

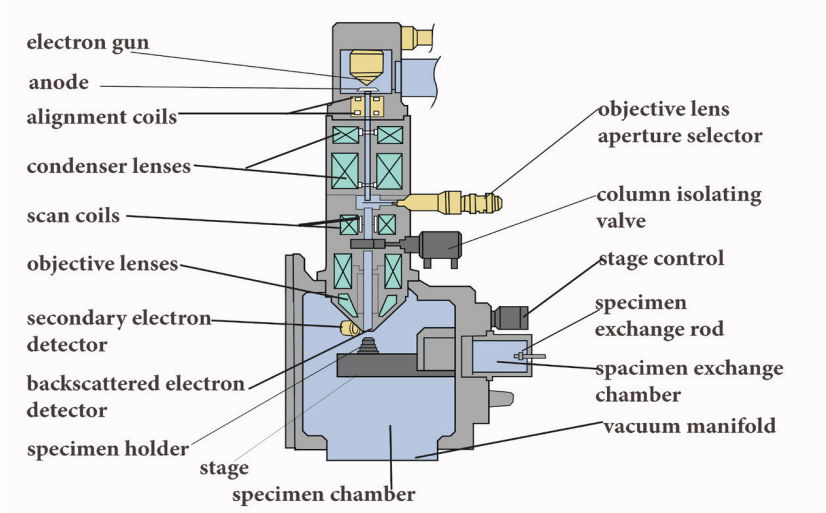


Figure 9.3: Transmission electron microscope

- The electron beam passes through a very thin prepared sample.
- Electrons pass through the denser parts of the sample less easily, so giving some contrast.
- The final image produced is two-dimensional.
- The magnification possible with a TEM is X500 000.

ii. Scanning electron microscope (SEM)

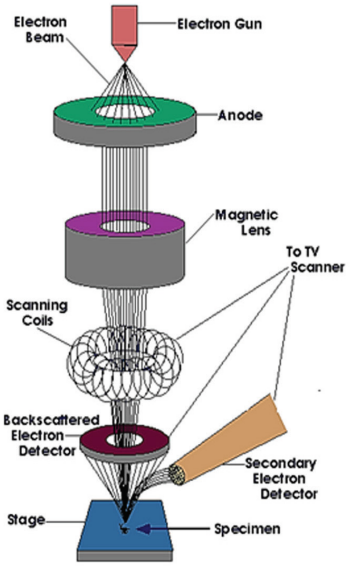


Figure 9.4: Image of Scanning Electron Microscope (SEM)

- The electron beam is directed onto a sample. Electrons do not pass through the specimen.
- They are bounced off the sample.
- The final image produced is a three-dimensional.
- The magnification possible with an SEM is about X100 000.

Advantages and disadvantages of the electron microscope

a. Advantages of electron microscope

Light microscope has a higher resolution and are therefore able of a higher magnification estimated up to 2000 X more than in the light microscope. Electron microscopes therefore allow for the visualization of too small structures, such as cell organelles that would normally be not visible by optical microscopy. The SEM produces a 3D images that can reveal the detail of cellular and tissue arrangement. This is not possible by using light microscopes.

b. Disadvantages of electron microscope

Despite the advantages, electron microscope presents a number of disadvantages and limitations.

- These type of microscope are extremely expensive and the maintenance costs are high.
- Samples must be completely dry so that it is impossible to observe living specimens and moving specimens (they are dead).
- It is not possible to observe colours because electrons do not possess a color. The image is only black-white images/ grey images.
- The energy of the electron beam is very high, the sample is therefore exposed to high radiation, and therefore not able to live.
- The space requirements are high, so that they may need a whole room.
- Electron beams are deflected by molecules in air, so samples have to be placed in a vacuum.

Application activity 9.1

1. Make a comparison between light and electron microscope, highlighting the advantages and disadvantages for each type of microscope.
2. Discuss the advantages and disadvantages (limitations) of an electron microscope.
3. Calculate the magnification of an image measuring 50mm, while the object measures 5µm.

4. If a nucleus measures 100mm on a diagram, with a magnification of X10 000, what is the actual size of the nucleus?
5. Complete the table below:

Parts of microscope	Functions
Base	……
……	rotates to allow use of different power objectives
Coarse focus adjustment	……
……	focuses and magnifies light coming through the slide
Eye piece / ocular lens:	……

6. What is the importance of a light microscope?
7. How can you apply microscope technique rules?
8. Discuss the advantages and disadvantages of the types of electron microscopes in medicine and biology research.
9. Make a comparative study between light and electron microscope focussing on the advantages of each type of microscope.

9.2. Eukaryotic and Prokaryotic cells

Activity 9.2

With microscope, observe mounted slides of bacteria, and plant cells. Draw and label the parts that are common in both plant and bacterial specimens

Eukaryotic cells contain membrane-bound organelles, including a true nucleus enclosed in a nuclear envelope. They include cells of: plants, animals, fungi and protoctista.

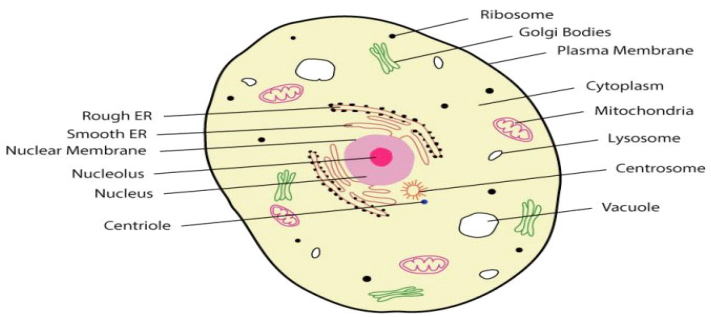


Figure 9.5: Eukaryotic cell

Prokaryotes are organisms having cells with no true nuclear envelope. Prokaryotic cells do not contain a nucleus or any other membrane-bound organelle. Prokaryotes include bacteria and blue-green algae. They make up the monera kingdom.

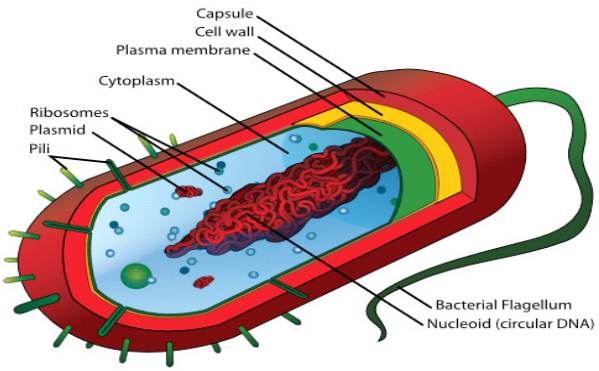


Figure 9.6: Prokaryotic cell

Comparison between prokaryotic and eukaryotic cells

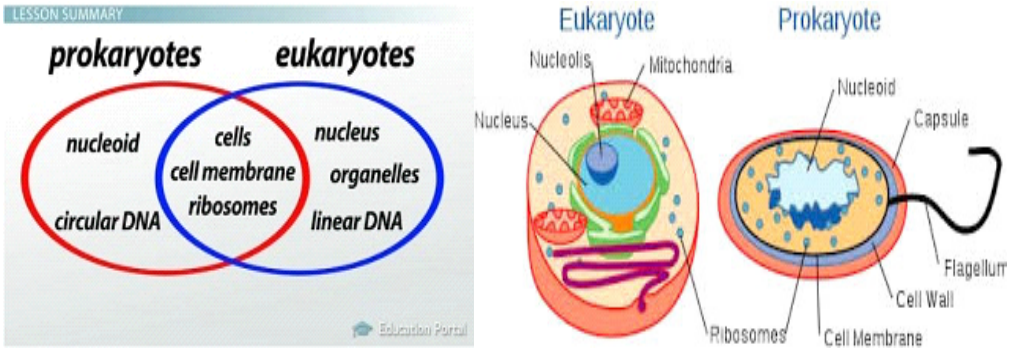


Figure 9.7: Comparison between prokaryotic and eukaryotic cells

Table 9.3: Comparison between prokaryotic and eukaryotic cells

Criterion	Prokaryote	Eukaryotic Plant cell	Eukaryotic Animal cell
Cell membrane	Present	Present	Present
Cell wall	Present	Present	Absent
Nuclear envelope	Absent	Present	Present
Chromosome	Circular	Threadlike	Threadlike
Mitochondria	Absent	Present	Present
Chloroplast	Absent	Present	Absent

Endoplasmic reticulum	Absent	Present	Present
Golgi body	Absent	Present	Present
Ribosomes	Small (70S)	Big (80S)	Big (80S)
Vacuole	Absent	Big	Absent
Lysosomes	Absent	Present	Present
Centrioles	Absent	Absent	Always present

Application activity 9.2

1. Define a prokaryote and eukaryote.
2. In a form of a table differentiate prokaryotic cells from eukaryotic cells.

9.3. Plant and animal cells

Activity 9.3

Observe two electron photographs, one containing a plant cell another an animal cell.Record a description of their features, such as shape and internal parts.

When viewed under light microscope, the most obvious features observed are the very large nucleus and a clear cytoplasm surrounded by a cell membrane. However, under electron microscope, it is possible to identify a range of organelles in plant and animal cells.

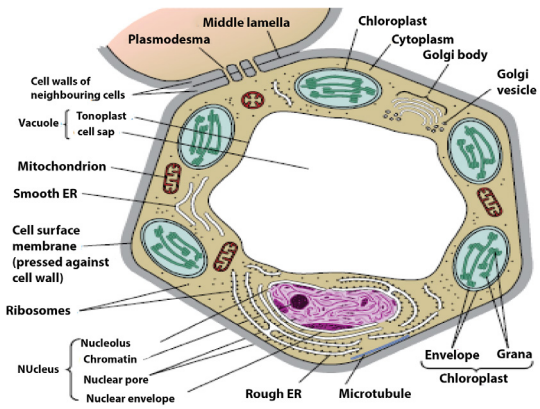


Figure 9.8: Ultrastructure of a typical plant cell (Adapted from Cambridge International AS and A Level Biology Course Book Fourth Edition)

Ultrastructure of a plant cell contains different parts like cell wall, cell membrane, cytoplasm with organelles. Organelles found in the cytoplasm of a plant cell include: chloroplast, mitochondria, Golgi apparatus, endoplasmic reticulum, ribosomes, big central vacuole, and the nucleus which contains chromosomes. The plant cell also has a regular shape, with a relatively bigger size than animal cell.

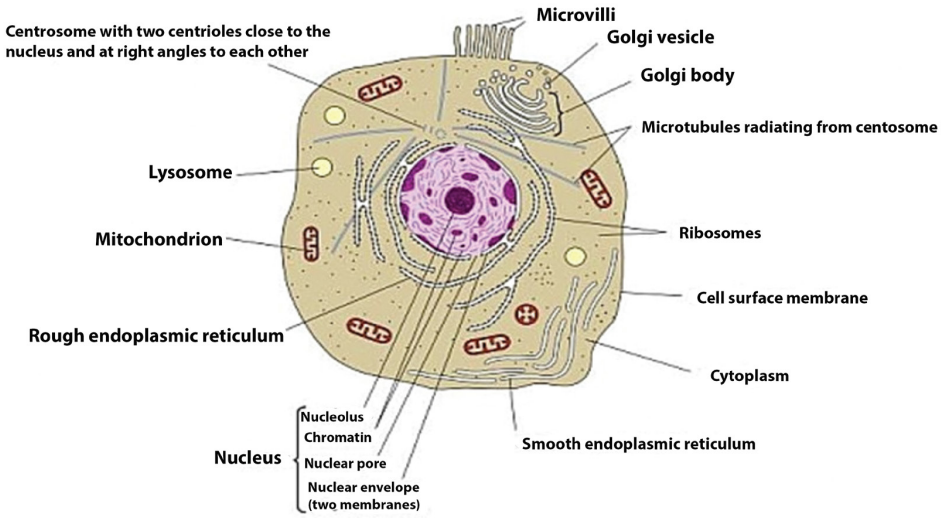


Figure 9.9: A generalized animal and cells showing structures visible under electron microscope. Adapted from Cambridge International AS and A Level Biology Course Book Fourth Edition.

Ultrastructure of an animal cell contains different parts like cell membrane, cytoplasm with organelles. Organelles found in the cytoplasm of an animal cell include: mitochondria, Golgi apparatus, endoplasmic reticulum, centrioles, ribosomes, small or absent vacuole, and the nucleus which contains chromosomes. The animal cell also has irregular shape, with a relatively smaller size than a plant cell.

Application activity 9.3

1. Describe the structure of:
   a). Animal cell and
   b). Plant cell

9.4. Organelles and their functions

Activity 9.4

Under microscope, observe animal cell and plant cells. Explain the functions on all structures of both animal cell and plant cell.

The previous sections described the structures of plant and animal cells. This unit will explain the function of each part of both animal cell and plant cell.

1) Organelles surrounded by membranes

a. Nucleus

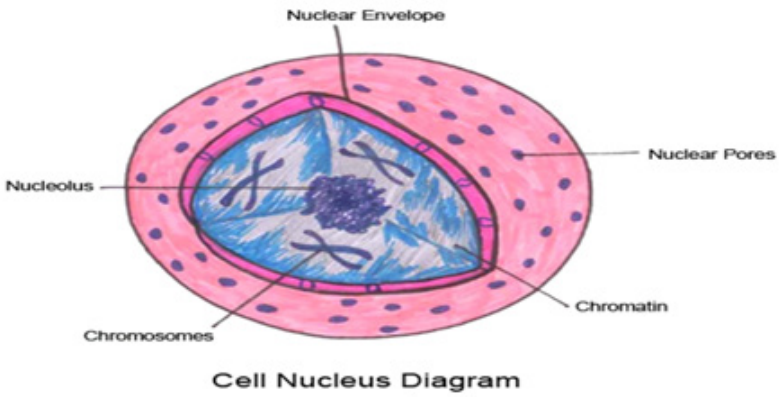


Figure 9.10: Cell nucleus

The cell nucleus contains nearly all the cell's DNA with the coded instructions for making proteins and other important molecules. The nucleus is surrounded by a double nuclear envelope, which allow materials to move into and out of the nucleus through nuclear pores. The granules found in the nucleus are called chromatin which consist of DNA bound to protein. When a cell divides, the chromatin condenses into chromosomes containing the genetic information that is passed from parents to their offspring. The nucleus contains a dense spherical structure called nucleolus in which assembly of ribosomes begins. The nucleus controls all activities of the cell.

b. Endoplasmic reticulum (ER)

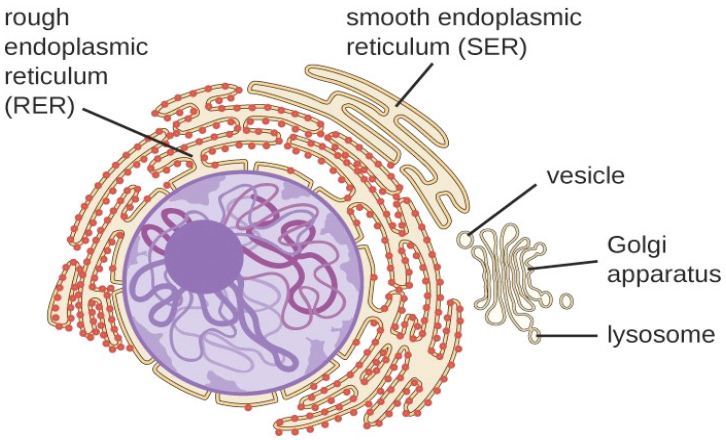


Figure 9.11: Endoplasmic reticulum

The ER consists of a series of flattened membrane-bound sacs called *cisternae*. The rough ER is surrounded with ribosomes. The rough ER transports proteins made on attached ribosomes. The smooth ER does not have ribosomes, and it involves in making lipids that the cell needs.

c. Golgi apparatus

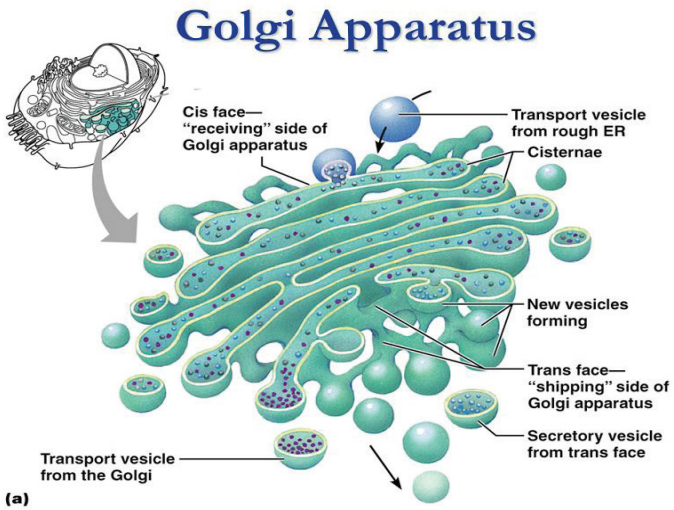


Figure 9.12: Golgi apparatus

The Golgi apparatus is a stack of membrane-bound, flattened sacs, which receives proteins from the ER and modify them. It may add sugar molecules to them. The Golgi apparatus then packages the modified substances into vesicles that can be transported to their final destinations throughout the cell or outside of the cell.

d. Mitochondria

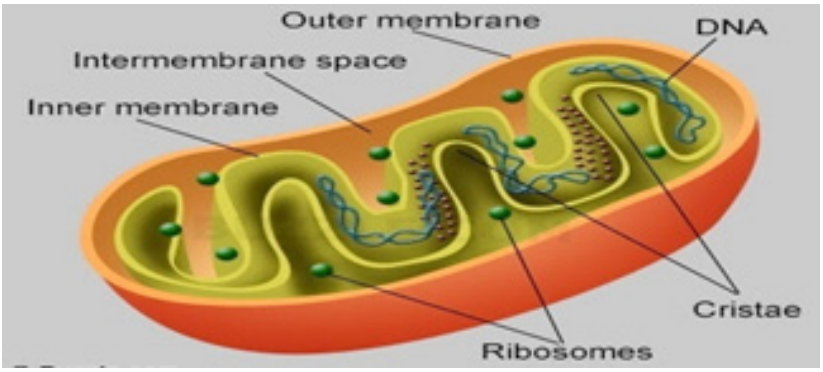


Figure 9.13: Mitochondrion. © Buzzle.com

Mitochondrion have two membranes separated by a fluid-filled space. The inner membrane is highly folded to form cristae. The central part of the mitochondrion is called matrix. The mitochondria are the site where the process of cell respiration takes place to produce Adenosine triphosphate (ATP), a universal energy carrier to be used in cell metabolism.

e. Chloroplasts

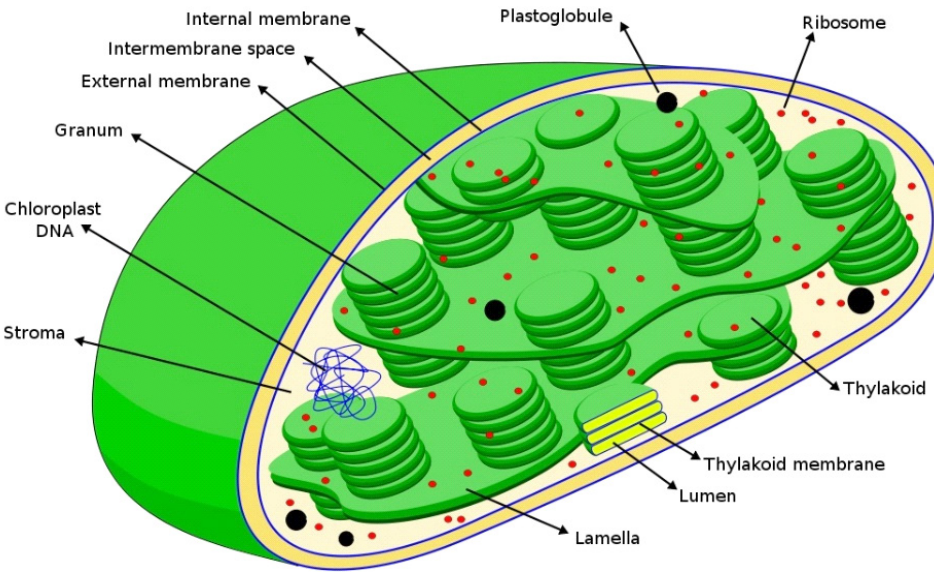


Figure 9.14: Chloroplast

Chloroplasts are the site of photosynthesis in plant cells. These are found in plant cells and in cells of some protoctists. They also have two membranes separated by a fluid-filled space. The inner membrane is continuous, with thylakoids. A stalk of thylakoids is called a granum (plural: grana). Chlorophyll molecules are present on the thylakoid membranes.

f. Lysosomes

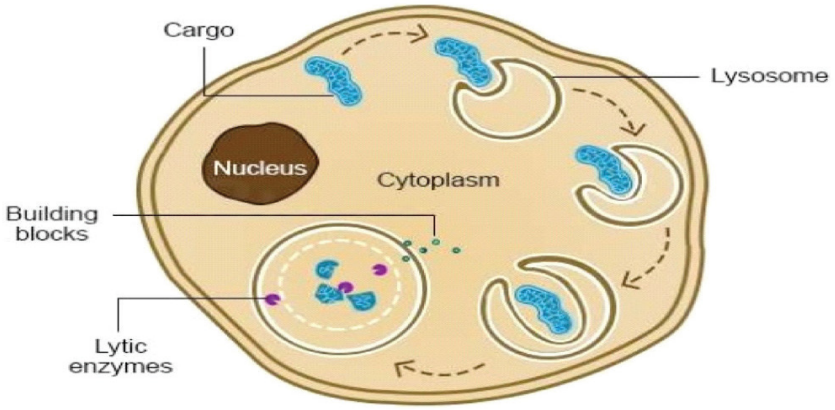


Figure 9.15: Lysosome

These are spherical sacs surrounded by a single membrane. They contain powerful digestive enzymes. Their role is to break down materials such as white blood cells, and destroy invalid microorganisms. In acrosome, lysosomes help the sperm to penetrate the egg by breaking down the material surrounding the egg.

2) Organelles without surrounding membranes

a. Ribosomes

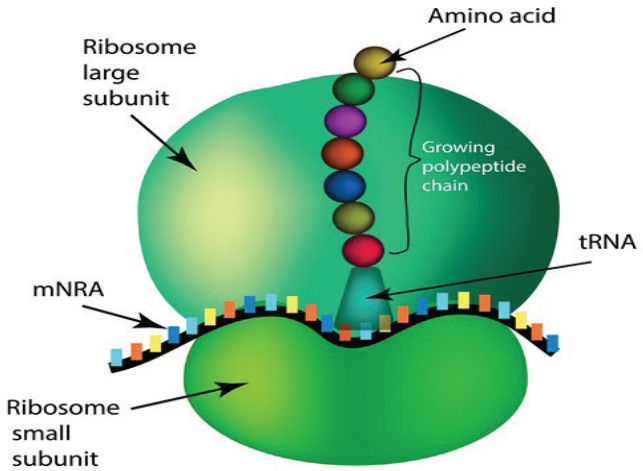


Figure 9.16: Ribosome

Ribosomes are the site of protein synthesis in the cell. Some are in cytoplasm; others are bound to ER. Ribosomes consist of two major components (two subunits): the small ribosomal subunit, which reads the RNA, and the large subunit, which joins amino acids to form a polypeptide chain. Each subunit

is composed of one or more ribosomal RNA (rRNA) molecules and a variety of ribosomal proteins (r-protein). Ribosomes act as an assembly line where coded information (mRNA) from the nucleus is used to assemble proteins from amino acids. Cells that are more active in protein synthesis are often packed with ribosomes.

b. Centrioles

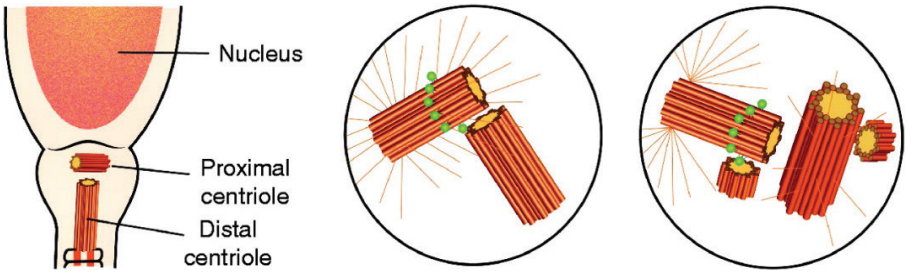


Figure 9.17: Centrioles

Centrioles are small tubes of protein fibres (microtubules), which are involved in animal cell division. They form fibres, known as spindle, which move chromosomes during nuclear division.

c. Vacuole

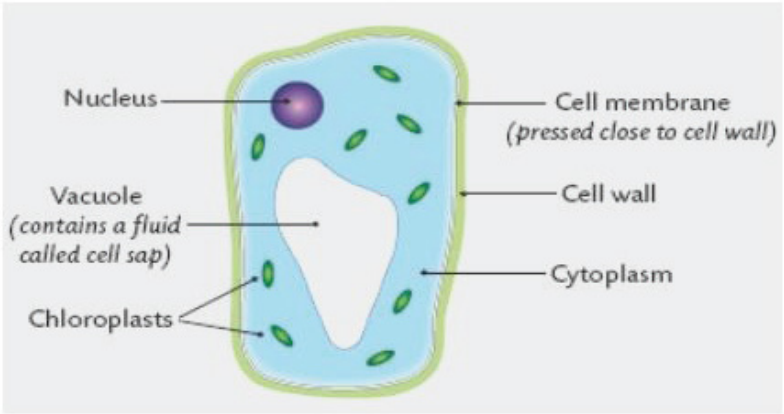


Figure 9.18: Vacuole

A vacuole is a saclike structure that is used to store materials such as water, salts, proteins, and carbohydrates. In many plant cells there is a single, and large central vacuole filled with liquid. The pressure of central vacuole in this cells makes it possible for plants to support heavy structures such as leaves and flowers. Some animals and some unicellular organisms contain contractile vacuoles which contract rhythmically to pump excess water out of the cell.

d. Cytoskeleton

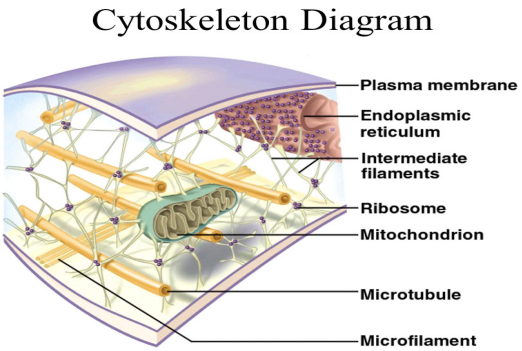


Figure 9.19: Cytoskeleton

Cytoskeleton is a network of protein filaments that helps the cell to maintain its shape. It is also involved in in movement. The main components of cytoskeleton are microfilaments made of a protein called actin, microtubules made of a protein called tubulins, and intermediate filaments.

Definitely, a cell has many organelles with different functions and work together for the survival of the cell.The structure of the cell membrane is based on fluid mosaic model. The term fluid mosaic is used to describe the molecular arrangements in membranes. The main features of the fluid mosaic model are:

- A bilayer of phospholipid molecules forming the basic structure.
- Many protein molecules floating in the phospholipid bilayer. Some are free, others are bound to other components or to structures within the cell.
- Some extrinsic proteins are partially embedded in the bilayer on the inside or the outside face while other intrinsic proteins are completely spanning the bilayer.

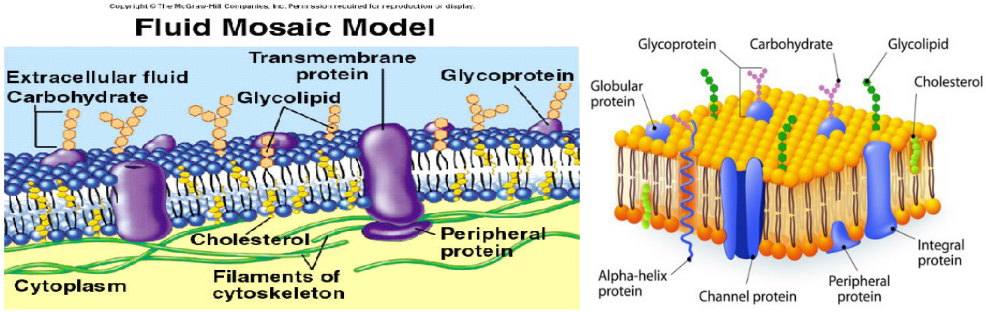


Figure 9.20: Fluid mosaic model of a cell membrane

The basic structure of phospholipids has two parts: hydrophilic part which means water loving and which consists of the phosphate head, and hydrophobic part which means water hating and which consist of fatty acids. If phospholipid molecules are completely surrounded by water, a bilayer can form phosphate heads on each side of the bilayer stick into water, while the hydrophobic fatty acid tails point towards each other.

Types of protein found in the cell membrane

Various types of proteins are found in the cell membrane. They include:

- Carrier proteins: They fix or attach molecules and facilitate them to cross through the cell membrane by active transport.
- Channel proteins: they act as pores by pumping substances and allow facilitated diffusion.
- Receptors: They act as receptors of enzymes and neurotransmitters
- Glycoproteins: They act as receptor proteins which recognize the substance to pass through the membrane
- Integrated proteins: They define the shape of the cell
- Immune proteins (antigens): found in the membrane on the red blood cell, they recognize the antibodies.

Roles of different components of cell membrane

a). Cholesterol

- Gives the membranes of some eukaryotic cells the mechanical stability.
- It fits between fatty acid tails and helps make the barrier more complete, so substances like water molecules and ions cannot pass easily and directly through the membrane.

b). Channel proteins

- Allow the movement of some substances across the membrane.
- Large molecules like glucose enter and leave the cell using these protein channels.

c). Carrier proteins

- Actively move some substances across the cell membrane. For example, magnesium and other mineral ions are actively pumped into the roots hair cells from the surrounding soil.
- Nitrate ions are actively transported into xylem vessels of plants

d). Receptor sites

- Allow hormones to bind with the cell so that a cell response can be carried out.

- Glycoproteins and glycolipids may be involved in cells signaling that they are self to allow recognition by the immune system.
- Some hormone receptors are glycoprotein and some are glycolipid.

e). Enzymes and coenzymes

- Some reactions including metabolic processes in photosynthesis take place in membranes of chloroplasts.
- Some stages of respiration take place in membranes of mitochondria, where Enzymes and coenzymes may be bound to these membranes.
- The more membrane there is, the more enzymes and coenzymes it can hold and this helps to explain why mitochondrial inner membranes are folded to form cristae, and why chloroplasts contain many stacks of membranes called thylakoids.

Properties of the cell membrane

- It is mainly made of lipids, proteins and carbohydrates.
- It is semi-permeable or partially permeable and allow some substances to pass through but prevents others to cross depending on their size, charges and polarity.
- It is positively charged outside and negatively charged inside and has a hydrophilic pole and a hydrophobic pole
- It is a bilayer, sensitive, flexible, has inorganic ions and its proteins and lipids may be mobile and contains different types of enzymes and coenzymes.
- It is perforated of pores and recognizes chemicals messengers including hormones and neurotransmitters.

Functions of a cell membrane

- The cell membrane acts as a selective barrier at the surface of the cell, and controls the exchange between the cell and its environment.
- The membrane proteins act as pores where chemicals pass through.
- Glycocalyx including glycoprotein and glycolipid are involved in the cell protection, the process by which cell adhesions are brought about and also in the uptake and entry of selected substances.

Comparison between animal and plant cell

By referring to the diagrams which show the structure of animal cell and plant cell, similarities and differences between animal cell and plant cell can be identified.

Similarities between animal cell and plant cell

- Both animal and plant cells have a cell membrane, a cytoplasm and a nucleus.
- Both animal and plant cells have a true nucleus bounded by an envelope.
- Both animal and plant cells have mitochondria, Golgi apparatus, Reticulum endoplasmic, lysosome, big ribosomes (80S), peroxisome, microtubules.
- The protoplasm is enveloped by a bounding cell membrane called plasmalemma.
- The protoplasm is composed of a dense round structure called nucleus which is surrounded by a less dense jelly-like cytoplasm.
- The cytoplasm contains numerous organelles such as mitochondria, Golgi bodies, secretory vacuoles, endoplasmic reticulum.
- Mitochondria appear as very small darkly staining, rod-like structures.
- Golgi bodies are semi-transparent irregular, and membrane bound structures.
- Vacuoles contain secretions, food- particles, or decomposing organic substances.
- Chemically, both plant and animal cells are made up of water (80-90%), proteins (7-13%), lipids (1-2%), carbohydrates (1-1.5%) and inorganic salts.
- The cytoplasmic organelles are suspended in a semi-fluid jelly matrix called cytosol.

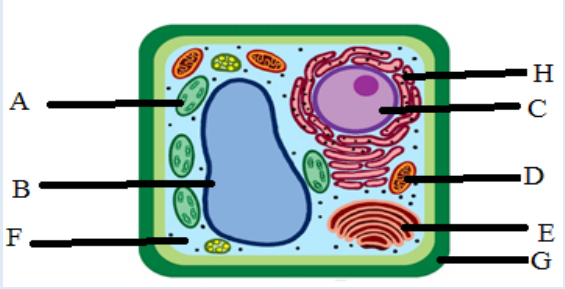
Difference between animal and plant cells

Table 9.2: Comparison between animal and plant cell

Criterion	Animal Cell	Plant Cell
Shape	Ovoid or spherical/ irregular	Polygon / regular
Plastids/chloroplasts	Absent	Present
Centrioles	Present	Absent
Vacuole	Small or absent	Central /Big with a tonoplast
Cell wall	Absent	Present
Microvilli	Present	Absent
Plasmodesmata	Absent	Present
Size	Relatively small	Relatively large

Application activity 9.4

1. What is meant by the fluid mosaic model of the cell membrane?
2. State the properties of the cell membrane.
3. Discuss at least 4 types of the proteins in the cell membrane and their roles.
4. What does partially permeable membrane mean?
5. Explain why muscle cells contain a lot of mitochondria, whereas most fat storage cells do not.
6. What kind of information is contained in chromosomes?
7. You examine an unknown cell under the microscope and discover that the cell contains chloroplasts. What type of organism could you decide that the cell came from?
8. The diagram below shows the structures which would be visible in a plant cell examined under an electron microscope.



   a). Identify the parts labelled in this plant cell and:

   b). State one function for A,B, C, E,D,G, F, and H

9. Explain two functions of the cytoskeleton?
10. What structures do both animal and plant cells have in common?
11. Answer by true or false:

    a). All organelles of a cell are well seen through a compound light microscope.

    b). Chloroplasts are found in both animal and plant cells.

    c). Mitochondria are found only in animal cells.

9.5. Specialized cells

Activity 9.5

Observe the figure below and explain the importance of specialized cells

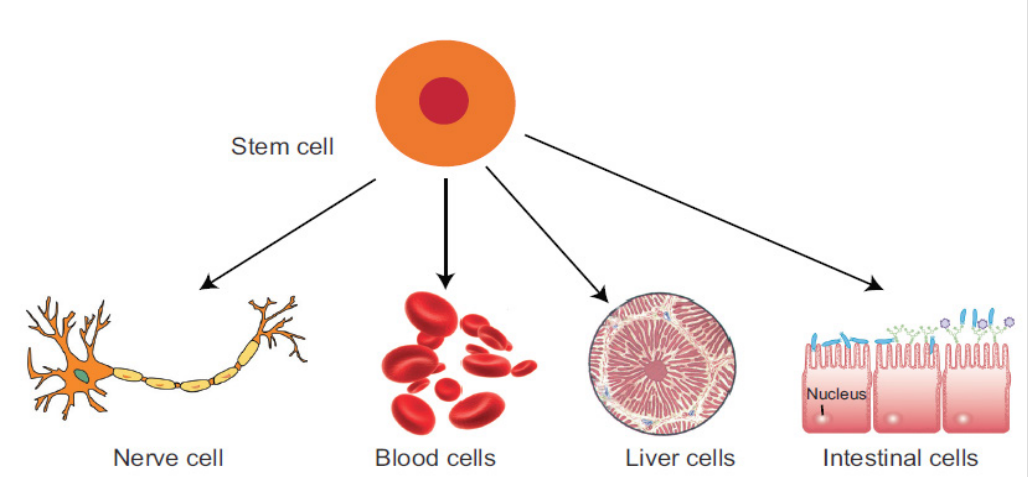


Figure 9.21: Specialized animal cells

9.5.1. Specialized animal cells and their functions

Differentiation refers to the changes occurring in cells of a multicellular organism so that each different type of cell becomes specialised to perform a specific function.

a. Red blood cells

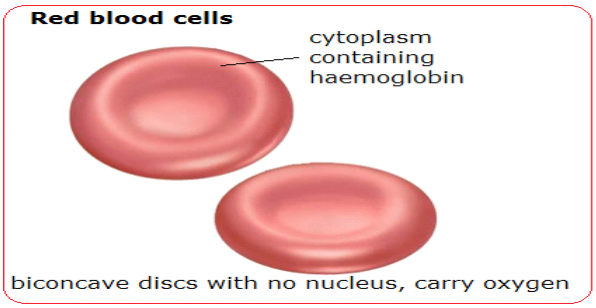


Figure 9.22: Red blood cell

All blood cells are produced from undifferentiated stem cells in the bone marrow but the cells destined to become erythrocytes (red blood cells) lose their nucleus, mitochondria, Golgi apparatus and rough endoplasmic reticulum. They are packed full of the protein called haemoglobin. The shape

of this cells change so that they become biconcave discs, and they are then able to transport Oxygen in the body.

b. Sperm cell

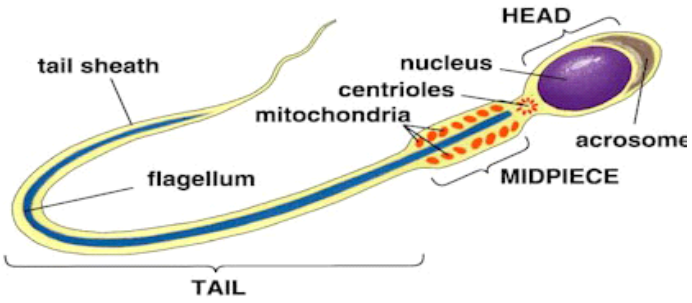


Figure 9.23: Sperm cell

Sperm cells are specialized to fertilize the egg. Its specialization involves many changes in shape and organelles content.

By shape, the sperm cells are very small, long and thin to help them to move easily, and they have a flagellum which helps them to move up the uterine tract towards the egg.

By organelles content, sperm cells contain numerous mitochondria which generate much energy for their movement. Their acrosome releases specialized lysosomes contains many enzymes on the outside of the egg. These enzymes lyse the wall of the egg, and facilitate the sperm nucleus to penetrate easily. In content, the sperm cell nucleus contains the half number of chromosomes of the germ cell in order to fulfil its role as a gamete of fertilizing the egg.

c. Nerve cells

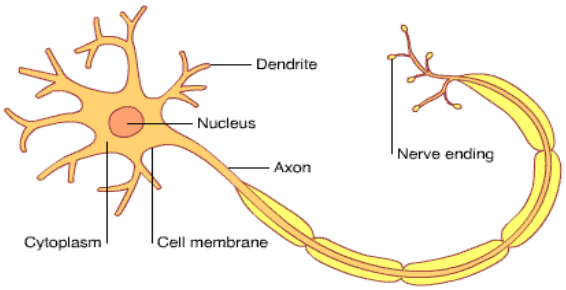


Figure 9.24: Neuron

Nerve cells also known as a neuron are specialized cells to carry nervous information in the body. It is an electrically excitable cell that receives, processes, and transmits information through electrical and chemical signals. These signals between neurons occur via specialized connections called synapses. Specialized animal cells have different functions. Some of them are summarized in the following table.

Table 9.3: Specialized animal cells and their functions.

The specialized cell	Specialization	function
Erythrocytes	They don't have nucleus, mitochondria, Golgi apparatus, and rough ER. However, they have biconcave shape and possess haemoglobin	Transport of Oxygen in the body
Sperm	They have many mitochondria, flagella, acrosome with enzymes, half number of chromosomes, being small, thin and long	Fertilization of the egg
Nerve cell	Have excitability and conductivity characteristics, acrosome, axon, node of Ranvier, Myelin sheaths, and dendrite	Conduction of nervous information.
Epithelial cells	Having different size and shape	Protection of the body, secretion of mucus, digestive enzymes, hormones and sweat. They reproduce to replace dead ones
Smooth muscle cells	Smooth muscle fibres do not have striation	Allow the motion
Pigment cells	They contain colouring substances which determine their colour. For example Melanocytes which secrete melamine for the skin colour	Determine the colour
Flame cells	They contain excretory structures. They are specialized excretory cells found in the simplest freshwater invertebrates, including flatworms.	They function like a kidney, removing waste materials by excretion.
Nematocysts	Nematocysts are specialized stinging cells specific to Coelenterates. They are also called Cnidae and hence the coelenterates are also called Cnidarians. The cells that produce nematocysts are called nematoblasts.	They act as organs of offence and defense. Their venomous coiled thread (bladder) can be projected in self-defense or to capture prey.

9.5.2. Specialized plant cells and their functions

a. Root hair cells

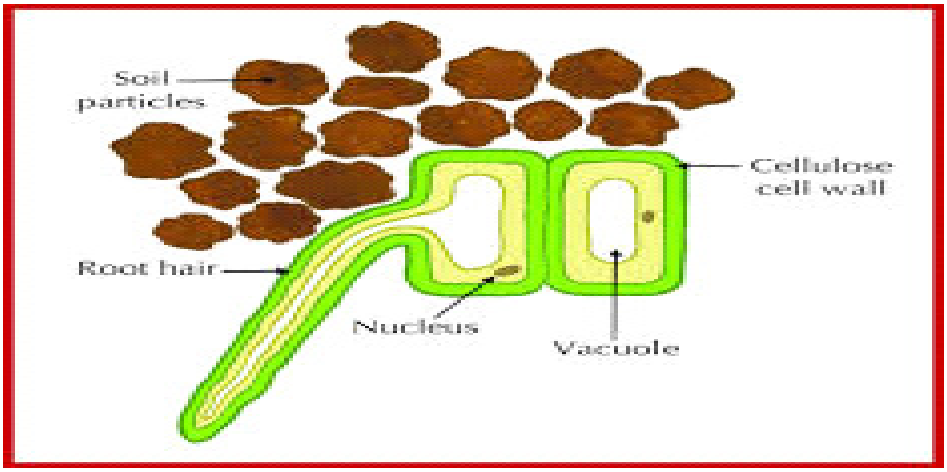


Figure 9.25: Root hair cell

The root hair cells have hair-like projection from their surface out into the soil. This increase the surface area of root available to absorb water and minerals from the soil. Root hairs are tip-growing cells that originate from epidermal cells called trichoblasts. Their role is to extend the surface area of the root to facilitate absorption of mineral nutrients and water.

b. Palisade cells

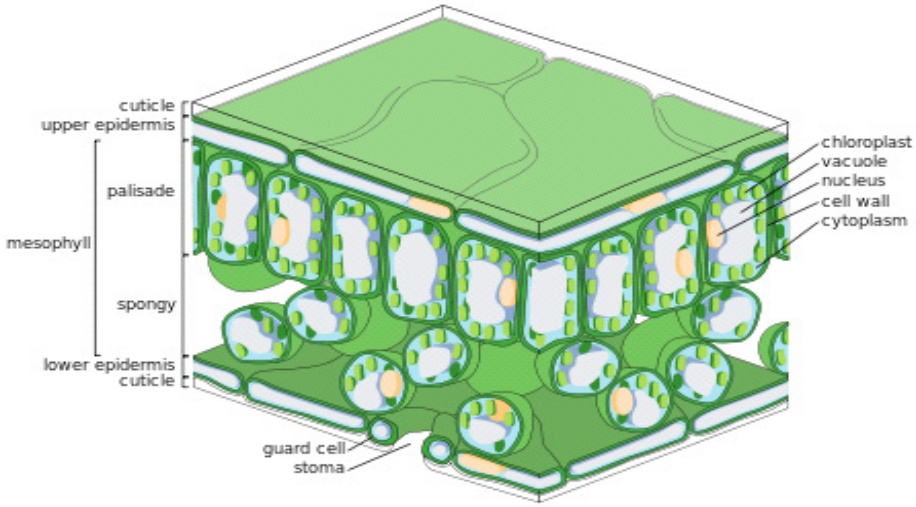


Figure 9.26: Palisade cells

Palisade cells are plant cells located in leaves, right below the epidermis and cuticle. They are vertically elongated, a different shape from the spongy

mesophyll cells beneath them in the leaf. Their big number of chloroplasts allow them to absorb a major portion of the light energy used by the leaf in the process of photosynthesis.

c. Parenchyma cells:

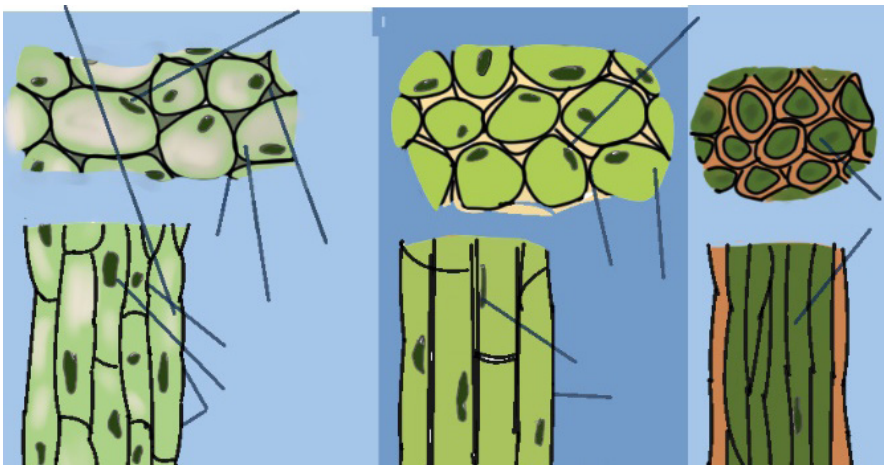


Figure 9.27: Parenchyma cells

Parenchyma cells are alive at maturity. They have function in storage, photosynthesis, and as the bulk of ground and vascular tissues. Palisade parenchyma cells are elongated cells located in many leaves just below the epidermis. Parenchyma is composed of relatively simple, undifferentiated parenchyma cells. In most plants, metabolic activity such as respiration, digestion, and photosynthesis occurs in these cells because they retain their protoplasts (the cytoplasm, nucleus, and cell organelles) that carry out these functions. Parenchyma cells are capable of cell division, even after they have differentiated into the mature form.

d. Guard cells

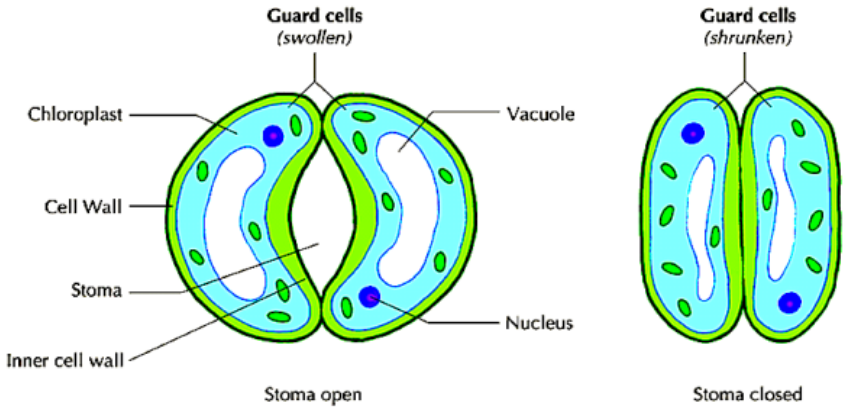


Figure 9.28: Guard cells

Guard cells are cells surrounding each stoma. They help to regulate the rate of transpiration by opening and closing the stomata. Guard cells are specialized cells in the epidermis of leaves, stems and other organs that are used to control gas exchange. They are produced in pairs with a gap between them that forms a stomatal pore.

Application activity 9.5

1. Explain why differentiation to produce erythrocytes involves a change in shape.
2. Red blood cells cannot divide as they have no nucleus. State two other biological processes that red blood cells cannot carry out.
3. Describe how the following are specialized for their roles:
   a) Neutrophil
   b) Sperm cell
   c) Root hair cell
4. Explain why photosynthesis is carried out in palisade mesophyll more than in spongy mesophyll.
5. In what kinds of organisms' cell specialization is a pronounced characteristic?
6. Using what you know about the ways muscles move, predict which organelles would be most common in muscle cells.
7. Discuss the advantages of cell specialization for living things

Skills lab 9

Take tissue cells from a rabbit and from a young leaf of a sweet potato. Prepare the slides containing those tissues. Observe them under the light microscope, and draw their structures on the paper and present to the class.

End unit assessment 9

Section A. Multiple choice questions

1. Which organelle converts the chemical energy in food into a form that cells can use?
   a) Chromosome
   b) Chloroplast
   c) Nucleus
   d) Mitochondrion

2. The cell membranes are constructed mainly of:
   a) Carbohydrate gates
   b) Protein pumps
   c) Lipid bilayer
   d) Free-moving proteins
3. In many cells, the structure that controls the cell's activities is the:
   a) Nucleus
   b) Nucleolus
   c) Cell membrane
   d) Organelle
4. Despite differences in size and shape, all cells have cytoplasm and a
   a) Cell wall
   b) Cell membrane
   c) Mitochondria
   d) Nucleus
5. If a cell of an organism contains a nucleus, the organism is a (an)
   a) Plant
   b) Eukaryote
   c) Animal
   d) Prokaryote
6. Match each part of the cell to its correct statement:

Nucleus	**controls movement of substances in and out of the cell**
Mitochondrion	where photosynthesis takes place
Chloroplast	where aerobic respiration takes place
Smooth ER	controls the activity of the cell
Ribosomes	where lipids including steroids are made

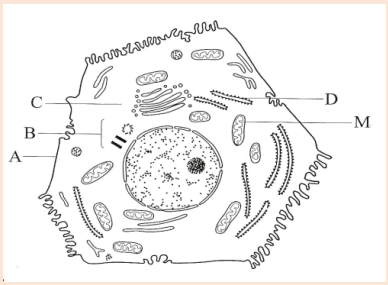
Section B: Questions with short answers

1. How does a cell membrane differ from a cell wall?
2. Name the structures that animal and plant cells have in common, those found in only plant cells, and those found only in animal cells.

3. List:
   a) Three organelles each lacking a boundary membrane
   b) Three organelles each bounded by a single membrane
   c) Three organelles each bounded by two membranes (an envelope)
4. Identify each cell structure or organelle from its description below.
   a) Manufactures lysosomes and ribosomes
   b) Site of protein synthesis
   c) Can bud off vesicles which form the Golgi body
   d) Can transport newly synthesized protein round the cell
   e) Manufactures ATP in animal and plant cells
   f) Controls the activity of the cell, because it contains the DNA
   g) Carries out photosynthesis
   h) Can act as a starting point for the growth of spindle microtubules during cell division
   i) Contains chromatin
   j) Partially permeable barrier only about 7 nm thick
   k) Organelle about 25 nm in diameter
   l) Which two organelles other than the nucleus contain their own DNA?

Section C: Essay questions

1. Describe the structure and function of the cell membrane and cell wall.
2. Describe the basic structure of the cell membrane.
3. Explain two common characteristics of chloroplasts and mitochondria. Consider both function and membrane structure.
4. The diagram below shows the structure of a liver cell as seen using an electron microscope.

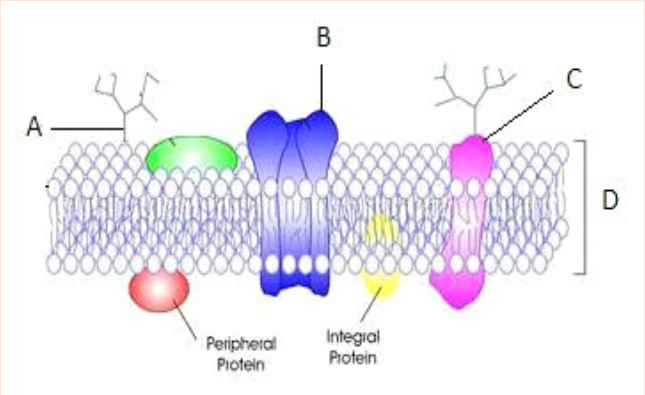


a) Name the parts labelled A, B, C and D.

b) The magnification of the diagram above is 12 000. Calculate the actual length of the mitochondrion labelled M, giving your answer in μm. Show your working.

c) Explain the advantage to have a division of labor between different cells in the body.

5. The diagram below shows the structure of a cell membrane. Observe carefully the diagram and answer the questions that follows:



a) Name parts labelled A – D and give the function of the part B.

b) State four factors that can affect diffusion across the cell membrane.

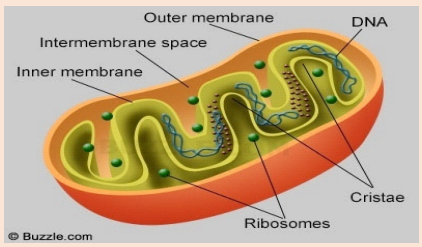
6. What types of molecule are likely to be involved in?

a) Cell signalling and recognition?

b) Allowing small charged molecules to pass through the cell membrane?

7. What is the difference between rough and smooth endoplasmic reticulum?

8. The photograph in the figure below shows an organelle of the living cell.



a). Name this organelle.

b). What is the function of this organelle?

UNIT 10:

CELL AND NUCLEAR DIVISION

Key Unit competence: Describe the stages of the cell cycle and explain the significance of cell and nuclear division in organisms

Introductory Activity 10

Observe the figure 10.1 of the house built in bricks below. The bricks came from the valley where they are made, and then have been used to build this house.

Figure 10.1: A house in bricks

1. Is it possible to build a house like this by using only one brick?
2. Link the analogous example of building a house by using bricks with growth of and increasing in size of a human body from the body size of a one day new baby to the body size of 35 years person.
3. Is it possible for the body of a person to be made by only one cell?
4. Where are cells which make the human body come from?
5. How are they produced?

10.1. Cell cycle

Activity 10.1

The male and female gametes fuse to form a zygote and after the zygote goes through different phases and a baby is born.visit your smart classroom, and discuss more information about the main phases of the cell cycle.

The cell cycle is a series of events of cellular growth and division that has five phases such as:

- The first growth phase (G_1),
- The synthesis phase (S),
- The second growth phase (G_2),
- Mitosis (M),
- Cytokinesis.

The three first phases (G1, S, and G2) of the cell cycle collectively are known as interphase (the phase between two mitotic divisions).

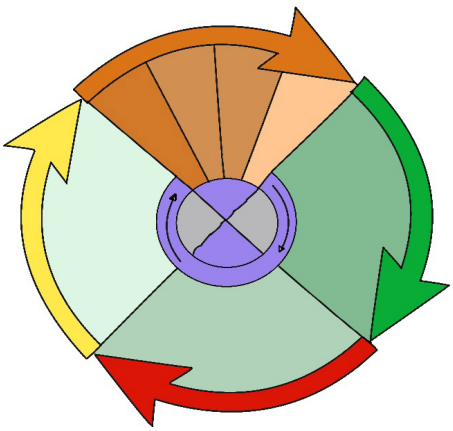


Figure 10.2: Cell cycle

During the cell cycle, the cell grows, grows its DNA and divides into two daughter cells.

10.1.1. The first growth phase (G_1)

During this phase, a cell undergoes rapid growth and the cell performs its routine functions. The cell spends most of its life in the G_1 phase. If the cell is not dividing, it remains in this phase. The time taken for the completion of G1 phase varies among species and the type of cells. But on an average, it takes around 11 hours for the completion of this phase.

10.1.2. The synthesis phase (S)

DNA synthesis or DNA replication takes place in this phase. The S phase takes around 8 hours to complete.

10.1.3. The second growth phase (G_2)

It is the short period in which the cell continues to grow, making proteins and manufacturing many organelles necessary for cell division. This phase serves as an intermediate between the synthesis phase and the mitotic phase. It takes around four hours to complete

10.1.4. Mitosis

This is the phase of nuclear division in which one nucleus divides through four phases (prophase, metaphase, anaphase and telophase) and becomes two nuclei.

10.1.5. Cytokinesis

In this phase, the cytoplasm divides in half, producing two daughter cells, each containing a complete set of genetic material.

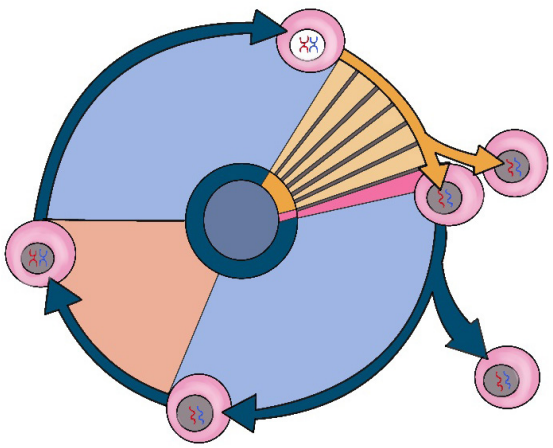


Figure10.3: Events of cell cycle

Application activity 10.1

1. What do you understand by the cell cycle?
2. Describe all phases of the cell cycle.
3. Predict what may happen if the cytokinesis does not take place in the succession of cell cycle.

10. 2. Mitosis and Meiosis: Stages and results

Activity 10.2

Many of schools in Rwanda have the smart classrooms. Visit your school smart classroom and download a video showing the phases of mitosis and meiosisand present them on manila papers.

10.2.1. Mitosis

Mitosis is a type of cell division that produces two daughter cells having the same number and kind of chromosomes as the mother cell. It takes place only in eukaryotes, where the nuclear division (karyokinesis) occurs. This phase takes around 1 hour to complete. The mitotic cell division is more rapid at the meristematic region of plant and root tip as it is the growing region of the plant. The mitotic phase is divided into four steps that include Prophase, Metaphase, Anaphase and Telophase.

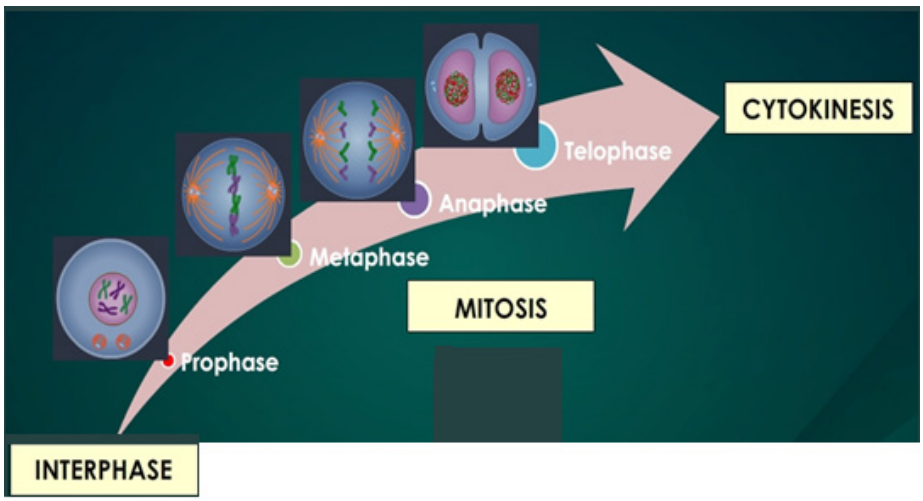


Figure 10.4: The process of mitosis

a. Prophase

Prophase is the first and longest phase of mitosis. During prophase:

- The DNA and histone proteins coils up into visible chromosomes, each made up of two sister chromatids held together by the centromere.
- The nucleus disappears as the nuclear envelope and nucleolus break apart.
- The centrioles begin to move to opposite ends, or poles, of the cell.
- As the centrioles migrate, the fiber-like spindle begins to elongate

between the centrioles. In plant cells, the spindle forms without centrioles. The spindle plays an essential role moving chromosomes and in the separation of sister chromatids.

End of interphase Prophase

Figure 10.5: Prophase

b. Metaphase

During metaphase, the spindle which attaches to the centromere of each chromosome helps the chromosomes to line up at the center of the cell by forming the equatorial plate also known as the metaphase plate. Each sister chromatid is attached to a separate spindle fiber, with one fiber extending to one pole, and the other fiber extending to the other pole.

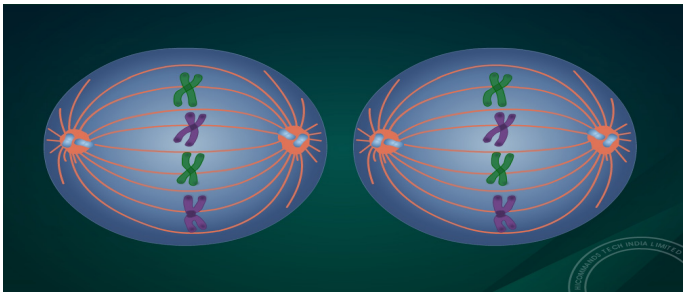


Figure 10.6: Process of metaphase

c. Anaphase

During Anaphase:

- Centromeres divide,
- The sister chromatids separate and pulled apart by the shortening of the spindles,
- One sister chromatid moves to one pole of the cell, and the other sister chromatid moves to the opposite pole, (sister chromatids take the name of chromosomes as soon as they separate).
- At the end of anaphase, each pole of the cell has a complete set of chromosomes, identical to the amount of DNA at the beginning of G1 of the cell cycle.

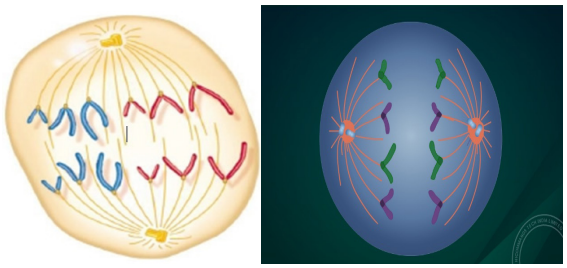


Figure 10.7: Process of Anaphase

d. Telophase

During telophase which is the opposite of prophase:

- The spindle disappears,
- Formation of two nuclei,
- The nuclear envelopes surround the two nuclei.

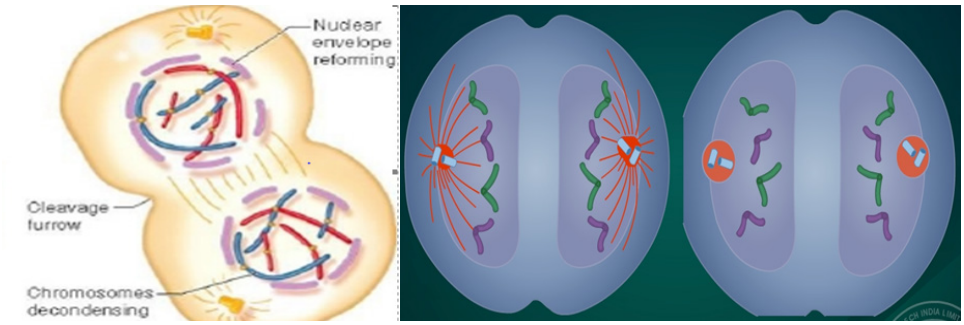


Figure 10.8: The process of telophase

e. Cytokinesis

During this phase, the cytoplasm divides in half, producing two daughter cells, each containing a complete set of genetic material as the mother cell.

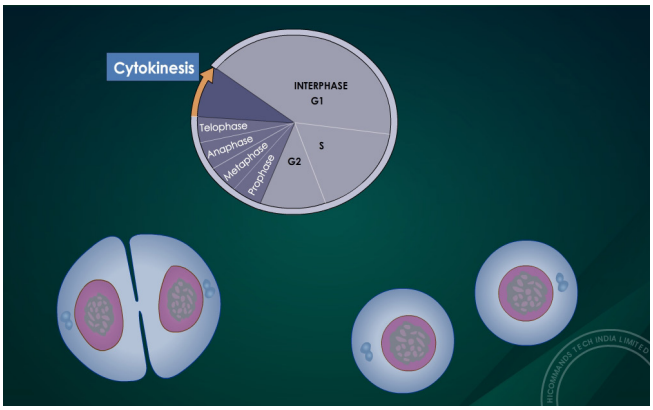


Figure 10.9: The process of cytokinesis

Table 10.2: A table showing phases of cell cycle with one important event at each phase

S. No.	Phase	Main Event
1.	MITOTIC PHASE (M)	
	Prophase	• Condensation of chromosomes in the nucleus
	Prometaphase	• Spindle fibres gets attached to the chromosomes and start aligning chromosomes at metaphase plate
	Metaphase	• Chromosomes align at the metaphase plate
	Anaphase	• Chromosomes are separated in two clusters at the two poles
	Telophase and Cytokinesis	• The two clusters of chromosomes have reached the poles and formed two daughter cells • In cytokinesis, separation of cytoplasm has taken place
2.	INTERPHASE	
	G0	• It is a resting phase.
	G1	• The cell prepares itself for DNA replication.
	S	• DNA replication takes place.
	G2	• The cell resumes growth before the cell enters the next phase i.e. mitotic phase.

10.2.2. Meiosis

In sexual reproduction, meiosis produces haploid gametes that fuse together during fertilization to produce a diploid zygote. Meiosis involves two divisions without an interphase in between, starting with one diploid cell and generating four haploid cells. Each division, named meiosis I and meiosis II, has four stages: prophase, metaphase, anaphase, and telophase.

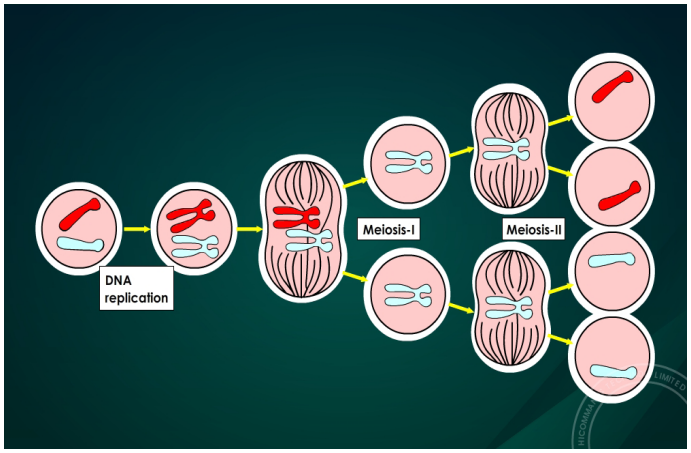


Figure 10.10: Process of meiosis

During meiosis the number of chromosomes is reduced from a diploid number (2n) to a haploid number (n). During fertilization, **haploid gametes** come together to form a **diploid zygote**, and the original number of chromosomes (2n) is restored.

Eight stages of meiosis are summarized below:

1. Prophase I

Prophase I is very similar to prophase of mitosis, but with one very significant difference. In prophase I, the nuclear envelope breaks down, the chromosomes condense, and the centrioles begin to migrate to opposite poles of the cell, with the spindle fibers growing between them.

During this time, the homologous chromosomes form pairs. These homologous chromosomes line up gene-for-gene down their entire length, allowing the crossing-over to occur.

This process permits the exchange of genetic material between maternal and paternal chromosomes. Thus, crossing-over results in genetic recombination by producing a new mixture of genetic material. This is an important step in creating genetic variation.

2. Metaphase I

In metaphase I, the 23 pairs of homologous chromosomes line up along the equator of the cell.

3. Anaphase I

During anaphase I the spindle fibers shorten, and the homologous chromosome pairs are separated from each other. One chromosome from each pair moves toward one pole, with the other moving toward the other pole, resulting in a nucleus with 23 chromosomes at one pole and the other 23 at the other pole. The sister chromatids remain attached at the centromere.

4. Telophase I

he spindle fibers disappear and the nucleus reforms. This is quickly followed by cytokinesis and the formation of two haploid cells, each with a unique combination of chromosomes, some from the father and the rest from the mother.

After cytokinesis, both cells immediately enter meiosis II, without replication of the DNA. Meiosis I is described as reductional division as it reduces by half the number of chromosomes of the mother cell. Meiosis II is equational division, and it occurs like a normal mitosis, separating the sister chromatids from each other.

5. Prophase II

Once again the nuclear membrane breaks down, and the spindle begins to reform as the centrioles move to opposite sides of the cell.

6. Metaphase II

The 23 chromosomes, each made out of two sister chromatids, occupy the equator of the cell.

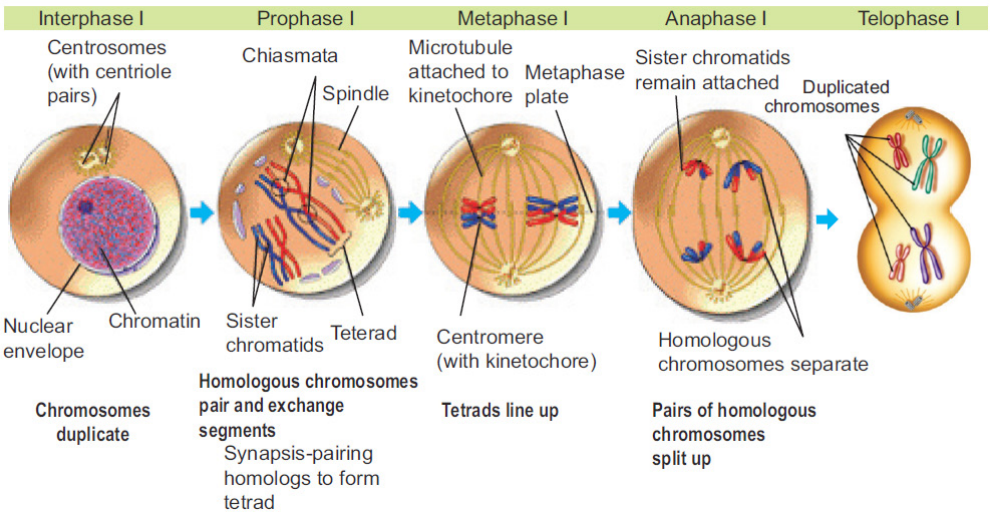
7. Anaphase II

The centromere divides and sister chromatids are separated and move to opposite poles of the cell. As the chromatids separate, each is known as a chromosome. Anaphase II results in a cell with 23 chromosomes at each end of the cell; each chromosome contains half as much genetic material as at the start of anaphase II.

8. Telophase II

The nucleus reforms and the spindle fibers break down. Each cell undergoes cytokinesis, producing **four haploid cells**, each with a unique combination of genes and chromosomes.

Meiosis I



Meiosis II

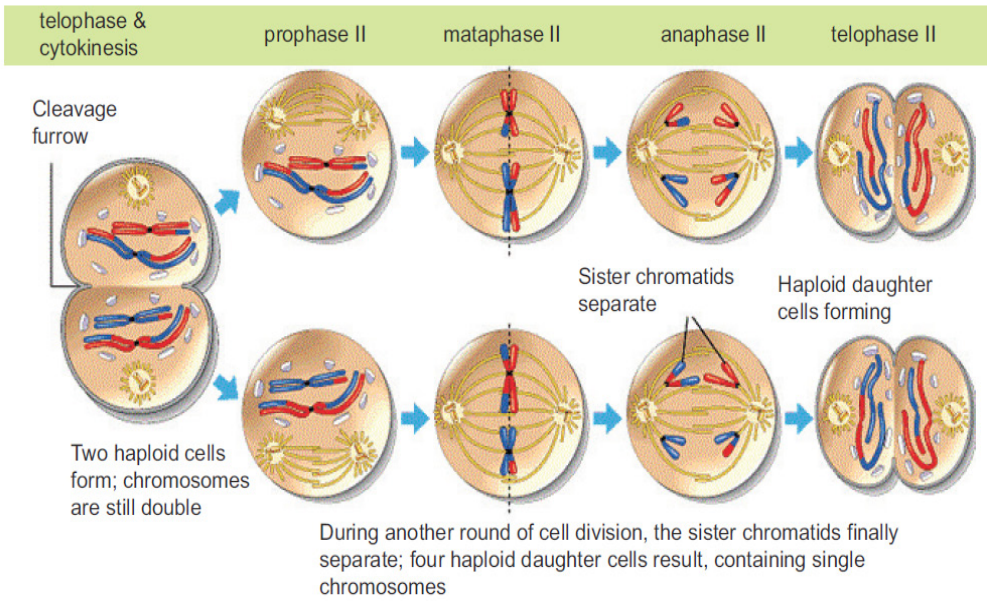


Figure 10.11: The process of meiosis

Comparison of Mitosis and Meiosis

Table 10.3: Comparison between mitosis and meiosis

S. No.	Property	Mitosis	Meiosis
1.	Occurrence	Somatic cells	Germ cells/sex cells
2.	Number of daughter cells	Two diploid cells (2n)	Four haploid cells (1n)
3.	Genetic composition	• The two daughter cells are genetically identical to the parent cell. • It is equational division	• The four daughter cells are genetically different from parent cell and form each other. • It is reduction division
4.	Number of divisions	Prophase, prometaphase, metaphase, anaphase, and telophase	Meiosis I: Prophase I, Metaphase I, Anaphase I, Telophase I Meiosis II: Prophase II, Metaphase II, Anaphase II, Telophase II
5.	DNA Replication	Occurs during S phase of interphase prior to mitosis	Occurs during S phase of interphase prior to meiosis I

6.	Functions in the animal body	• It enables multicellular adult to arise from zygote • It helps in production of cells in growth and repair • Asexual reproduction in some animals	• It produces gametes (sperms and eggs) • It reduces the number of chromosome by half from diploid (2n) to haploid (1n) • It introduces genetic variation among the gametes
7.	Synapsis	It does not occur	Synapsis through synaptonemal complex
8.	Crossing over	It does not occur	Crossing over occurs between two non-sister chromatids during meiosis I
9.	Chiasmata	No chiasmata formation	Chiasmata, sites of crossing over, formation occurs
10.	Homologs on the metaphase plate	Individual chromosomes aligned at metaphase plate	Homologous pairs of chromosomes are aligned at metaphase plate during metaphase I
11.	Sister chromatids	Sister chromatids separate at anaphase	During meiosis I, the replicated chromosomes of each homologous pair move toward opposite poles at anaphase I; however, sister chromatids separate only at anaphase II
12.	Cytokinesis	Cytokinesis occurs after mitosis	Cytokinesis doesn't occur after meiosis I but occurs after meiosis II
13.	Centromeres	Division of centromeres take place at anaphase	Division or cleavage of centromeres takes place only at anaphase II
14.	Chromosomes at metaphase plate	Chromosome pairs are aligned at metaphase plate	Duplicated chromosome pairs are aligned at metaphase plate

Haploid and diploid conditions of the cell

During the formation of **gametes**, the number of chromosomesis reduced by half, and returned to the full amount when the two gametes fuse during **fertilization**. The cells of human beings, most animals and many plants (except for their gametes) are diploid abbreviated as 2n. They contain two sets of chromosomes in their nuclei. **The haploid cells** have only one set of

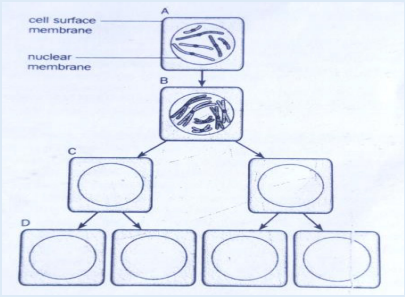
chromosomes, abbreviated as n. Ploidy is a term referring to the number of sets of chromosomes.

Organisms with more than two sets of chromosomes are termed polyploidy. Chromosomes that carry the same genes are termed homologous chromosomes. Meiosis is a special type of nuclear division which segregates one copy of each homologous chromosome into each new "gamete". Mitosis maintains the cell's original ploidy level (for example, one diploid 2n cell producing two diploid 2n cells; one haploid n cell producing two haploid n cells; etc.). Meiosis, on the other hand, reduces the number of sets of chromosomes by half, so that when gametic recombination **fertilization** occurs the ploidy of the parents will be re-established.

Two successive nuclear divisions occur, Meiosis I (Reduction) and Meiosis II (Division). Meiosis produces 4 haploid cells. Mitosis produces 2 diploid cells. The old name for meiosis was reduction/ division. Meiosis I reduces the ploidy level from 2n to n (reduction) while Meiosis II divides the remaining set of chromosomes in a mitosis-like process (division). Most of the differences between the processes occur during Meiosis I.

Application activity 10.2

1. Give two reasons why cells divide.
2. As a cell increases in size, which increases more rapidly, its surface area or its volume?
3. Describe what happens during each of the four phases of mitosis.
4. How is cytokinesis in plant cells similar to cytokinesis in animal cell?
5. The diagram shows meiosis in an animal cell.



a). What is the diploid number of chromosomes in this cell?

b). Where do you think this cell could be found in an animal?

c). What is the stage of cell division shown at B? Give a reason for your choice.

10.3. Mitosis and Meiosis roles in living organisms

Activity 10.3

Observe figures below and explain the significance of mitosis in living organisms.

Figure 10.12: Growth by mitosis A

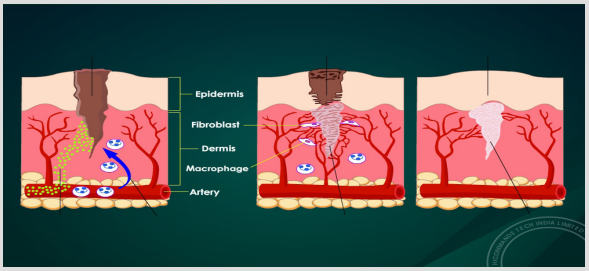


Figure 10.13: Repair of tissue by mitosis B

The Significance of Mitosis in Cell Replacement and Tissue Repair by Stem Cells

a). **Mitosis allows growth:** The number of cells within an organism increases by mitosis and this is the basis of growth in multicellular cells.

b). **Mitosis allows to repairing and cell replacement:** Replacement of cells and tissues also involves mitosis. Cells are constantly dying and are being replaced, for example, in the skin

c). **Mitosis is involved in asexual reproduction:** a single parent cell divides into two genetically identical offspring.

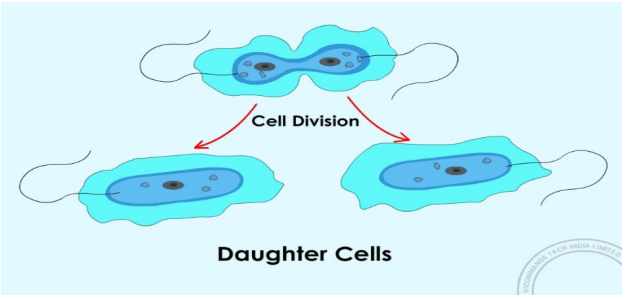


Figure 10.14: Asexual reproduction by mitosis

d). **Mitosis allows genetic stability** by producing two nuclei which have the same number of chromosomes as the parent cell.

e). **Regeneration:** Some animals are able to regenerate whole parts of the body, such as legs in crustacean and arms in starfish. Production of the new cells involves mitosis.

Cancer and uncontrolled cell division

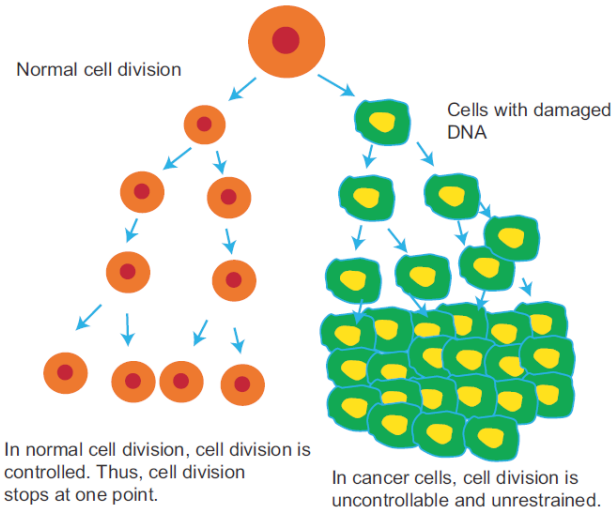


Figure 10.15: For normal cell division and uncontrolled cell division (cancer cells)

Many cancers result from uncontrolled cell division, when the regulation of the cycle is lost. Cancerous cells divide much more rapidly than healthy cells. These cells use the blood and nutrients that other cells need and they can stress the environment of the healthy cells. As cancerous cells do not provide any useful function to the organism, they are extremely harmful. **If cancerous cells are allowed to grow uncontrolled**, they will kill the host organism.

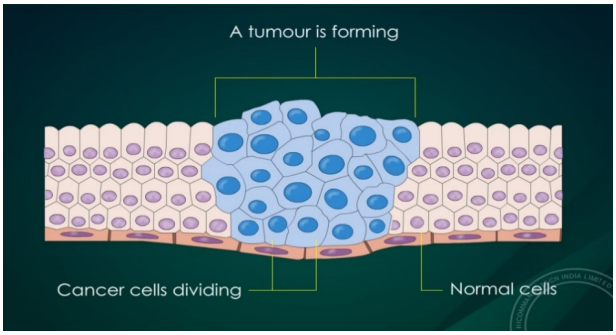


Figure 10.16: Tumour formation

The problem begins when a single cell in a tissue undergoes transformation, the process that convertsa normal cell to a cancer cell. The body's immune system normally recognizes a transformed cell asan abnormal and destroys it.

However, if the cell escapes immune system, it may proliferate to form atumor (a mass of abnormal cells within an otherwise normal tissue). There are three types of tumors: benign tumors, malignant tumors and metastasis.

i. **Benign tumor:** It is a lump of the abnormal cells that remains at the original site. Most benign tumors do not cause seriousproblems and can be removed by surgery.

ii. **Malignant tumors:** These are abnormal cells that have become invasive enough to impair with the functions of one or more organs. An individual with a malignant tumor is said to have cancer.

iii. **Metastasis:** Cancer cells may also separate from the original tumor, enter the blood and lymph vessels, and invade other parts of the body, where they proliferate to form more tumors. This spread of cancer cells beyond their original site is called **metastasis**.

Many cancers can be inherited, such as breast cancer, others are triggered by viral infections, tobacco smoke (e.g. lung cancer) and radiations (e.g. skin cancer). All cancers have one thing in common: the control over the cell cycle has broken down.

Significance of Meiosis

a. Cells undergo Reduction Division Prior to Sexual Reproduction

Generally, a cycle of reproduction consists of meiosis and fertilization. Before sexual reproduction occurs, gametes undergo meiosis and produce haploid cells. Thus during sexual reproduction, one haploid (1n) gamete comes from the paternal side and another haploid (1n) gamete comes from the maternal side; then, they both fuse to form a zygote, which is diploid (2n). The fusion of gametes to form zygote or new cell is called as fertilization or syngamy. If meiosis does not occur before sexual reproduction, the chromosome number would double up with each fertilization. And after few generations, the number of chromosomes in each cell would become impossibly large. For example, in humans, in just 10 generations, the 46 chromosomes would increase to about 47104 (46×2^{10}).

b. Role and Significance of Meiosis in Producing Gametes

Gametogenesis is a biological process by which diploid cells undergo cell division and differentiation to form mature haploid gametes. It occurs through meiosis. In humans, the male gamete (sperm) is produced by a process called spermatogenesis and the female gamete (egg) is produced by a process called oogenesis through meiotic division. Here gametes function takes place soon after meiosis but in plants it happens after gametophyte formation sexual reproduction of plants starts with spore formation. Sporophyte is a diploids generation of flowering plant where haploid spores are produced by meiosis which in turns undergoes mitosis to form multi-celled haploid gametophytes. These haploid gametophytes differentiate to

produce gametes—sperm and egg cells. Similarly, embryo sac is formed by reduction division. Each of the cells of embryo sac is haploid. Two of the nuclei fuse to produce diploid nucleus.

c. The Role of Meiosis in Reproduction of Plants

Generally, plants reproducing sexually have life cycle consisting of two phases

- A multicellular gametophyte or haploid stage:

  It is a haploid stage with n chromosomes. It alternates with a multicellular sporophyte stage.

- A multicellular sporophyte or diploid stage:

  It is a diploid stage with 2n chromosomes, made up of n pairs. A mature sporophyte produces spores by meiosis, a process which reduces the number of chromosomes from 2n to 1n.

Alternation of generations (also known as mutagenesis) refers to the occurrence in the plant life cycle of both a multicellular diploid organism (sporophyte) and a multicellular haploid organism (gametophyte), each giving rise to the other. This is in contrast to animals, in which the only multicellular phase is the diploid organism (such as the human man or woman), whereas the haploid phase is a single egg or sperm cell. In bryophytes (mosses and liverworts), the dominant generation is haploid, so that the gametophyte comprises the main plant. On the contrary, in trichophytes (vascular plants), the diploid generation is dominant and the sporophyte comprises the main plant.

d. Independent Assortment of Chromosomes

Specifically, at metaphase I, each homologous pair of chromosomes positioned independently of the other pairs. As a result, each homologous pair sorts out its maternal and paternal homologue into daughter cells independently of every other pair. This act of separating homologous pairs independently is called independent assortment. The random orientation of homologous pairs of chromosomes due to independent assortment in meiosis I (metaphase) increases genetic variation in organisms.

e. Crossing Over and Random Fertilization

During crossing over, DNA segments of the two parents-paternal and maternal are combined into a single chromosome producing recombinant chromosomes, which are non-identical with their sister chromatids. In humans, an average of one to three crossing over events occurs per

chromosome pair, depending on the position of their centromeres and on the size of the chromosome. Thus, crossing over is an important event of meiosis that brings genetic variation in sexual life cycles.

Besides independent assortment and crossing over, the random fertilization during sexual reproduction also increases genetic variation in organisms. During random fertilization, the male gamete and female gamete fuse to form zygote. The most interesting thing is that this zygote has the possibility of about 70 trillion diploid combinations. The number 70 trillion comes from possible combinations of male and female gametes which are $2^{23} \times 2^{23} = 70$ trillion. The possibility of this enormous number of combinations makes each and every one of us unique physically and genetically.

f. Non-disjunction of Chromosomes

Proper separation of chromosomes during meiosis is essential for the normal growth in humans. Any set of chromosomes that do not separate properly during meiosis results in improper separation of chromosomes or non-disjunction, which is a serious issue in human genetics. Non-disjunction is a condition in which the homologues or sister chromatids fail to separate properly during meiosis. It can lead to the gain or loss of chromosome, a condition called as aneuploidy. Example: Down syndrome is an autosomal trisomy. It is also called as trisomy 21, where non-disjunction results in an embryo with three copies of chromosome 21 instead of the usual two copies of chromosome 21. The origin of trisomics condition is through non-disjunction of chromosome 21 during meiosis. Failure of paired homologues to separate during either anaphase I or II may lead to gametes with 23 + 1 chromosome composition instead of the normal 23 gamete chromosome composition. Therefore, instead of 46 normal chromosomes, Down syndrome patient will have 47 chromosomes with three copies of chromosome 21 instead of the normal 2 copies. It was first discovered by John Langdon Down. The chance of occurrence is one infant in every 800 live births.

The most common symptoms of Down syndrome or trisomy 21 are:

- They are short.
- They may also have protruding, furrowed tongues, which causes the mouth to remain partially open.
- They are mentally retarded.
- They have a prominent epicanthic fold in the corner of each eye; and typical flat face and round head.
- Usually, there is a wide gap between the first and the second digits on their feet.

a. Spindle fibres formation

Spindle fibres are microtubules that move chromosomes during cell division. They are found in eukaryotic cells. Spindle fibres move chromosomes during mitosis and meiosis to ensure that each daughter cell gets the correct number of chromosomes. The spindle apparatus consists of spindle fibres, motor proteins, chromosomes, and, in some cells, structures called asters (which are star-shaped structures form around each pair of centrioles during mitosis. They help to manipulate **chromosomes** during **cell division** to ensure that each **daughter cell** has the appropriate complement of chromosomes). In **animal cells**, spindle fibres are produced from cylindrical microtubules called **centrioles.** Centrioles are separated by asters to generate spindle fibres during the **cell cycle.** Centrioles are located in a region of the cell known as the **centrosome**.

Synapsis

In prophase I, homologous chromosomes become closely associated in synapsis. Synapsis includes the formation of an elaborate structure called the synaptonemal complex, consisting of homologous chromosomes paired closely along a lattice or zipper-like structure of proteins between them. The components of synaptonemal complex include a meiosis-specific form of cohesion that helps the two homologous chromosomes to be closely associated along their length.

Some events that occur along with synapsis are:

- The nuclear envelope breaks down.
- Two pairs of centrosome migrate to opposite poles.
- Spindle fibres formation occurs.

Bivalents

These arethe two homologous chromosomes attached at chiasmata. The homologous chromosomes consist of two sister chromatids each.

Chiasma formation and movement of chromosomes.

Chiasmata is the region of crossing over between two homologous chromosomes during prophase I of meiosis. At the chiasmata, homologous chromosomes exchange genes, allowing genetic information from both the paternal and maternal chromatids to be exchanged, and a recombination of paternal and maternal genes can be passed down to the progeny. This process is important in diploid organisms to ensure variation in the progeny.

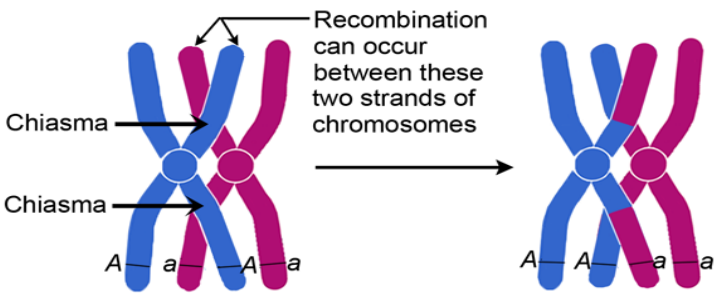


Figure 10.17: Process of crossing over

Process of chiasmata formation

At prophase I of meiosis, after the homologous chromosomes pair up in the process called synapsis, the non-sister chromatids overlap, forming an X-shape. They then exchange their alleles at the point of crossing over. The X-shape the homologous pair together until the cell progresses to anaphase.

Movement of chromosomes

In early Prophase I, chromosomes change their size and become short and thick. This is the first movement they make during meiosis

In late Prophase I, homologous chromosomes become fully shortened and thickened lie side by side (in a process called synapsis).

In metaphase I, Pairs of homologous chromosomes migrate to the equatorial plane of the cell. Each chromosome moves independently of all the others and the phenomenon is called independent assortment.

During anaphase I, the homologous chromosomes separate and begin to move to the opposite poles of the cell, pulled by the shrinking spindle fibers.

In telophase I, the movement of the homologous chromosomes to the poles is completed.

During Prophase II chromosomes again become thicker and shorter begin to move to the equatorial plane of the cell. Spindle fibers once again begin to form at the poles.

During metaphase II the chromosomes become aligned on the equatorial plane and spindle fibers attach to the centromeres.

In anaphase II, the centromeres divide, separating the sister chromatids, that move to the opposite poles due to the spindle fibers pulling.

In telophase II, the movement of the chromosomes to the poles is completed and the spindle disappears.

Application activity 10.3

1. Discuss possible functions of mitosis in living organisms.
2. Appreciate the role of cyclins in the process of cell cycle.
3. Describe the process of tumor formation.
4. Describe the three types of tumors.

Skills lab 10

Use microscope to compare the sizes of the similar cells in larger and small plants. For example, you might compare the leaf cells of grass to the leaf cells of a tree. Be sure you use the same magnification when comparing the sizes of the cells. Do the same from the small and large animal e.g. from frog and from human.

Are the cells of small plants larger or smaller than those of the larger plants? Are the cells of the small animal larger or smaller than those of larger animal? Make a general statement that compares the number and the sizes of cells in small organisms to those in large organisms.

End unit assessment 10

I. Choose whether the following statements are True (T) or False (F)

1. A typical cell spends most of its time in interphase.
2. Mitosis is a process where a single cell divides into three identical daughter cells.
3. Cytokinesis is a division of cytoplasm.
4. The process of mitosis is basically divided into 5 phases.
5. Meiosis is divided into three stages: Meiosis I, Meiosis II and Meiosis III.
6. The unrestrained, uncontrolled growth of cells in human beings results into a disease called cancer.
7. Cancer occurs due to failure in controlling cell division.
8. Proper separation of chromosomes during meiosis is not essential for the normal growth in humans.
9. The life span of blood cells ranges from less than one day to a few months.

II. Multiple Choice Questions

1. In telophase, the nuclear envelope re-forms around the..... set of haploid daughter chromosomes.

   (a) one (b) two (c) three (d) four

2. ..... is a condition in which the homologues or sister chromatids fail to separate properly during meiosis.

   (a) Disjunction (b) Non-disjunction (c) Down syndrome (d) None of these

3. Which of the event is correct in anaphase

   (a) Sister chromatids separate and give rise to daughter chromosomes.

   (b) Chromosomes are aligned at metaphase plate.

   (c) Cytokinesis starts occurring.

   (d) Chromosomes begin to uncoil.

4. One round of oogenesis produces

   (a) One egg (b) Two eggs (c) Three eggs (d) Four eggs

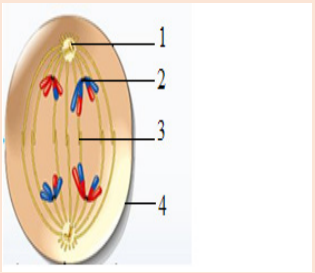
5. In mitosis, which of the following occurs?

   (a) Chiasmata formation (b) DNA replication (c) Synapsis (d) None of these

III. Long Answer Type Questions

1. Describe the main stages of cell cycle.
2. In your own words, explain what is meant by homologous pairs of chromosomes.
3. In your own words, describe the process of mitosis.
4. In your own words, describe the process of meiosis.
5. Outline the significance of mitosis in cell replacement and tissue repair by stem cells.
6. In your own words, explain how uncontrolled cell division can result in the formation of a tumor.
7. What is the need for reduction prior to fertilization in sexual reproduction?
8. In your own words, explain the importance of effective cell division.

9. Outline the role of meiosis in gametogenesis in humans and in the formation of pollen grain and embryo sacs in flowering plants.
10. Explain how crossing over and random assortment of homologous chromosomes during meiosis and random fusion of gametes at fertilization leads to genetic variation, including the expression of rare recessive alleles.
11. Analyze the following diagram and answer the questions below:



a) Identify the stage of cell division shown in the figure.

b) Label the structures marked as (1), (2), (3) and (4).

c) Which type of cell is involved in this division?

d) What will happen if the structure marked (3) is not formed?

12. How can you correlate the spread of HIV virus with the process of Mitosis? Knowing the viral disease and its spread.

UNIT 11:

AUTOTROPHIC NUTRITION

Key Unit competence: Describe the process of photosynthesis and explain the various environmental factors that influence the rate of photosynthesis.

Introductory Activity 11

Read the passage below and then answer questions that follow.

Vital process that makes up our life

The life is all about autotrophic nutrition. There is a number of reactions that take place during the autotrophic nutrition but all are simpled summarised by the saying inorganic matters are transformed into organic matters. And as said in physics, the energy is neither created nor lost but only transformed from one form to another, the process of autotrophic nutrition transforms light energy into chemical energy. But, opposite to most of metabolic reactions which produce harmful waste products, the waste product from photosynthesis is useful for life. So, no autotrophic nutrition, no life either.

1. Suggest a name to autotrophic nutrition
2. Which waste product from autotrophic nutrition is useful for life?
3. Do you agree that no autotrophic nutrition, no life either? Why?
4. The overall equation of autotrophic nutrition should be written as:
   $6\,CO_2 + 12\,H_2O \rightarrow C_6H_{12}O_6 + 6\,H_2O + 6O_2$ and not
   $6\,CO_2 + 6\,H_2O \rightarrow C_6H_{12}O_6 + 6\,O_2$. Prove it.
5. What requirements are missing on the above equation?
6. What factors can affect the rate of autotrophic nutrition.

11.1. Types of autotrophic nutrition

Activity 11.1

Do a research on different types of autotrophic nutrition and present them on manila paper.

Autotrophic nutrition is a process by which living organisms make their own food. This process is carried out by photoautotrophs like green plants, green algae and green bacteria; and chemoautotrophs. Living organisms which make their own food are called autotrophs, while others, including humans, which cannot make their own food but depend on autotrophs are called heterotrophs.

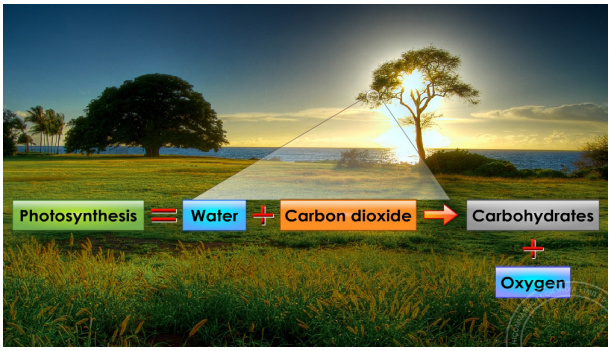


Figure 11.1: Photoautotrophism

There are two types of autotrophic nutrition such as chemoautotrophic and photoautotrophic nutrition.

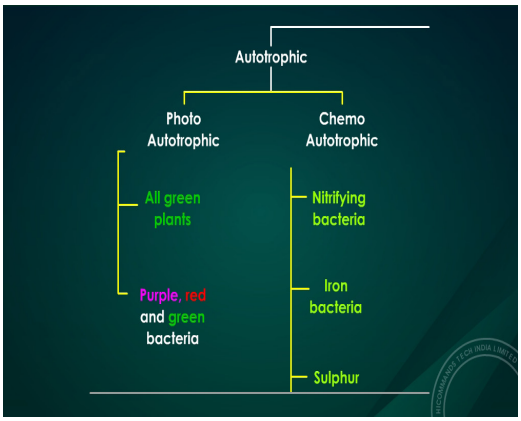


Figure 11.2: Types of autotrophic nutrition

Chemoautotrophic nutrition

It is an autotrophic nutrition where organisms (mainly bacteria) get energy from oxidation of chemicals, mainly inorganic substances like hydrogen sulphide and ammonia.

$$H_2S + O_2 \longrightarrow 2H_2O + 2S + E.$$

$$(NH_4)_2CO_3 + 3O_2 \longrightarrow 2HNO_2 + CO_2 + 3H_2O + E.$$

Photoautotrophic nutrition

It is an autotrophic nutrition where organisms get energy from sunlight and convert it into sugars. Green plants and some bacteria like green Sulphur bacteria can make their own food from simple inorganic substances by a process called photosynthesis. Photosynthesis is a process by which, autotrophs make their own food by using inorganic substances in presence of light energy and chlorophyll.

$$6CO_2 + 6H_2S \xrightarrow{\text{Light and bacteriochlorophylly}} C_6H_{12}O_6 + 6S$$ (Green Sulphur bacteria)

$$6CO_2 + 6H_2O + \xrightarrow{\text{Light and chlorophylly}} C_6H_{12}O_2 + 6O_2$$ (Green plants)

Application activity 11.1

1. Define Photosynthesis
2. Differentiate:
   a) Autotrophs and heterotrophs.
   b) Chemoautotrophs and photoautotrophs.
3. Animals' life depends on plants. Defend this statement by providing two convincing reasons.

11.2. Structure of the chloroplast and Adaptations for photosynthesis

Activity 11.2

With a compound microscope and well prepared slides of plant cell

- Observe the slides under the compound microscopes, and draw the structure of chloroplast.
- Search and relate the structure of chloroplast with the process of photosynthesis.

In eukaryotes photosynthesis takes place in chloroplasts. A chloroplast contains many sets of disc like sacs called thylakoids, which are arranged in stacks known as grana. Each granum looks like a stack of coins where each coin being a thylakoid. In the thylakoid, proteins are organized with the chlorophyll and other pigments into clusters known as photosystems. The

photosystems are the light-collecting units of the chloroplast.

The function of thylakoids is to hold the chlorophyll molecules in a suitable position for trapping the maximum amount of light. A typical chloroplast contains approximatively 60 grana, each consisting of about 50 thylakoids. The space outside the thylakoid membranes are made by watery matrix called stroma. The stroma contains enzymes responsible for photosynthesis.

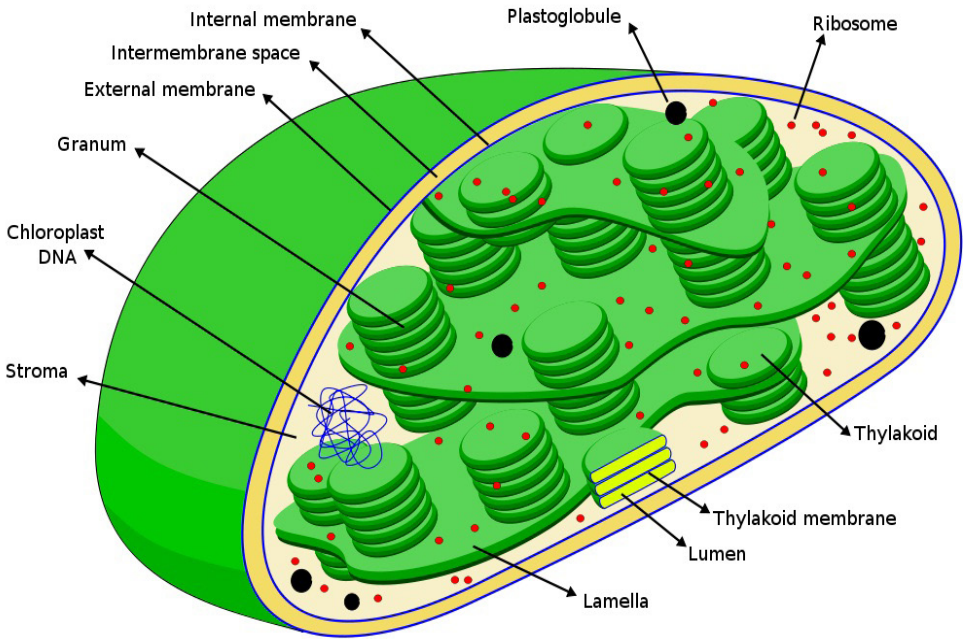


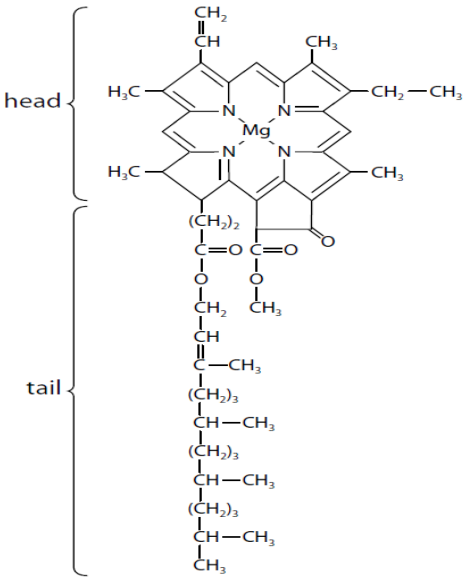
Figure 11.3: Structure of a chloroplast

Note: Photosynthetic prokaryotes have no chloroplasts, but thylakoids often occur as extensions of the plasma membrane and are arranged around the periphery of the prokaryotic cell.

The structure of chlorophyll

The chlorophyll molecule is made of atoms of Carbon and Nitrogen joined in a complex porphyrin ring containing an atom of Magnesium in the center of the ring. The chlorophyll also has long hydrophobic carbon tail of 20 carbon atoms (phytol) which hold it in the thylakoid membrane. In short, the chlorophyll consists of a porphyrin ring and a phytol tail.

The chlorophyll a differs from the chlorophyll b in that: the porphyrin of the chlorophyll a has the methyl group ($-CH_3$) as a functional group, which is replaced by an aldehyde group (-CHO) for chlorophyll b.



The structure of chlorophyll

The difference between the chlorophyll a and the chlorophyll b shifts the wavelength of light absorbed and reflected by chlorophyll b, so that the chlorophyll b is yellow-green, whereas the chlorophyll a is bright-green.

Adaptations for photosynthesis

By considering both external and internal structures of the leaf, we can recognize several adaptations for photosynthesis.

Adaptation of leaf for photosynthesis considering to its internal structure

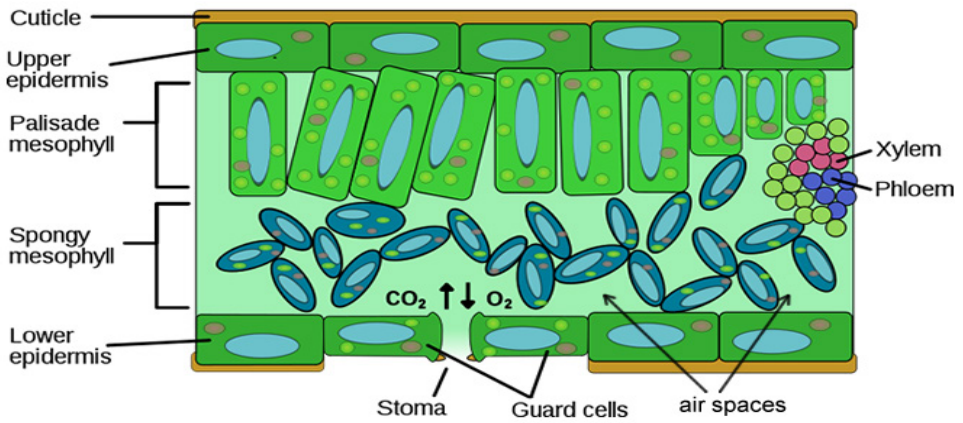


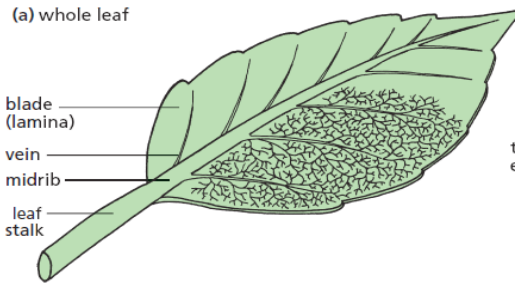
Figure 11.4: Internal structure of the leaf and its adaptations to photosynthesis

Parts of the leaf	Its adaptation for photosynthesis
Transparent cuticle	It allows to absorb much light. It prevent excessive water from the plant
Upper epidermis	It is made by thin and transparent cells for absorption of much light. They act as barriers to pathogens.
Palisade mesophyll	Its cells are packed with many chloroplasts, so it is the main site for photosynthesis
Spongy mesophyll	It is relatively less packed by chloroplasts, and it contains air spaces which allow gaseous exchange: CO_2 in cells, and O_2 from cells during photosynthesis.
Vascular bundle	It consists of leaf veins: Xylem which bring water and minerals to leaves, phloem which transport sugars and amino acids away to other parts.
Lower epidermis	It is the site for gaseous exchange in and out of the leaf, and it prevent excessive water loss through stomata.
Stomata	Surrounded by guard cells, which allow stomata to open: water moves out during transpiration, CO_2 diffuses in and Oxygen diffuses out during photosynthesis.

Note: When stomata are opened, the rate of photosynthesis may be 10 to 20 times as fast as the maximum rate of respiration. If the stomata are closed, photosynthesis still can continue, using CO2 produced during cell respiration. The equilibrium can be reached between photosynthesis and cell respiration.

Photosynthesis uses CO_2 from respiration, and respiration uses Oxygen from photosynthesis. However, the rate of photosynthesis under these circumstances will be much slower than when an external source of CO_2 is available. The stomata cannot remain closed indefinitely, they have to be open in order to maintain transpiration of the plant.

Adaptation of leaf for photosynthesis considering its external structure



- Leaves are thin and flat, this facilitate absorption of the maximum amount of light.

- The cuticle is transparent to allow absorption of light into tissues.
- Presence of a waxy substance on the cuticle to prevent excessive water loss from photosynthetic tissues.
- Presence of the midrib and veins containing vascular tissues like: the Xylem which brings water and minerals from soil to photosynthetic tissues, and Phloem which carry away manufactured organic food from photosynthetic tissues to other parts (translocation).
- Having the leaf stalk which holds the lamina in a good position to receive the maximum amount of the light.

Application activity 11.2

1. Describe the structure of a chloroplast.
2. Explain adaptations of both thylakoid and stroma for their functions.
3. Relate the internal structure of the leaf with the process of photosynthesis

11.3. Absorption, action spectra and other carbon dioxide fixation pathways

Activity 11.3

1. Is it possible for a plant to grow without solar light? Justify your answer.
2. Why do plants have different colours? Justify each color of plant observed in environment.

In addition to water and CO_2, photosynthesis requires light and chlorophyll. The chlorophyll pigment is found in the chloroplasts. The light that our eyes perceive as white light is a mixture of different wavelengths. Most of them are visible to our eyes and make up the visible spectrum. Our eyes see different wavelengths of visible spectrum as different colors (violet, blue, green, yellow, orange and red) except indigo which is not visible to our eyes. Plants absorb the light energy by using molecules called pigments such as: chlorophyll a, chlorophyll b, carotene (orange), xanthophyll (yellow) and phaeophytin (grey) but chlorophyll a is the principle pigment in photosynthesis.

The chlorophyll absorbs light very well in blue-violet and red regions of visible spectrum. However, chlorophyll does not absorb well the green light, instead it allows the green light to be reflected. That is why young leaves and other parts of the plants containing large amount of chlorophyll appear green.

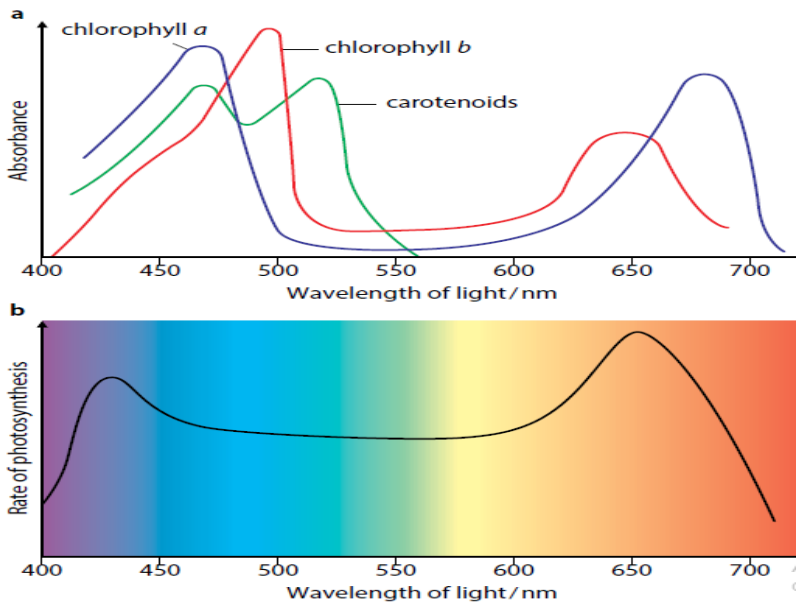


Figure 10.5: Light absorption by leaf pigment

The chlorophyll a as a principle and abundant pigment, it is directly involved in light reactions of photosynthesis. Other pigments (chlorophyll b, carotene, xanthophyll and phaeophytin) are accessory pigments. They absorb light colours that chlorophyll a cannot absorb, and this enables plants to capture more energy from light.

The amount of energy that the pigment can absorb from the light, depends on its intensity and its wavelengths. So, the greater the intensity of light, the greater amount of energy will be absorbed by the pigment in a given time.

The process of photosynthesis occurs through two main stages such as:

- The light-dependent reactions: which take place in thylakoids, and
- The light-independent reactions (Calvin cycle): which take place in stroma.

a) The light-dependent reactions

They require light energy and occur in thylakoids. They produce Oxygen gas and convert ADP and $NADP^+$ into ATP and NADPH.

The light-dependent reactions involve the following steps:

Step 1: Photosynthesis begins when the chlorophyll a in photosystem II absorbs light at different wavelengths of light.

- When the light energy hits the chlorophyll a, the light energy is absorbed by its electrons, by raising their energy level.

- These electrons with high potential energy (electrons with sufficient quantum energy) are passed to the electron-transport chain.
- Excited electrons are taken up by an electron acceptor ($NADP^+$: oxidized Nicotinamide Adenine Dinucleotide Phosphate), and pass along electron transfer chain from photosystem II to the photosystem I. (Note: The photosystems are the light-collecting units of the chloroplast).

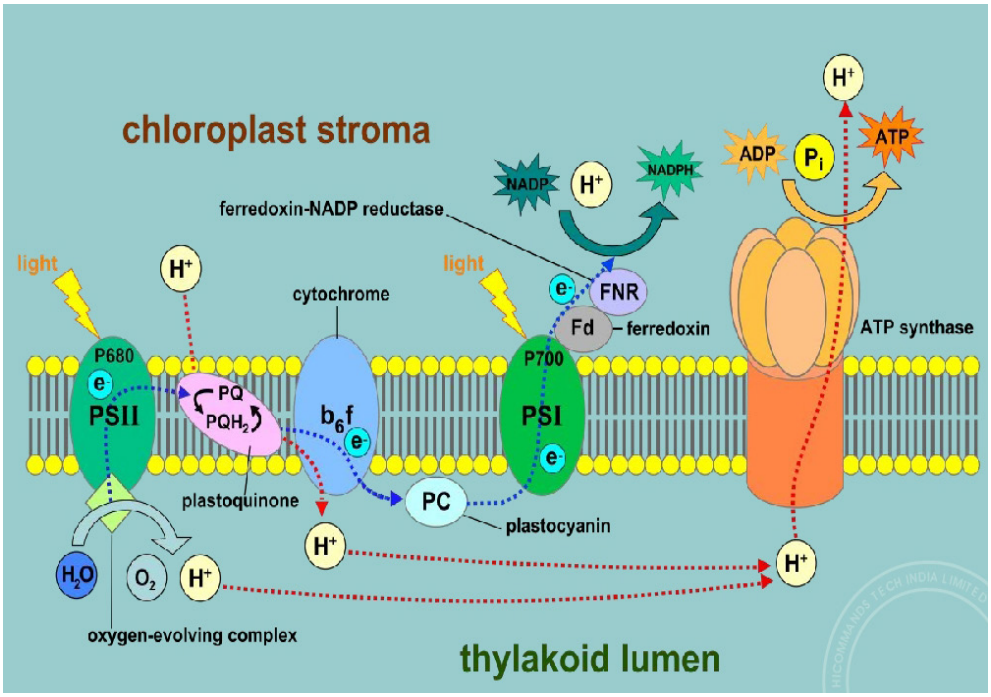


Figure11.6. Electron transport chain

Step 2: Enzymes in thylakoids and light absorbed by photosystem II are used to break down a water molecule into energized electrons, hydrogen ions H^+, and Oxygen.

$$H_2O \xrightarrow{Enzymes\ and\ light} 2H^+ + \frac{1}{2}O_2 + 2e^-$$

- Oxygen produced is released to be used by living things in respiration.
- Electrons and H^+ from photolysis of water are used to reduce NADP+ to NADPH (Reduced Nicotinamide Adenine Dinucleotide Phosphate).
- The light-dependent reactions also allow generation of ATP (Adenosine Triphosphate) by adding inorganic phosphate to ADP^+ (Adenosine Diphosphate):

$$ADP + P_i \xrightarrow{Energy} ATP$$

Generally, the light-dependent reactions use light energy, ADP, Pi, NADP+ and water to produce ATP, NADPH and Oxygen. Or simply:

$$12H_2O + 12NADP^+ + 18ADP^+ + 18P_i \xrightarrow{\text{Light and chlorophyll}} 6O_2 + 12NADPH_2 + 18ATP$$

Both ATP and NADPH are energy carriers which provide energy to sugars (energy containing sugars) in Light-independent reactions.

Step 3: The fixation of Pi to ADP+ to form ATP is called photophosphorylation.

Photophosphorylation can be done into two processes: cyclic photophosphorylation, and non-cyclic photophosphorylation.

i) Cyclic photophosphorylation

It involves only photosystem I and not photosystem II. There is no production of NADPH and no release of Oxygen. When the light hits the chlorophyll in PSI, the light-excited electron leaves the molecule.

This light-excited electron is taken up by an electron acceptor which passes it along an electron transfer chain (a series of electron carriers) until it returns to the chlorophyll molecule that it left (cyclic process). As an excited electron moves along an electron transfer chain, it loses energy which will be used for the synthesis of ATP from ADP+ and inorganic phosphate in the process called chemiosmosis. Electron carriers can vary, but the principle include the cytochromes.

ii) Non-cyclic photophosphorylation

It is the main route of ATP synthesis. It is done in the following steps:

- When the photosystem II (in chlorophyll) absorbs light, an electron is excited to a higher energy level and captured by the primary electron acceptor.
- Enzymes extract electrons from a water molecule replacing each electron that the chlorophyll molecule lost when absorbed light energy. This reaction dissociates a water molecule into hydrogen ions ($2H^+$) and Oxygen which is released for animals' respiration.
- Excited electron moves from the primary electron acceptor of photosystem II to photosystem I, via an electron transport chain.
- When excited electron moves from the primary electron acceptor of photosystem II to photosystem I, via an electron transport chain its energy level lowers. The energy removed is used to synthesize ATP from ADP and P_i in a process called: Non-cyclic phosphorylation.
- The hydrogen ions ($2H^+$) produced from dissociation of water molecule combines with $NADP^+$ to form $NADPH_2$.

- Both ATP and $NADPH_2$ will be used in the light-independent reactions (Calvin cycle) for synthesis of sugars.

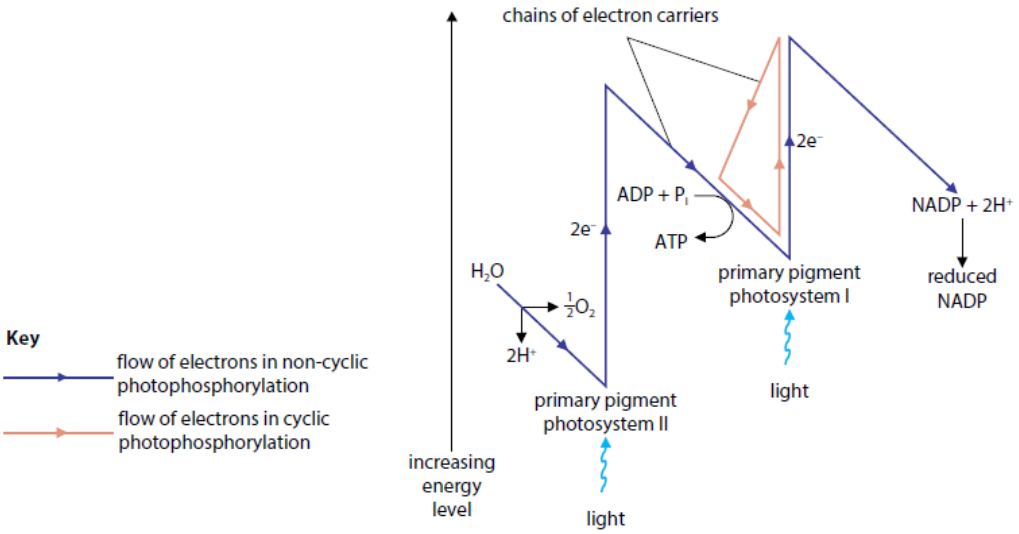


Figure 11. 7: Chain of electron carriers

The significance of the cyclic phosphorylation

Non-cyclic photophosphorylation produces ATP and NADPH in equal quantities, but the Calvin cycle consumes more ATP than NADPH. The concentration of NADPH in a chloroplast may determine which pathway (cyclic versus non-cyclic) electrons pass through.

If a chloroplast runs low on ATP for the Calvin cycle, NADPH will accumulate as the cycle slows down. The rise of NADPH may stimulate a shift from non-cyclic (which produces ATP only) to cyclic electron pathway until ATP supply catches with the demand.

Table 11.4: Comparison between Non-cyclic and cyclic photophosphorylation

	Cyclic photophosphorylation	Non-cyclic photophosphorylation
Pathway of excited electrons	Cyclic pathway	Non-cyclic pathway
Source of electrons (1st electron acceptor)	Photosystem I	water
Destination of electron (Last electron acceptor)	Photosystem I	$NADP^+$
Photosystem involved	Photosystem I only	Photosystem II and Photosystem I
Products	ATP only	ATP, NADPH and Oxygen

b) The light-independent reactions (Calvin cycle)

The light-independent reactions occur in stroma, and consist of reducing CO_2 into sugars by using ATP and NADPH both coming from light-dependent reactions in thylakoids. The Calvin cycle involves three main stages such as:

- Carbon fixation in form of CO_2.
- Carbon reduction from CO_2 to glucose.
- Regeneration of RuBP.

Step 1: Carbon fixation (Carboxylation) in form of CO_2

The Calvin cycle begins with a 5-Carbon sugar phosphate called Riburose-1, 5 biphosphate (RuBP) which fixes the CO_2 from air. This reaction is catalyzed by an enzyme called RuBPcarboxylase-oxygenase (RUBISCO), which makes up about 30% of the total protein of the leaf, so it is probably one of the most common proteins on the Earth.

The combination of RuBP and CO_2 results in a theoretic 6-carbon compound which is highly unstable. It immediately splits into two molecules of 3-carbon known as phosphoglyceric acid (PGA) or glycerate 3-phosphate, or 3-phosphoglycelate.

Step 2: Carbon reduction from CO_2 to glucose

With energy from ATP and reducing power from NADPH, the phosphoglyceric acid is reduced into 3carbon molecules known as glyceraldehyde-3-phosphate or phosphoglyceraldehyde (PGAL).

Each molecule of PGA receives an additional phosphate group from ATP, becoming 1, 3-biphosphoglycerate, and a pair of electrons and H^+ from NADPH reduces the carboxyl group of 3-phosphoglycerate to the aldehyde group of PGAL which stores more potential energy.

ATP gives one phosphate group becoming ADP^+, and NADPH gives H^+ and electrons to become $NADP^+$. Both ADP^+, and $NADP^+$ will be used again in light-dependent reactions.

With 6 turns of Calvin cycle, the plant cell fixes $6CO_2$ molecules which are used to synthesize 2 molecules of PGAL which leave the cycle and combine to make one molecule of glucose or fructose. This glucose can be converted into:

- Sucrose: When Oxygen combined with fructose. It is a form by which carbohydrates are transported in plants.
- Polysaccharides like starch for energy storage, and cellulose for structural support.

- Amino acids when combined with nitrates,
- Nucleic acids when Oxygen combined with phosphates, and
- Lipids.

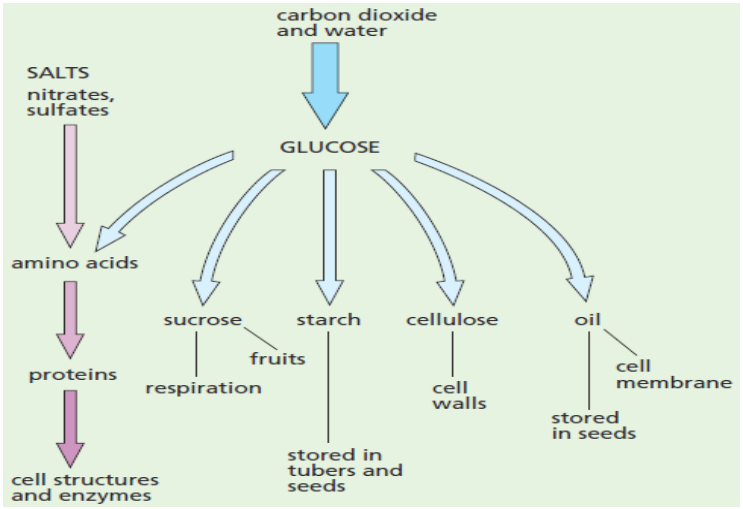


Figure 11.8: Conversion of glucose into other organic substances

Step 3: Regeneration of RuBP

The remaining ten 3-carbon molecules (PGAL) are converted back into six 5-carbon molecules, ready to fix other CO_2 molecules for the next cycle. The light-independent reactions can be summarized as:

$$12NADPH + 18ATP + 6CO_2 \rightarrow C_2H_{12}O_6 + 12NADP^+ + 18ADP^+ + 18P_i + 6H_2O$$

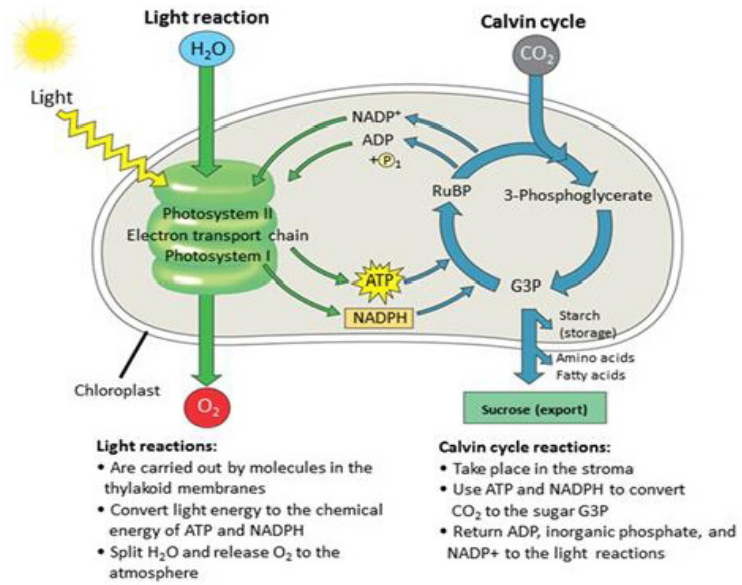


Figure 11.9: Light reaction and Calvin cycle

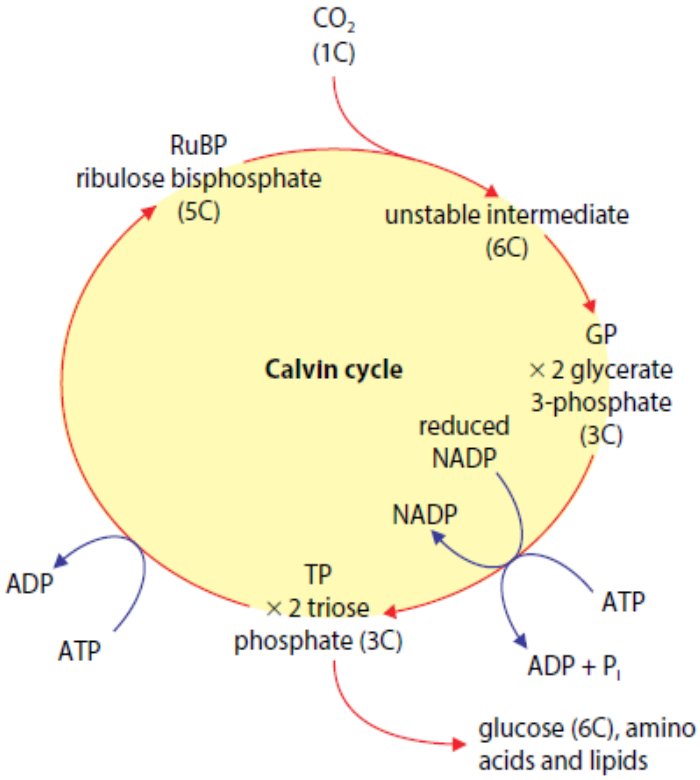


Figure 11.10: Calvin cycle

Other carbon dioxide fixation pathways (C4 CAM)

The most common pathway combines one molecule of CO_2 with a 5-carbon sugar called ribulose biphosphate (RuBP). The enzyme which catalyzes this reaction (nicknamed "Rubisco") is the most abundant enzyme on earth! The resulting 6-carbon molecule is unstable, so it immediately splits into two 3-carbon molecules. The 3 carbon compound which is the first stable molecule of this pathway gives this largest group of plants the name "**C-3 plants**".

Dry air, hot temperatures, and bright sunlight slow the C-3 pathway for carbon fixation. This is because **stomata**, which normally allow CO_2 to enter and O_2 to leave, must close to prevent loss of water vapor. Closed stomata lead to a shortage of CO_2. Two alternative pathways for carbon fixation demonstrate biochemical adaptations to differing environments. Plants such as **corn** solve the problem by using a separate compartment to fix CO_2.

Here CO_2 combines with a 3-carbon molecule, resulting in a 4-carbon molecule. Because the first stable organic molecule has four carbons, this adaptation has the name **C-4**. Shuttled away from the initial fixation site, the

4-carbon molecule is actually broken back down into CO_2, and when enough accumulates, Rubisco fixes it a second time!

In some temperate plants such as wheat, rice, potato and bean only Calvin cycle occurs. Such plants are called C-3 plants. While in some other plants dual carboxylation takes place: (1) carboxylation of phosphoenol pyruvate (PEP) and (2) carboxylation of RuBP. Such plants are called C-4 plants e.g. maize, sugar cane and sorghum. In these, the first product formed during carbon dioxide fixation is a four carbon compound oxalo acetic acid (OAA). C-4 plants have special type of leaf anatomy called **Kranz** Anatomy. They have special large cells around vascular bundles called **bundle sheath cells**. These are characterized by having large number of chloroplasts, thick walls and no intercellular spaces. The shape, size and arrangement of thylakoids in chloroplasts are also different in bundle sheath cell as compared to mesophyll cell chloroplasts.

The pathway followed by C-4 plants is called **C-4 cycle** or **Hatch and Slack pathway**. This was discovered by Hatch and Slack in sugar cane. The primary CO_2 acceptor is a 3-carbon molecule phosphoenol pyruvate (PEP). The reaction is catalyzed by **PEP carboxylase** or **PEP case** in mesophyll cell chloroplast. It forms 4-carbon compounds like OAA, malic acid or aspartate, which are transported to the bundle sheath cells. In bundle sheath cells, these acids are broken down to release CO_2 and 3-carbon molecule. The 3-carbon molecule is transported back to mesophyll cells and converted to PEP again, while CO_2 enters into C-3 cycle to form sugars. C-4 plants are more efficient than C-3 plants as in C-4 plants, photosynthesis can occur at low concentration CO_2 and photorespiration is negligible or absent.

Cacti and succulent (water-storing) plants such as the jade plant avoid water loss by fixing CO_2 only at night. These plants close their stomata during the day and open them only in the cooler and more humid nighttime hours. Leaf structure differs slightly from that of C-4 plants, but the fixation pathways are similar. The family of plants in which this pathway was discovered gives the pathway its name, **Crassulacean Acid Metabolism**, or **CAM**. All carbon fixation pathways lead to the Calvin cycle to build sugar.

The CAM pathway is similar to the C4 pathway in that carbon dioxide is first incorporated into organic intermediates before it enters the Calvin cycle. The difference is that in C4 plants, the initial steps of carbon fixation are separated structurally from the Calvin cycle whereas in CAM plants, the two steps occur at separate times.

The CAM pathway and the C4 pathway compared:

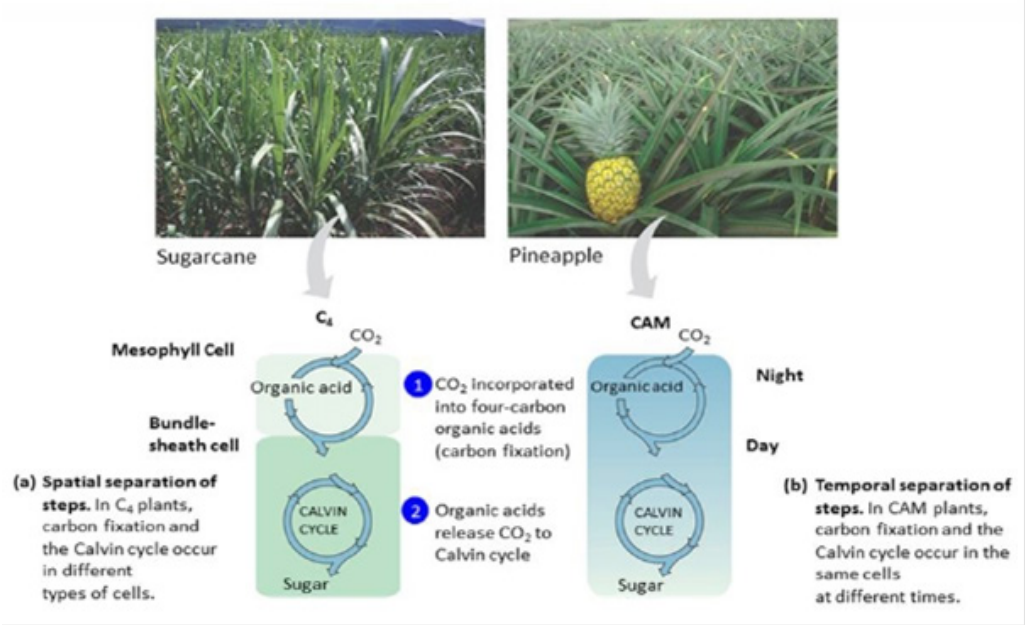


Figure 11.11: The C4 and CAM pathways compared

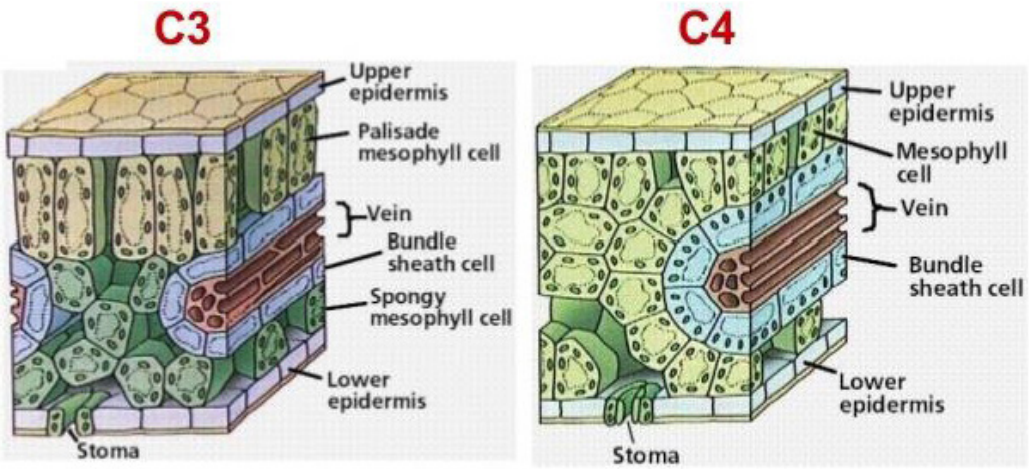


Figure 11.12: C3 leaf anatomy and the C4 leaf anatomy compared

Table 11.4: Comparison between C3 and C4 plants

	C3 Plants	C4 Plants
Perform the Calvin cycle	Yes	Yes
Primary CO_2 acceptor	RuBP	PEP
First product of CO_2 fixation	PGA	Oxaloacetate
CO2 fixing enzyme	RUBISCO	PEP carboxylase (This enzyme adds CO_2 to phosphoenolpyruvate (PEP), forming the four-carbon product oxaloacetate
Affinity of carboxylase for CO_2	Moderate	High
Photorespiration	Extensive	Minimal
Classes of chloroplasts	One	Two
Leaf anatomy, photosynthetic cells	Mesophyll	Mesophyll and Bundle sheath

Photorespiration

In most plants, initial fixation of carbon occurs via Rubisco, the Calvin cycle enzyme that adds CO_2 to ribulose biphosphate. Such plants are called C3 plants because the first organic product is a three carbon organic compound, PGA. These plants produce less food when their stomata close on hot and dry days.

The declining level of CO_2 in the leaf starves the Calvin cycle. Making matter worse, Rubisco can accept O_2 in place of CO_2. As O2 concentration overtakes CO_2 concentration within the air space, Rubisco adds O_2 instead of CO_2. The product splits and one piece, a two-carbon compound is exported from the chloroplast. Mitochondria then break the two-carbon molecule into CO_2.

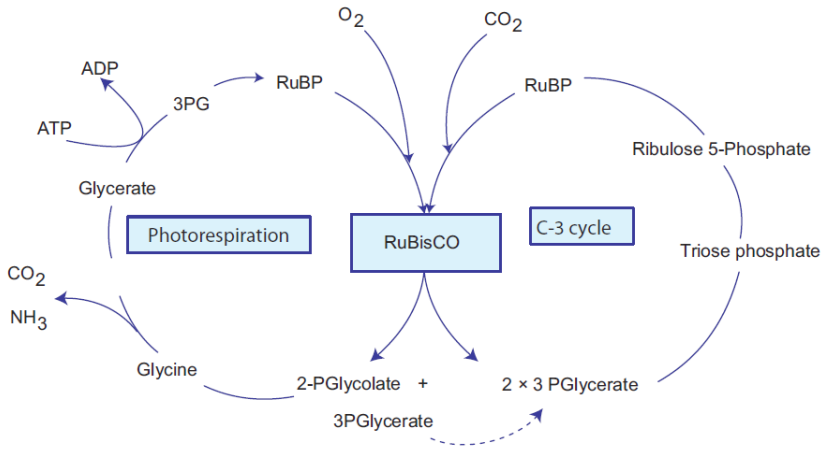


Figure 11.13: Photorespiration and C-3 cycle

The process is called photorespiration because it occurs in presence of light (photo) and consumes O_2 (respiration). However, unlike normal cellular respiration, photorespiration generates no ATP, and unlike photosynthesis, photorespiration generates no food. In fact, photorespiration decreases photosynthetic output by using material from the Calvin cycle.

Application activity 11.3

1. Why are light and chlorophyll needed for photosynthesis?
2. Describe the relationship between the chlorophyll and the color of plants.
3. How well would a plant grow under pure yellow light? Explain your answer.
4. Appreciate the presence of accessory pigments in leaves for the process of photosynthesis.
5. differentiate the light-dependent stage and light-independent stage of photosynthesis.
6. Distinguish the cyclic and non-cyclic photophosphorylation.
7. Explain the stages of the Calvin cycle.
8. Appreciate the significance of the cyclic phosphorylation.
9. In a tabular form, compare the process of photosynthesis in C_3 and C_4 plants.
10. What do you understand by photorespiration?

11.4. Rate of photosynthesis: limiting factors of photosynthesis

Activity 11.4

By using prior knowledge, what do you think should influence the photosynthesis process in plants?

The photosynthesis rate varies with the species but also varies within individuals for a same species; this varies under the influence of certain external factors which are: the temperature, CO2 concentration in the atmosphere, light intensity and soil humidity.

a. Temperature

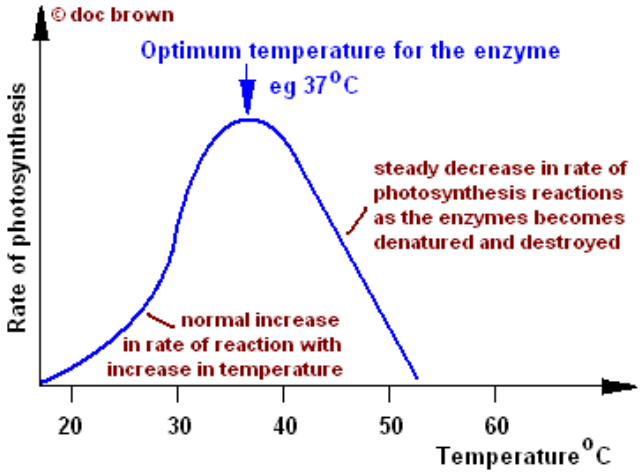


Figure 11.14: Effect of temperature on the rate of photosynthesis

Photosynthesis is low at 0˚C, certain alpine plants do photosynthesis at has less -15˚C.

Photosynthesis presents an optimum towards 35 to 40 ˚C, then it decreases and is cancelled at around 50˚C.

b. CO_2 concentration in the atmosphere

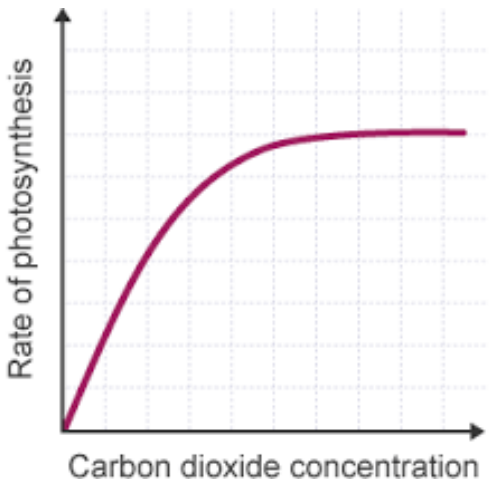


Figure 11.15: Effect of carbon dioxide on the rate of photosynthesis

The photosynthetic rate is zero in place lacking CO_2, it increases with the increase concentration of CO_2 in the atmosphere and reaches an optimum ranging between 5 and 8%CO_2 concentration.

11.4.3. Light intensity

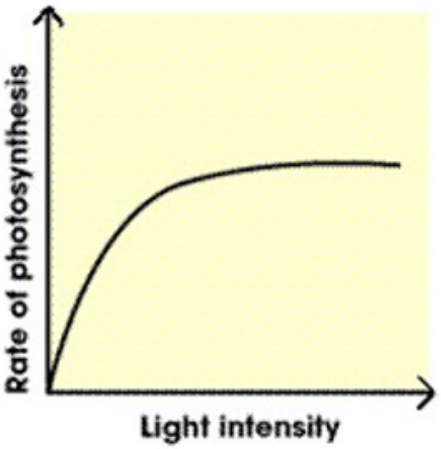


Figure 11.16: Effect of light intensity on the rate of photosynthesis

The photosynthesis rate is low during night, it increases when the light intensity increases but the optimum varies according to the plants.

11.4.4. Availability of water for the plant

The photosynthesis rate is low when the soil is dry, it increases when the content of water increases for the terrestrial plants, and for the aquatic plants it remains constant as long as they are fixed in water.

Note: The limiting factors work together to influence the rate of photosynthesis

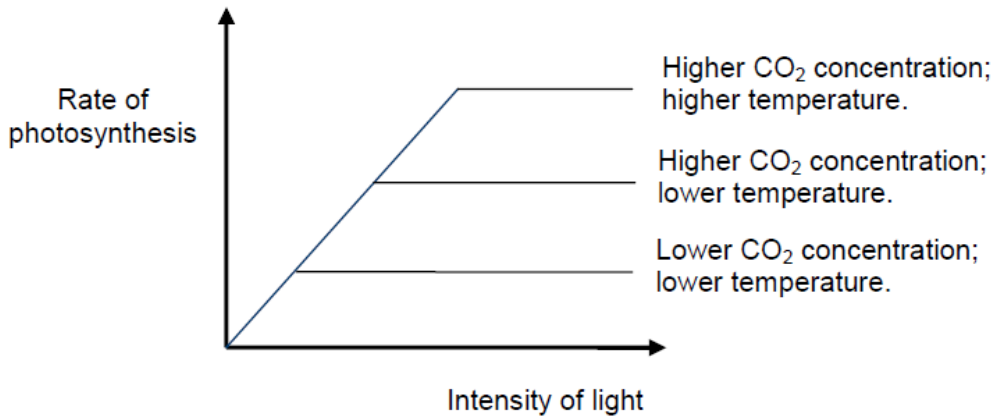


Figure 11.17: Effect of combined limiting factors on the rate of photosynthesis

Application activity 11.4

1. Use the graphs to explain how the limiting factors below may influence the rate of photosynthesis:

   a) Temperature

   b) Light intensity

   c) Concentration of CO_2 in air.

2. A student who studies Biology talked to his Biology group members that:

   a) "In Rwanda, the rate of photosynthesis is generally lower at 5:30 AM than it is at 12:30 PM, during a sunny day". Defend him by providing two convincing reasons.

   He also said that:

   b) "The rate of photosynthesis is generally higher in Rwanda during the sunny day than in Sahara Desert". Defend him with a convincing reason.

11.5. Importance of autotrophic nutrition and Tests for starch in terrestrial plants and for oxygen in aquatic plants

Activity 11.5

Describe the following topic: Animals depend on Plants, rather than plants depend on animals

A. Importance of autotrophic nutrition

Autotrophic nutrition is a process by which living organisms (autotrophs: photoautotrophs and chemoautotrophs) make their own food. The autotrophism is very essential as it allows production of Oxygen and food for not only themselves but also for heterotrophs. The roles of autotrophic nutrition include:

- **Independence of green plants from other living organisms to the nutrition point of view**

  This importance relates to their capacity for synthesizing organic molecules from glucose produced by CO_2 and water, this completely make them independents of the other living organisms to the nutrition point of view.

- **Energy storage**

  The autotrophs like green plants, by the process of photosynthesis synthesize certain substances like: the cellulose, starch… which are variables sources of energy.

- **Production of O_2 for the living organisms' respiration**

  The oxygen produced by the photosynthesis is necessary for the living organisms' respiration. Thus without photosynthesis, no oxygen; without oxygen no respiration; without respiration no life on Earth.

- **Cleaning the atmosphere**

  Photoautotrophs absorb carbon dioxide from surrounding air, and release Oxygen (produced by photosynthesis) in atmosphere.

- **Formation of Ozone layer**

  Ozone layer is a thick layer in the atmosphere which is formed Ozone molecule (O_3). Oxygen atoms which make ozone molecule are produced by photosynthesis. Ozone layer protects the Earth from high solar radiations, and this allows the existence of the life on the Earth.

  Synthesis of the organic substances: food for the heterotrophs (animal and mushrooms): The organic substances produced by photosynthesis are the food for the heterotrophs which are unable to synthesize these substances by their own means.

B. Tests for starch in terrestrial plants

The process of photosynthesis results in production of Oxygen and Organic substances like simple sugars, double sugars starch, amino acids among other. It is evident that, if someone need to be sure that photosynthesis has occurred, he can simple test for the presence of some of those substances produced by photosynthesis.

Materials and substances needed to test for the presence of starch in terrestrial plants

Bunsen burner, tripod stand, wire gauze,forceps, the petri dish, test tube holder, 250 cm^3 beaker, dropping pipette, boiling tube, a water bath, leaf to be tested (hibiscus leaves are excellent),90% ethanol,cold water, iodine/ potassium iodide solution.

The process of testing for starch in a leaf

1. Prepare a water bath; heat the water to boiling point and keep it boiling.
2. Detach a leaf from a dicotyledonous plant that has been exposed to sunlight for at least three hours.

3. With forceps, hold the leaf in boiling water for about 10 seconds. This will kill the cells, stop all chemical reactions and allow alcohol and iodine to penetrate the leaf more easily, and will removes the waxy cuticle which prevents entry of iodine/potassium iodide solution, will also denatures enzymes, particularly those which convert starch to glucose e.g. diastase. Boiling arrests all chemical reactions, since enzymes which catalyse the reactions are denatured. Denatured enzymes have altered or destroyed active sites due to heat, pH, and ionic concentration.
4. Extinguish the Bunsen burner.
5. Half fill a test tube with ethanol (ethyl alcohol).
6. With forceps, place the leaf in the alcohol.
7. Place the test tube containing ethanol and the leaf into the hot water. The ethanol will boil and dissolve out the chlorophyll in the leaf. When the leaf is nearly colorless, remove it from the ethanol with forceps and wash it in cold water. This will soften the leaf by replacing water removed by the ethanol.
8. Place the softened leaf in a Petri dish and, using a dropping pipette, cover it with iodine solution. Leave for 3 minutes.
9. If the colour of the leave turns blue-black, it means that it contains starch, and therefore photosynthesis has occurred in the leave. If no color change (Iodine Solution remains brown), starch is not present.

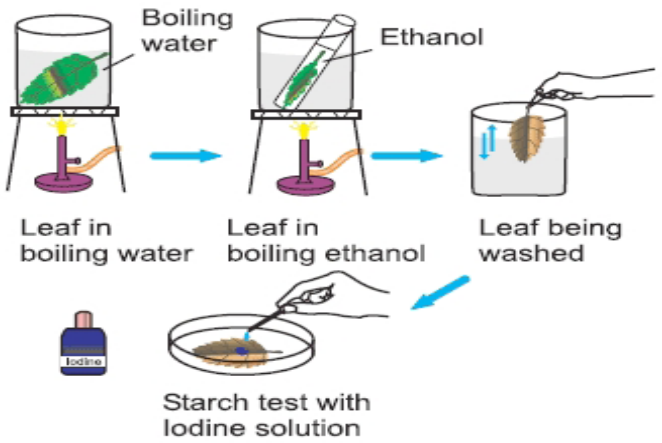


Figure 11.18: The process of testing for starch in terrestrial plants

We can also measure the process of photosynthesis via the production of oxygen, or by using the Audus apparatus to measure the amount of gas evolved over a period of time. Oxygen can be measured by counting bubbles evolved from pondweed. If you illuminate a *plant* (Presume you are using an aquatic/water *plant*) it will *produce oxygen*. You can put the *plant* under a

funnel and collect the bubbles in an upturned *test tube*. If the gas is collected fresh, it will relight a glowing splint. You can also use *Elodea*, but we find *Cabomba* more reliable. Put the weed in a solution of $NaHCO_3$ solution. You can then investigate the amount of gas produced at different distances from a lamp.

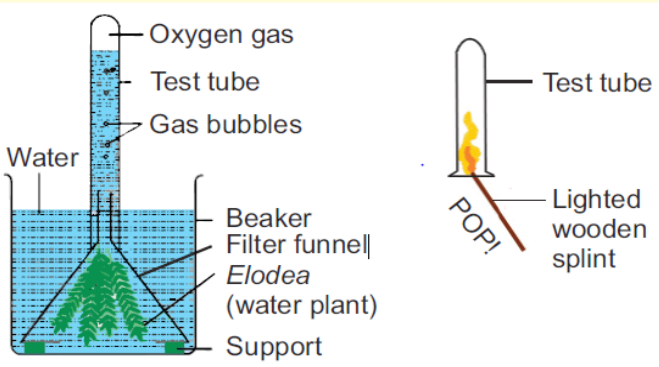


Figure 11.19: Test for oxygen in aquatic plants.

Application activity 11.5

1. Which products of photosynthesis may be present but not revealed by the iodine test?
2. Write down the steps to follow during testing for starch in terrestrial plant.
3. Appreciate the necessity of boiling the leaf to be used in testing for starch in plants.
4. Summarize the process of testing for Oxygen in aquatic plants.
5. Explain why the leaf which is nearly colorless, should first be washed in cold water.
6. Summarize the importances of photosynthesis to the living organisms.

Skills lab 11

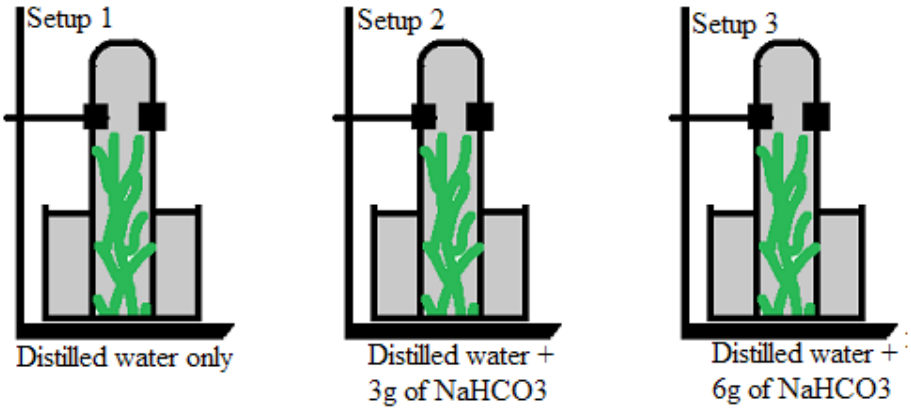
Evidence of photosynthesis

Materials: Large clear plastic cup, sodium bicarbonate solution (baking soda), aquatic filamentous green algae, large test tube, retor stand.

Procedure:

- Fill three large clear plastic cups with water and 3g or 6g of sodium bicarbonate as shown below:

- Place aquatic green algae in three large test tubes. Fill the tube with water or sodium bicarbonate solution in the cup. Caution: handle the test tube carefully.
- Hold your thumb over the mouth of the test tube. Turn the tube over, and lower it to the bottom of the cup. Make sure there is no air trapped in the tube.
- Fix each setup in a stand.



- Place your setups in bright light.
- After at least 20 minutes, record your observations and draw a conclusion.

End unit assessment 11

I. Choose whether the given statements are True (T) or False (F)

1. Organisms that are heterotrophic can make their own food.
2. Photosynthesis has two stages - light reaction and dark reaction.
3. CAM cycle includes triple carboxylation.
4. A pigment is a material that changes colour of reflected or transmitted light.
5. Within leaves, chloroplasts are responsible for respiration.

II. Multiple Choice Questions

1. Green plants does not require which of the following for photosynthesis?

   (a) Sunlight (b) CO_2 (c) O_2 (d) Water

2. C-4 cycle occurs in

(a) Wheat (b) Rice (c) Sugar cane (d) All of the above

3. Autotrophs are commonly called producers because they

(a) Produce young plants

(b) Produce CO_2 from light energy

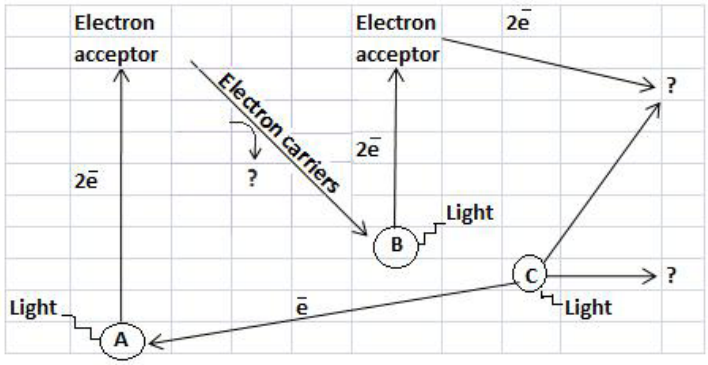
(c) Produce sugars from chemical energy

(d) Produce water from light energy

III. Long Answer Type Questions

1. State and explain the types of autotrophic nutrition.
2. Relate the structure of a chloroplast with its function and structure of leaf in photosynthesis.
3. State the pigments involved in light absorption.
4. Outline the three main stages of Calvin cycle.
5. Compare anatomy of C4 and CAM plants.
6. Differentiate between C4, CAM and C3 plants during carbon dioxide fixation.
7. Investigate the effect of light intensity on the rate of photosynthesis.

The chart below shows the sequence of events that takes place in the light dependent reactions.

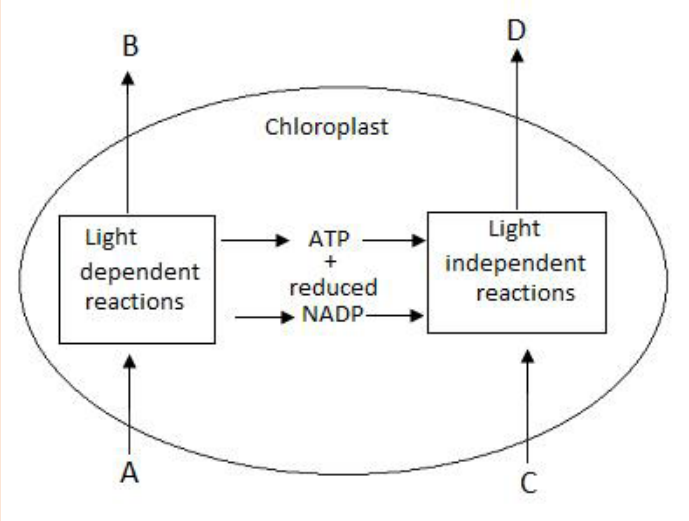


a) Identify the point A and B

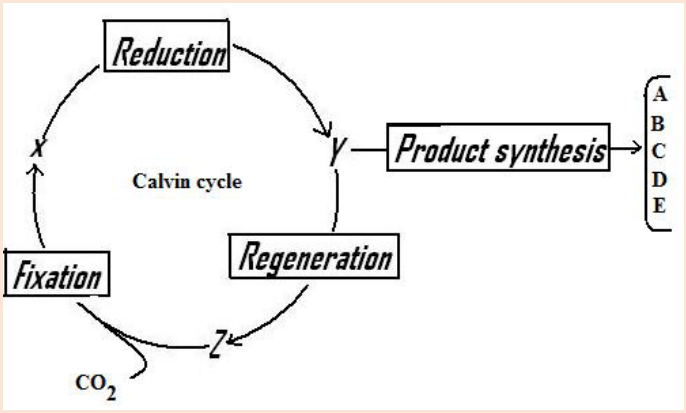
b) What process is taking place at C?

c) What are the products of the light dependent reaction? (They are indicated by? on the diagram).

8. The diagram below summarizes the movement of materials into and out of chloroplast. Identify the substances moved, indicated by labels A-D.



9. The diagram below represents the *CALVIN'S Cycle* also known as the light independent reactions of the photosynthesis.



a). Name substances X, Y and Z.

b). (i) State any two steps of the Calvin cycle which are endothermic.

(ii) State the source of the energy consumed in these two steps.

c). Why is it said that the reduction is the most important step of the Calvin cycle?

d). Name any two products represented by A, B, C, D or E.

UNIT 12:

THE CHEMICAL BASIS OF LIFE

Key Unit competence: Explain the use of biological molecules in living organism.

Introductory Activity 12

Analyze all foods in the figure above and answer the questions below:

a). Among the foods observed in figure, plan a list for your menu for breakfast, lunch and supper. Justify your choices.

b). Are there foods you missed on the list which you prefer to eat? Why?

c). List the nutrients our body gains from different food and justify why all are needed.

12.1. Biological molecules

Activity 12.1

1. Discuss chemical elements, sub-units of different types of carbohydrates, lipids and proteins.
2. What do you know about the function of water to living organisms?

12.1.1. The chemical elements that make up carbohydrates, lipids and proteins

a. Carbohydrates

Carbohydrates comprise a large group of organic compounds which contain carbon, hydrogen and oxygen. The word carbohydrate indicates that these organic compounds are hydrates of carbon. Their general formula is

$C_x(H_2O)_y$

.In carbohydrates the ration hydrogen-oxygen is usually 2:1.

Carbohydrates are divided into three groups including the monosaccharide (single sugars), disaccharides (double sugars) and polysaccharides (many sugars). The most common monosaccharide of carbohydrates is glucose with molecular formula $C_6H_{12}O_6$.

Lipids

Lipids are a broad group of naturally occurring molecules which include waxes, sterols, fat soluble vitamins (such as vitamins A, D, E and K), monoglycerides, diglycerides, triglycerides, Phospholipids and others.

Lipids are made by carbon, hydrogen and oxygen, but the amount of oxygen in lipids is much smaller than in carbohydrates.

Lipids are grouped into fats which are solid at room temperature and oils which are liquid at room temperature.

b. Proteins

Proteins are organic compounds of large molecular mass. For example, the hemoglobin has a molecular mass of 64500. In addition to carbon, hydrogen and oxygen, proteins always contain nitrogen, usually Sulphur and sometimes phosphorus.

c. Water

Living organisms contain between 60% and 90% of water, the remaining being the dry mass. Water is made up of only two elements, Hydrogen and Oxygen.

The function of water is defined by its physical and chemical properties that differ from those of most liquids and make it effective in supporting life.

12.1.2. The sub-units that make up biological molecules

a. Sub-units of Carbohydrates

In carbohydrates the following three categories are identified: Monosaccharides , disaccharides and polysaccharides.

i. Monosaccharides

Monosaccharides are the smallest subunits and are made up of single sugar molecules. Monosaccharides are the sugars like galactose, fructose and glucose with a general formula $C_6H_{12}O_6$, and these typically take on a ring-shaped structure.

All monosaccharides are reducing sugars capable of acting as a reducing agents because they have a free aldehyde group or a free ketone group.

Sources of Monosaccharides:

- Glucose: Fruits and vegetables are natural sources of glucose. It's also commonly found in syrups, candy, honey, sports drinks, and desserts.
- Fructose: The primary natural dietary source of fructose is fruit, which is why fructose is commonly referred to as fruit sugar.
- Galactose: The main dietary source of galactose is lactose, the sugar in milk and milk products, such as cheese, butter, and yogurt.

ii. Oligosaccharides

These are complex carbohydrate chains made up of two to twenty simple sugars joined together with a covalent bond. The most common oligosaccharide is the disaccharide, and examples of this include sucrose, maltose and lactose whose general formula is $C_{12}H_{22}O_{11}$.

A disaccharide is the sugar formed when two monosaccharides are joined by glycosidic bond. Like monosaccharides, disaccharides are soluble in water, have sweet taste, and they are reducing sugars because they are able to reduce Copper II Sulfate of benedict solution directly by heating into copper II oxide except sucrose which is non-reducing sugar which are unable to reduce the copper ions in Benedict's solution.

This makes the color of Benedict's solution to persist when this sugar is boiled with it.

Sucrose is made up of two monosaccharides: Glucose and fructose

Maltose is made up of two monosaccharides: Glucose and glucose

Lactose is made up of two monosaccharides: Glucose and galactose

In maltose ring, the two ring of glucose are bonded by the -1, 4-glycosidic bond while in sucrose the glucose and fructose are bonded by -1, 2-glycosidic bond.

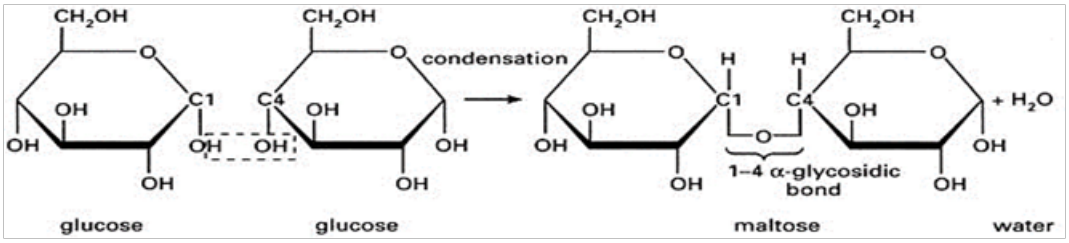


Figure 12.1: Illustration of the formation of maltose

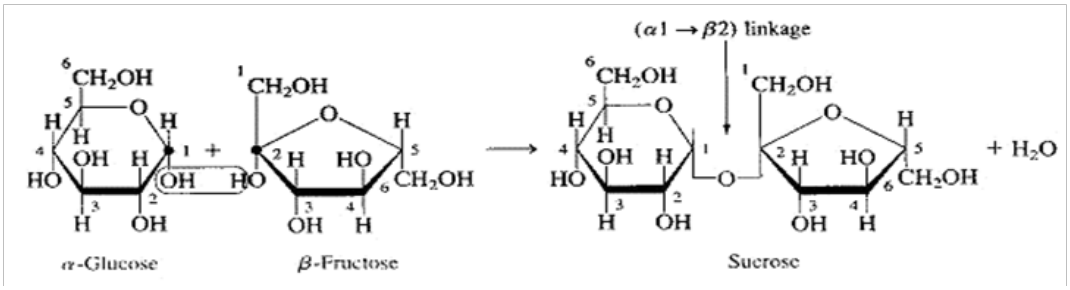


Figure 12.2: Illustration of the formation of sucrose

Table 12.1: Different groups of Disaccharides, their structure, enzyme and source

Disaccharide	Structure	Enzyme for hydrolysis	Sources
Sucrose (Saccharose)	Glucose+glucose	Sucrase	Table sugar, molasses, maple syrup, fondant, cakes, candies, chocolate, ice cream, fruits, honey
Lactose (milk sugar)	Glucose+galactose	Lactase	Milk, ice cream, infant formula, milk chocolate, milk candies, certain pills
Maltose (malt sugar)	Glucose+ glucose	Maltase	Spelt, kamut, sweet potatoes, barley malt syrup, high maltose corn syrup beer, or produced from corn

iii. Polysaccharides

These are known for their ability to store energy and are made up of long chains of glucose sugars. The most common polysaccharides are starch (sugar of plant tissues), glycogen (glucose in the human liver and muscles), cellulose (structural polysaccharide in plants; which acts as a dietary fiber

when consumed), chitin (sugar found in exoskeleton of arthropods) and peptidoglycan (sugar found in the bacteria cell membrane).

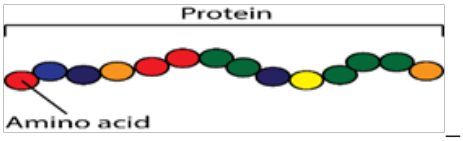
Table 12.2: Different groups of polysaccharides, their composition and source

Polysaccharide	Composition	Source
Digestible		
Starch	Glucose	Cereal grains (wheat, oats, barley, corn, rice…) and their products (bread, pasta, pastries, cookies), potatoes, tapioca, yam, legumes
Glycogen	Glucose	Shellfish, animal liver
Non-digestible (dietary fiber)		
Cellulose	glucose	Whole grains, green leafy vegetables, beans, peas, lentils
Hemicellulose	Arabinose + xylose	Cereals bran
Inulin	Fructose and glucose	Wheat, onions, chicory root, leeks; a food additive
Chitin and chitosan	Glucosamine	Dietary supplements, derived from shells of crustaceans
Pectin	Various monosaccharides	Fruits, carrots, sweet potatoes; a food additive

b. Sub-units of Proteins

These are also referred to as macro-nutrients. The protein are also called body- building food.

The protein molecules are made up of small units called amino acids joined together like links in a chain.



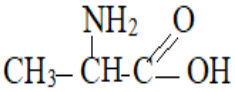


Figure 12.3: Structure of Amino acid

There are 21 different amino acids and each has its own chemical name. Different proteins are made when different numbers and types of amino acids combine through a covalent peptide bond. Proteins are therefore known as polypeptides.

Examples of proteins

a). Collagen, myosin and elastin found in meat,

b). Caseinogen, lactalbumin, lacto globulin found in milk,

c). Avalbumin, mucin and liporitellin found in eggs,
d). Zein found in maize

The 21 different amino acids found in protein are:

Arginine, Serine,Selenocysteine, Leusine, Histidine,Threonine, Glycine, Methionine, Lysine, Asparagine, Proline, Phenylalanine, Aspartic acid, Glutamine, Alanine, Tyrosine, Glutamic acid, Cysteine,Valine, Tryptophan, Isoleucine.

They are used to repair, to build, to maintain our bodies; to make muscles and to make breast milk during lactation period. The proteins are classified into two categories: animal or complete proteins and plant proteins or incomplete proteins.

c. Sub-units of lipids

Lipids are made by two components namely glycerol and fatty acids. The chemical formula for glycerol is $C_3H_8O_3$.

Structural formula of glycerolis

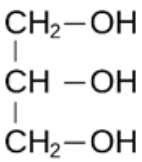


Figure 12.4: Structure of glycerol

Sources and classification of lipids

Fats and oils are obtained from both the plants and animals. And fat is present in food either as visible fat or invisible fat.

Visible fat is the one that is easily seen or detected in food for example; fat in meat, butter, margarine, lard, suet and cooking fat and oil.

Cooking oil

Cheese

Margarine

Figure 12.5: Examples of sources of visible fat

Invisible fat is the part of food that is not easily seen for example fat with in lean meat, egg yolk, flesh of oily fish, groundnuts, soya beans, avocado and fat found in prepared foods, for example, pastry, cakes, biscuits, French fries, pancakes, croquettes.

Cake

Meat pie

Figure 12.6: Examples of sources of invisible fat

Lipids are of different types as it is summarized in the following table.

Table 12.3: Lipids, structure, main role and features

Lipid	Structure	Main role	Other features
Triglyceride	Glycerol plus fatty acids	Compact energy store, insoluble in water so doesn't affect water potential.	Stored as fat, which also has thermal insulation and protective properties.
Phospholipid	Glycerol plus two fatty acids and a phosphate group	Forms a molecule that is part hydrophobic, part hydrophilic ideal for basis of cell surface membranes	Phosphate groups may have carbohydrate parts attached: These Carbohydrates are involved in cell signaling.
Cholesterol	Four carbon-based ring structures joined together		Used to form steroid hormones

Structure of fatty acid

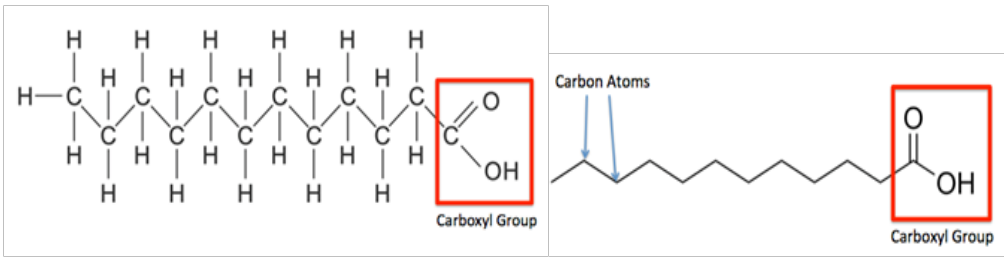


Figure 12.7: Structure of fatty acid

The following are three main types of lipids: Triglycerides, phospholipids and steroids:

- Triglycerides: These are lipids that are obtained from cooking oils, butter and animal fat. They are made up with: one molecule of glycerol and three molecules of fatty acids bonded together by Ester bonds. The triglycerides play the role like storing energy they have thermal insulation and protective properties

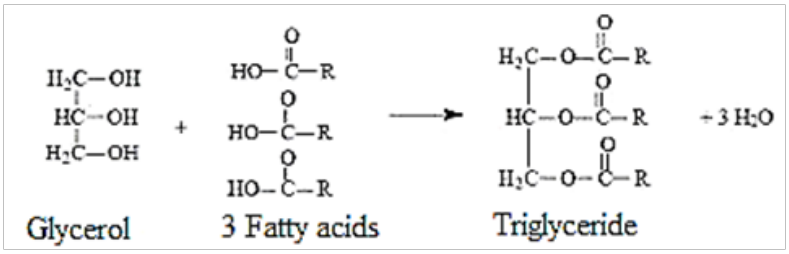


Figure 12.8: Structure of triglyceride formed from glycerol and fatty acid.

- **Sterols**: These are lipids that include steroid hormones like testosterone and oestrogen, cholesterol that is formed four carbon-based rings and it helps in regulation of fluid and strength of the cell membrane.
- **Phospholipids**: They are made up of one molecule of glycerol, two molecules of fatty acids and one phosphate group. The phospholipids form a molecule that is part hydrophobic, part hydrophilic, ideal for basis of cell surface membranes

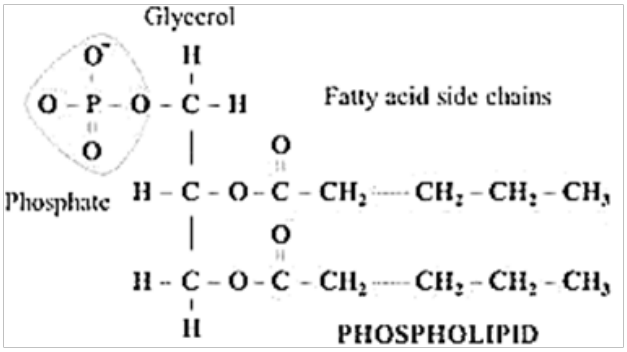


Figure 12.9: structure of phospholipids

Functions of lipids

- Fats are a source of energy. They supply energy to the body more than carbohydrates and proteins.
- Fat surrounds and protects important organs of the body such as the kidney and the heart, however too much fat around the organs is dangerous as it slows down their functioning.

- Fat forms an insulating layer beneath the skin to help keep us warm by preserving body heat and it also protects the skeleton and organs.
- Fat provides a source of fat soluble vitamins A, D, E and K in the body.
- Fat is a reserve of energy for long term storage and can be used if energy intake is restricted.
- Fat in foods provides texture and flavour in foods and it helps to make it palatable.

Food containing fat provides a feeling of satiety or fullness after a meal as fat is digested slowly.

12.1.3. The structure of proteins and their function

Activity 12.1.3

1. From the books make a research on proteins and answer to the following questions:
   a). What are different structures of proteins?
   b). Differentiate globular proteins and fibrous proteins.
2. Take a plastic cord, create the bulk on it and suppose that those are monomers of a long chain of polymer (the whole cord), heat it using a Bunsen burner or another source of fire. Discuss the change that takes place.

Proteins are organic compounds of large molecular mass ranging up to 40,000,000 for some viral proteins but more typically several thousand. For example, the hemoglobin has a molecular mass of 64,500. Proteins are polymers of amino acids and they are not truly soluble in water, but form colloidal suspensions. In addition to carbon, hydrogen and oxygen, proteins always contain nitrogen, usually Sulphur and sometimes phosphorus. Whereas there are relatively few carbohydrates and fats, the number of proteins is limitless. Coined by a Dutch chemist Mulder the word protein etymologically means “of the first importance” due to the fundamental role they play in living cells.

a. Amino acids

Amino acids are group of over a hundred chemicals of which around 20 commonly occur in proteins. They always contain a basic group, the amine group (-NH2) and an acid group (-COOH) together with -R group side chain (Figure 12.14). All the amino acid differs one to another by the structure of their side chain.

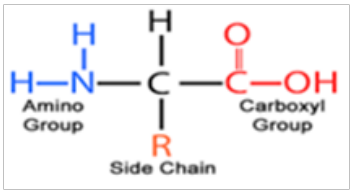


Figure 12.14: The structure of amino acid

Amino acids are divided into two categories including essential amino acid and non-essential amino acid. Essential amino acids are those amino acids which cannot be synthesized by the body. They include isoleucine, leucine, lysine, methionine, phenylalanine, threonine, tryptophan, valine, arginine and histidine. Non –essential amino acids are synthesized by the organism. They include alanine, asparagine, aspartic acid, cysteine, glutamine, glutamic acid, glycine, proline, serine and tyrosine. The simplest amino acid is glycine with H as -R group (Figure 12.15. a). The other one is Alanine with –CH3 as -R group (Figure 12.15.b). All 20 amino acids can be found in diet from animals such as meat, eggs, milk, fish…but diet from plant lack one or two essential amino acids such plant are beans, soy beans…

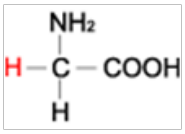


Figure 12.15.a: The structure of the simplest amino acid Glycine.

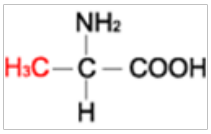


Figure 12.15.b: The structure of Alanine

When an amino acid is exposed to basic solution, it is deprotonated (release of a proton H+) to became negative carboxylate COO -while in acid solution it is protonated (gains of a proton H+) to became ammonium positive ion -NH3 +(Figure 12.16.a and Figure 12.16.b).

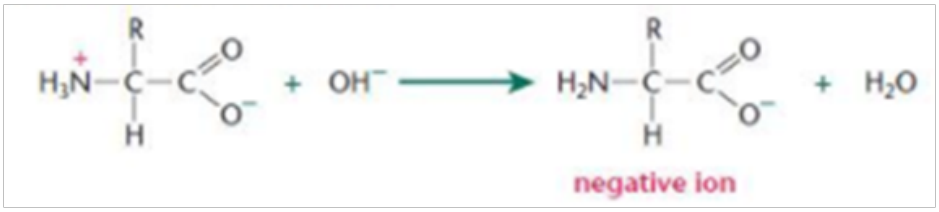


Figure 12.16.a: The amino acid in a basic solution

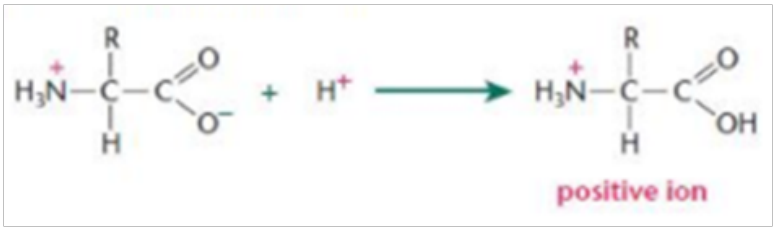


Figure 12.16.b: The amino acid in acid solution

At a physiological pH, usually around 7, the amino acid exists as ZWITTERION (from German means hermaphrodite) it is a molecule with two different charges (positive and negative) at the same time (Figure 12.17).

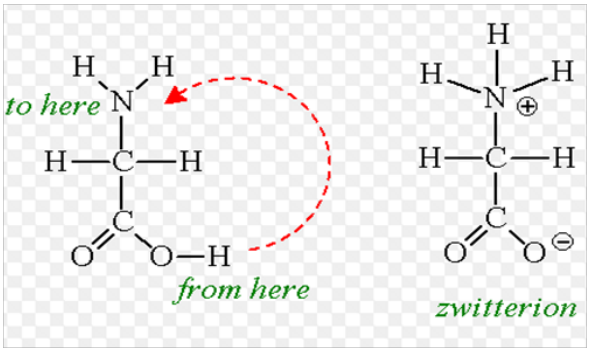


Figure 12.17: Formation of zwitterion in amino acid

b. Formation and breakage of peptide bond

The formation of peptide bond follows the same pattern as the formation of glycosidic bond in carbohydrates and ester bond in fats. A condensation reaction occurs between the amino group of one amino acid and the carboxyl group of another, to form a dipeptide (Figure 12.18).

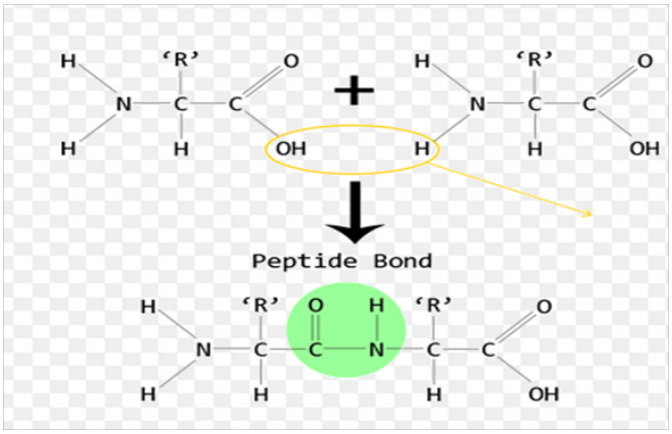


Figure 12.18: Formation of the peptide bond

A peptide bond is formed between two amino acids to form a dipeptide molecule, if three amino acids are assembled together it is a tripeptide, four amino acids form a tetrapeptide and so on. A long chain of amino acid it is called a polypeptide. The polypeptide chain or oligopeptide comprise more than 50 amino acids joined together by a peptide bond.

During digestion, proteins are hydrolyzed and give their monomer amino acids small molecules that can be diffused in the wall of intestine to the organism. In hydrolysis the peptide bond break down by the addition of a water molecule (Figure 12.19).

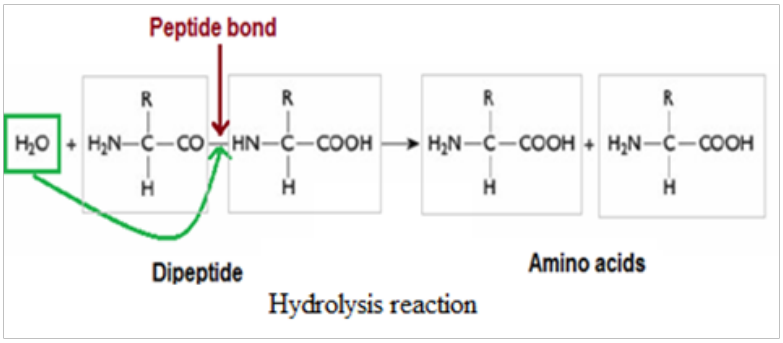


Figure 12.19: Hydrolysis of a peptide bond.

c. Structure and denaturation of proteins

The long chain of polypeptide can take different form according to its molecular weight and the types of bond that hold together atoms and molecules, those form are described as primary, secondary, tertiary and quaternary structure. A human has tens of thousands of different proteins, each with a specific structure and function. Proteins in fact are the most structurally sophisticated molecules known. Consistent with their diverse functions, they vary extensively in structure, each type of protein having a unique three-dimensional shape.

Structure of proteins

i. Primary structure of proteins

Primary structure of a protein is the sequence of amino acid in a linear shape, the amino acid are joined together with the peptide bond.The alteration of this linear sequence (change of the shape) can inhibit the proper function of the protein as well as the function of the protein depending on its three dimensional shape.

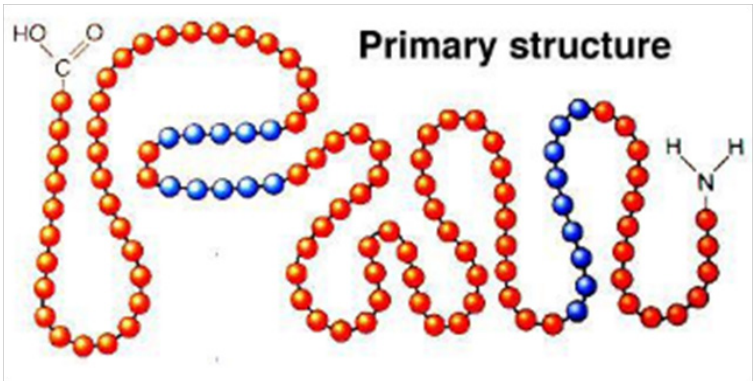


Figure 12.20: Primary structure of protein

ii. Secondary structure of proteins

The regular arrangement of amino acids in primary structure can induce the interaction of the back bone of the polypeptide chain (side chain) by hydrogen bonds. Those side chains are coiled and folded in the patterns that contribute to the protein's overall shape. One such secondary structure is α-helix and sometime ß-pleated sheet.

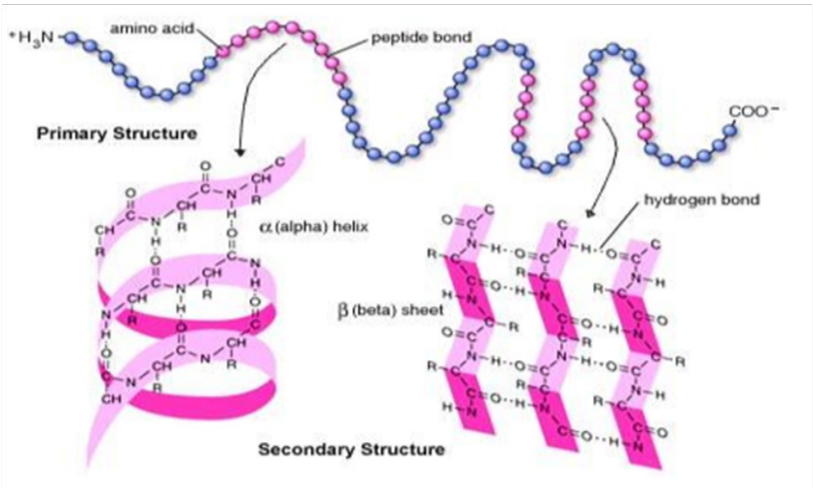


Figure 12.21: Secondary structure of proteins

iii. Tertiary structure of proteins

In addition to hydrogen and peptide bond in primary and secondary structure, the tertiary structure of protein has other types of interaction called hydrophobic interaction. Once the amino acid side chains are close together, Van der Waals interactions hold it together, and their stability depends on ionic bond between positively and negatively charged R groups. The cumulative effect of those week interactions reinforced with the covalent bonds called disulfide bridges (-S-S) give the protein a unique shape.

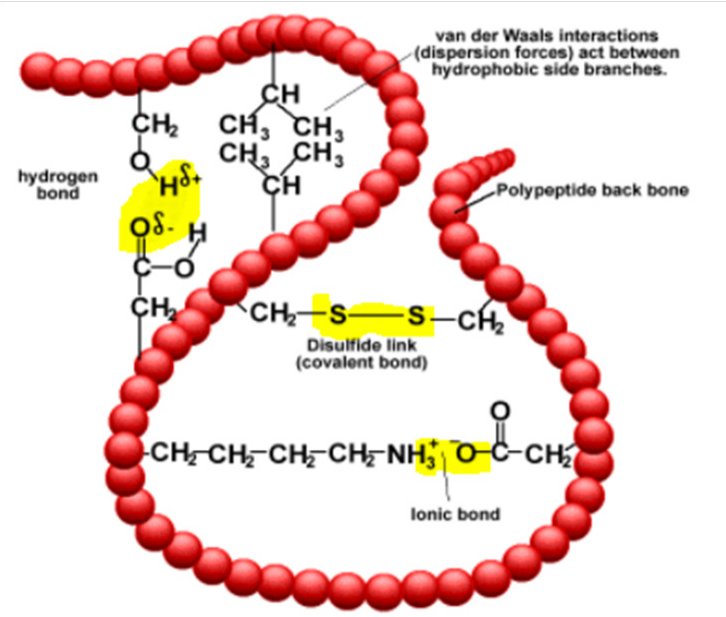


Figure 12.22: Tertiary structure of proteins

iv. Quaternary structure of proteins

Quaternary structure involves different polypeptide chains into one functional three dimensional molecules. For example, the protein that bind oxygen in red blood cells (Hemoglobin) is made by four polypeptide subunits, two of the same kind (α chains) and two of another kind (ß chains). Both α and ß sub-units primarily are α helical secondary structure with anon polypeptide chain iron that binds oxygen called Haeme.

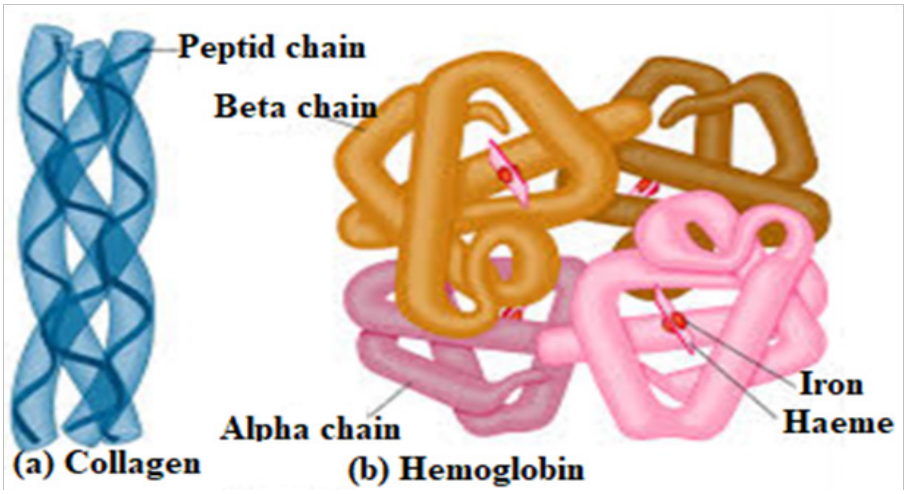


Figure 12.23: (a) Quaternary protein structure of Collagen, (b) Structure of haemoglobin

Globular proteinsuch as haemoglobin and enzymes are soluble in water. In coiling, the hydrophobic R group is dressed toward the inside while the hydrophilic is addressed toward the outside. The opposite is observed in fibrous protein such as collagen and keratin which are insoluble in water. Such protein may contain another group of compound which is not an amino acid called prosthetic group for example the iron (haeme) is prosthetic group in haemoglobin, magnesium in green pigment of plant (chlorophyll) is also a prosthetic group.

Protein denaturation

Protein denaturation is a mechanism by which quaternary, tertiary and secondary structure of protein changes their shape due to external stress called agent or factor of denaturation. Protein denaturation may be temporary or permanent due to a variety of factors. The agent of denaturation may include heat, changes in pH, Ultra Violet (UV°) rays, high salt concentration and heavy metals. Cooked egg is an example of a denatured protein due to the heat. This also explains why excessively high fever disease is fatal to the organism because protein in the blood denature at high temperature. The agents of denaturation will denature protein causing the loose of its shape and hence its ability to function.

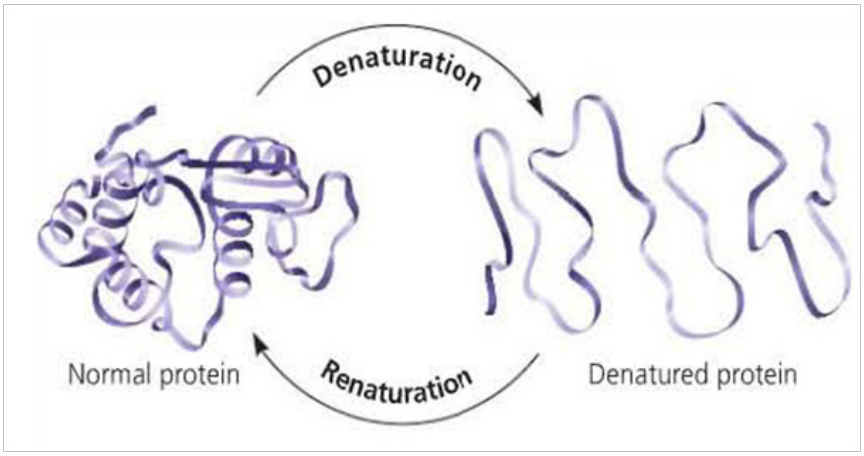


Figure 12.24: Denaturation and renaturation of a protein.

d. Functions of proteins.

- Proteins such as lipase, pepsin and protease act as enzymes as they play a crucial role in biochemical reaction where they act as catalysts.
- Proteins play an important role in coordination and sensitivity (hormones and pigments).

- Proteins have a transport functions. Example: Haemoglobin transport oxygen
- Proteins in the cell membrane facilitate the transport of substance across the cell membrane.
- Proteins provide a mechanical support and strength.
- Proteins such as myosin and actin are involved in movement.
- Proteins play the role of defense of the organisms. Example: Antibodies are proteins

12.1.4. Molecular structure and functions of polysaccharides, glycogen and cellulose

Activity 12.1.4

1. Based on the meaning of monosaccharide, what is the meaning of polysaccharide?
2. Classify the following compound into polysaccharide, monosaccharide and disaccharide
   a). Glucose, fructose and galactose
   b). Lactose, sucrose, and maltose
   c). Starch, cellulose and glycogen
3. Use glucose to form any polysaccharide of your choice

In the same way that two monosaccharides may combine in pairs to give a disaccharide, many monosaccharides may combine by condensation reactions to form a polysaccharide. The number of monosaccharides that combine is variable and the chain produced may be branched or unbranched. Polysaccharide are many but the most known are starch, glycogen and cellulose.

a. Starch

Starch is made up of two components: amylose and amylopectin. Amylose is a linear unbranched polymer of 200 to 1500 α-glucose units in a repeated sequence of α-1,4-glucosidic bonds. The amylose chain coils into helix held by hydrogen bonds formed between hydroxyl groups. A more compact shape is formed. The amylose helices are entangled in the branches of amylopectin to form a complex compact three dimensional starch molecule.

Amylopectin is a branched polymer of 200 to 200,000 α-glucose units per starch molecule. The linear chains of α-glucose units are held together by α-1,4-glucosidic bonds. Branches occur at intervals of approximately 25 to 30

where α-1,6-glucosidic bonds occur. Starch grains are found in chloroplast, potato tubers, cereals and legumes. Starch is insoluble in cold water. It is digested by salivary amylase and pancreatic amylase into maltose and the latter is hydrolyzed by maltase enzyme to form glucose. Therefore, diabetic people should avoid tubers since they are rich in starch which in turn gives glucose.

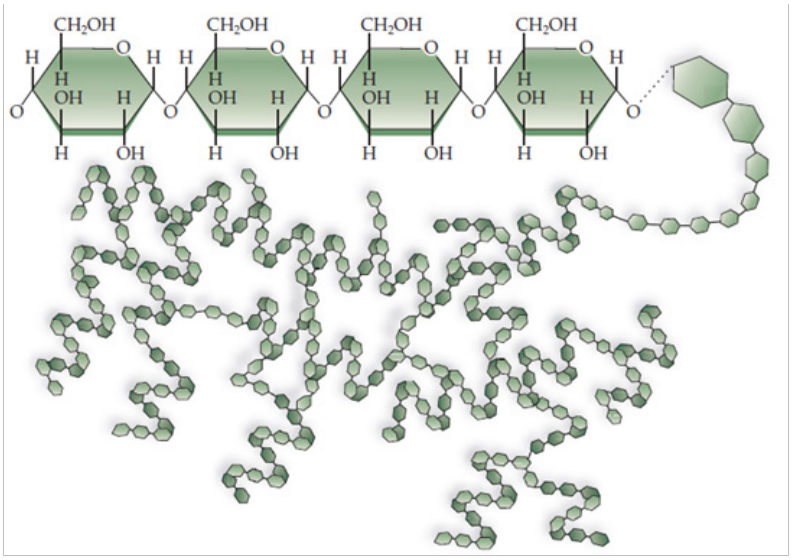


Figure 12.31: structures of starch

b. Glycogen

Glycogen is often called animal starch because it is a major polysaccharide storage material in animals and fungi. The brain and other tissues require constant supply of blood glucose for survival. Some tissues particularly the liver and skeletal muscles store glycogen in the form that can be rapidly mobilized to form glucose. Like starch, glycogen is made up of α-glucose and exists as granules. It is similar to amylopectin in structure but it has shorter chains (10-20glucose unit) and is more highly branched.

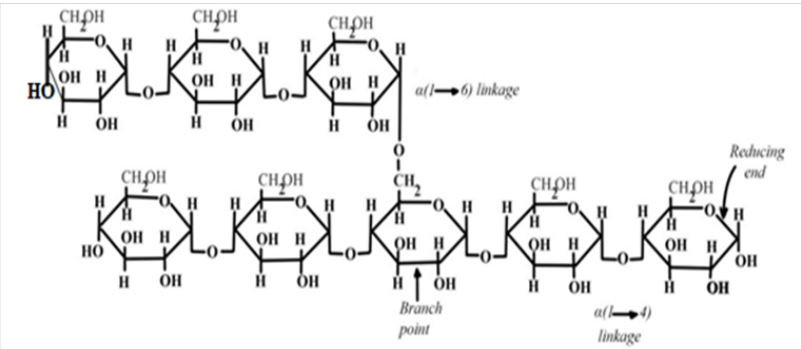


Figure 12.32: Illustration of glycogen

c. Cellulose

Cellulose is the structural polysaccharide in plant cell wall. It is found in vegetables and fruits but cannot be hydrolyzed by enzymes in the human digestive system. Cellulose is composed of long unbranched chains of up to 10,000 ß-glucose units linked by ß-1,4-glucosidic bonds. Each ß-glucose unit is related to the next by a rotation of 180°C with OH groups projecting outwards on either side of the chain.

Cellulose chains run parallel to one another. Unlike amylopectin and glycogen molecules, there are no side chains (no branch) in the cellulose. This allows the linear chains to lie close together. Many H-bonds are formed between the OH groups of adjacent chains. The chains group together to form microfibrils arranged in larger bundles of macrofibrils. The fibrils give the plant cell their high tensile strength and rigidity. The layers of fibrils are permeable to water and solutes.

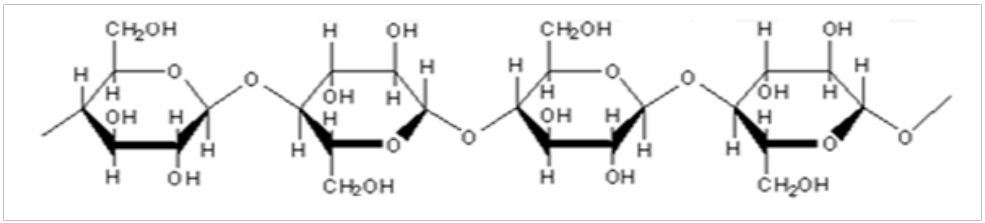


Figure 12.33: Illustration of cellulose

12.1.5. Isomerism of monosaccharide and formation of glycosidic bond

Monosaccharides are group of sweet and soluble crystalline molecules of relatively low molecular mass. They are named with the suffix –ose. The general formula for a monosaccharide is (CH_2O) , with n the number of carbon atoms. The simplest monosaccharide has n=3 and it is a triose sugar. When n = 5, this is a pentose sugar, and when n = 6, this is a hexose sugar. The two common pentose sugars are ribose and deoxyribose, while the most known hexose is glucose. Its molecular formula is C6H12O6. It is the most important simple sugar in human metabolism called simple sugar or monosaccharide because it is one of the smallest units which has the characteristics of this class of carbohydrates.

a. Isomerism and ring formation

Monosaccharides can exist as isomers. The isomer is defined as each of two or more compounds with the same formula but a different arrangement of atoms in the molecule and different properties. The isomer can also be each of two or more atomic nuclei that have the same atomic number and

the same mass number but different energy states. For example, glucose, fructose and galactose share the same molecular formula which is C6H12O6 however they differ by their structural formulae as follow:

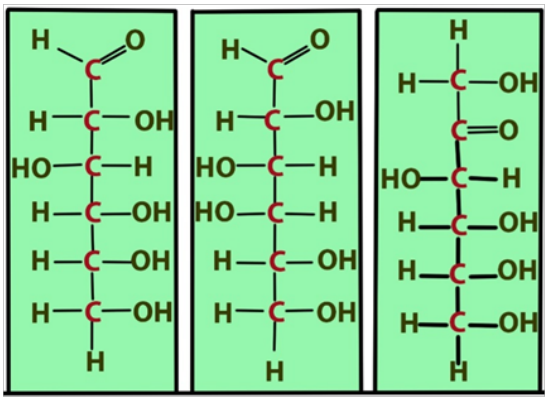


Figure 12.34: Structural formula of glucose, galactose and fructose

One important aspect of the structure of pentoses and hexoses is that the chain of carbon atoms is long enough to close up on itself and form a more stable ring structure. This can be illustrated using glucose as an example. When glucose forms a ring, carbon atom number 1 joins to the oxygen on carbon atom number 5.

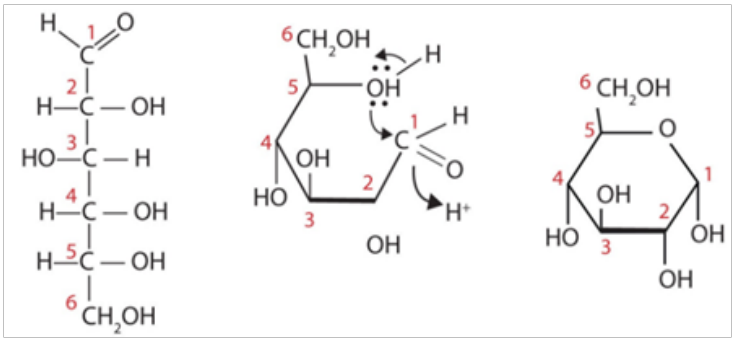


Figure 12.35: Ring formation in the molecule of glucose

All hexoses sugars can exist as straight-chain structures but they tend to form ring structures. Glucose, fructose, galactose can exist in ring structures.

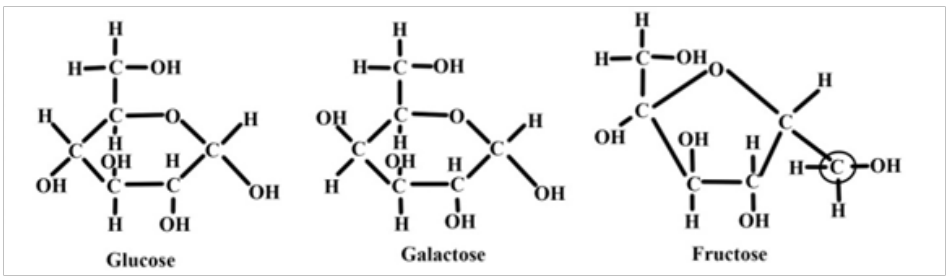


Figure 12.36: Ring structures of glucose, galactose and fructose

Ring monosaccharides are said to be alpha (α) if the -OH group located on carbon 1 is below the ring and beta (β) when the -OH group is above the ring. The molecule of glucose for example can exist as alpha and beta glucose denoted by α-glucose and β-glucose.

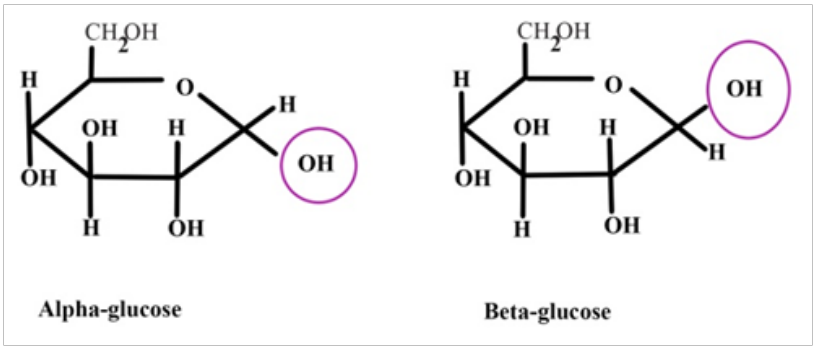


Figure 12.37: Ring form of α-glucose and β-glucose

b. Formation and breakdown of glycosidic bonds

Monosaccharides may combine together in pairs to give a disaccharide (double-sugar). The union involves the loss of a single molecule of water and is therefore a condensation reaction. The bond which is formed is called a glycosidic bond. It is usually formed between carbon atom1of one monosaccharide and carbon atom 4 of the other, hence it is called a -1, 4-glycosidic bond. Any two monosaccharides may be linked together to form a disaccharide of which maltose, sucrose and lactose are the most common.

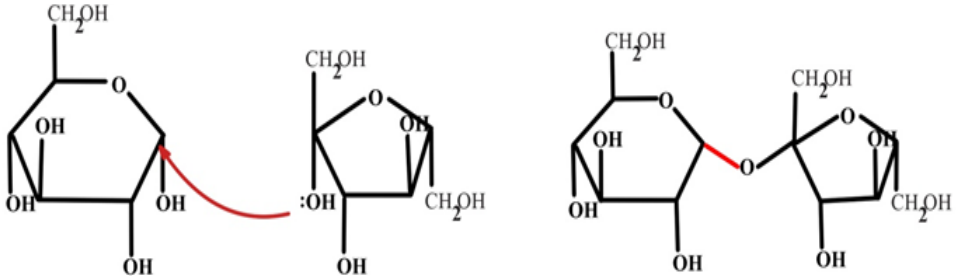


Figure 12.38: Formation of glycosidic bond

The addition of water under suitable conditions is necessary if the disaccharide is to be split into its constituent monosaccharide. This is called hydrolysis water-breakdown or more accurately, breakdown by water.

Application activity 12.1

1. Provided with different kinds of biological molecules such as carbohydrates, proteins, lipids, make a table to show their food source you always take and suggest their functions.
2. Explain what is meant the essential amino acids
3. Describe the formation of a peptide bond?
4. Alanine is an amino acid with -CH3 as a side chain. Writes its structural formulae.
5. Most of the plant lacks one or more of the essential amino acids needed by the body explain how a vegetarian can obtain the essential amino acids.
6. What are the structures of proteins?
7. How do we call the bond in a dipeptide?
8. What type of reaction is involved in the formation of glucose from starch?
9. Use the type of reaction above to form glucose from sucrose molecule
10. Describe how the glycosidic bond is formed.
11. Describe the major types of starch

12.2. Test for the presence of different biological molecules in variety of context

Activity 12.2

You are given solutions containing different food stuffs including maize flour, vegetable cooking oil, and egg white sugar cane liquid and passion fruit. Using prior knowledge of biological molecules to suggest the type of biological molecule in each one of them. Suggest the chemical tests used to identify each of the molecules.

12.2.1. Test for carbohydrates

Activity 12.2.1

Materials required:

Starch powder, Irish potatoes juice, prepared porridge, Iodine solution, beakers, droppers, source of heat and test tubes

A.Test for starch

Procedure

- Mix 1g of starch powder with 100ml of water
- Boil the mixture while stirring; then cool the solution
- Boil the mixture while stirring; then cool the solution
- Put 2ml of starch solution in a test tube labeled 1, 2ml of Irish potato juice in a test tube labeled 2 and 2ml of prepared porridge in a test tube labeled 3
- In each test tube put 2 drops of Iodine solution and shake
- Record your observation and draw a conclusion

B.Test for reducing sugar

Requirements:

Glucose powder, beaker and test tube, Benedict solution, Bunsen burner, droppers.

Procedure:

- In the beaker mix 1cm3 of water and 1g of glucose powder.
- Pour the prepared solution of glucose in a test tube and
- Add 2ml of benedict's solution and heat
- Record your observation.

Biological molecules are grouped into organic molecules including carbohydrates, proteins, lipids, nucleic acids and vitamins. They also contain inorganic molecules such as minerals and water. The first four organic molecules are called macromolecules because they are required in organism in large quantity. Carbohydrates including starch, reducing and non-reducing sugars appear in this category and are the main energy producers in the organisms. Others, including lipids and proteins are needed for building organisms while vitamins protect the organisms against diseases.

We need to ensure that what we take from diet have all required biological molecules.

a. Test for starch

Carbohydrates such as starch are tested by mixing a sample with 2-4 drops of iodine or Lugol's solution. If the sample contains starch the solution will turn from a yellow to brown color to a dark purple/dark blue (Figure 12.10). The color change is due to a chemical reaction between the large carbohydrate molecule and the iodine ions. If the sample does not contain starch the solution remains yellow-brown.

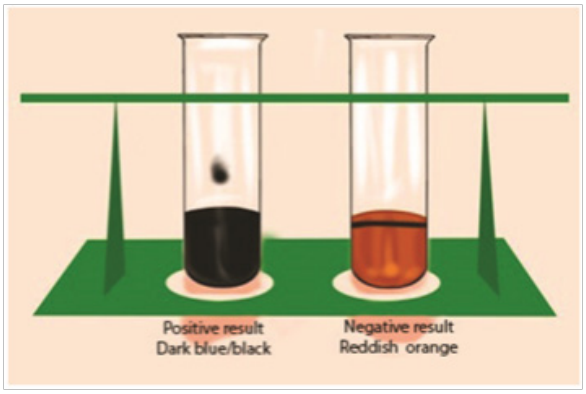


Figure 12.10: Color change during the test for starch

b. Testing for reducing and non-reducing sugar

The presence of reducing sugar can be tested by using benedict reagent. Benedict solution has copper ions that have a light blue color. When this solution is heated in the presence of simple reducing sugars such as glucose, the blue color of copper ions changes from a light green color to rusty orange-brown color (Figure 12.11).

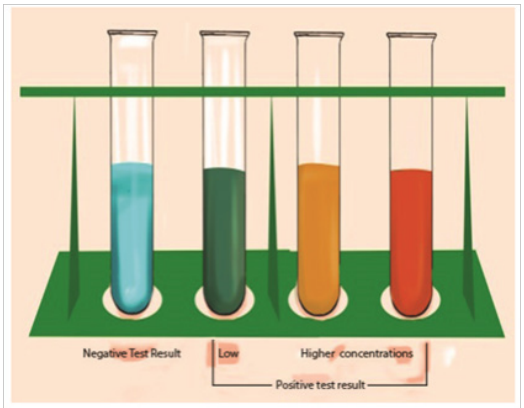


Figure 12.11: Color change during the test for reducing and non-reducing sugar

If the color of Benedict reagent persists, the sugar tested is not a reducing sugar. Note that there is no special reagent to test for non-reducing sugar, but by the addition of HCl, non-reducing sugars can be hydrolyzed to reducing sugars. To test the presence of reducing **sugars**, a **solution of sodium** hydroxide is needed to neutralize the acidity because Benedict reagent works better in neutral solution.

12.2.2. Test for proteins

Activity 12.2.2

Requirements

Milk, eggs, soybeans, test tubes, beakers, mortar for crushing beans, 1% NaOH or 1% KOH solution, 0.1M of $CuSO_4$ solution and Millon's reagent.

Procedure

- Extract the white fluid from an egg (albumen)
- Prepare an extra of soya bean and 10ml of fresh milk
- Put 2ml of albumen solution in a test tube labelled A1 and 2ml in A2
- Put 2ml of milk solution in a test tube labelled B1 and 2ml in B2
- Put 2ml of soya bean solution in a test tube labelled C1 and 2ml in C2
- Put 1ml of KOH or NaOH solution in each of the test tubes A1, B1, and C1. Shake the mixture and add 1ml of $CuSO_4$ solution in each (A1, B1, and C1) test tube
- Put 1ml of Millon's reagent in each of test tubes (A2, B2, and C2). Shake the mixture and thereafter boil the three test tubes (A2, B2, and C2).
- Record and interpret your observations.

The Biuret reagent is used to test for the presence of proteins. It contains copper ions with blue characteristic color. During the copper ions react with protein molecules and causes the biuret solution to turn from a light blue color to purple if proteins are present.

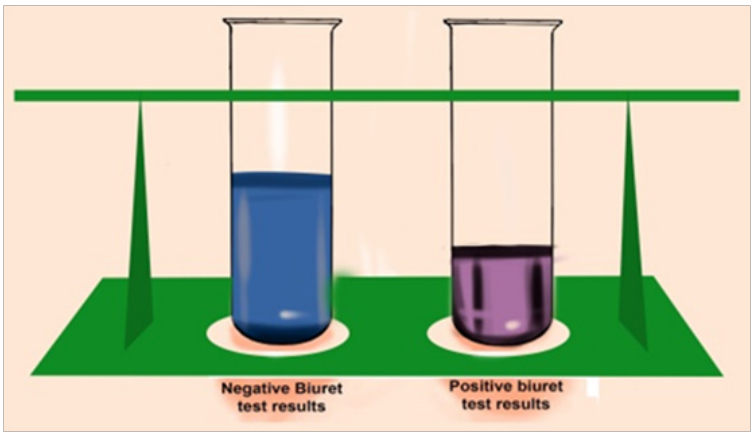


Figure 12.12: Color change during the test for proteins

The test can also be done by using Millon's reagent, which in the presence of proteins, the Millon reagent changes from colorless to pink.

12.2.3. Test for lipids

Activity 12.2.3

Laboratory experiments

Use olive oil to carry out the following experiments

To 2 cm^3 of olive oil in the test tube:

- Add 5 cm^3 of ethanol followed by 5 drops of water.
- Shake the mixture and record your observation.

To another test tube containing 2 cm^3 of olive oil:

- Add 5 drops of Sudan III solution
- Shake thoroughly and examine the mixture in the test tube after few minute and record your observations

The presence of lipids can be determined using Sudan III indicator, which is fat-loving molecules that are colored. During the test for a solution containing lipids, two results are likely to be found: there is either the separation of layers indicating the levels of water and lipid, or the dye migrates toward one of the layers. If the mixtures are all water soluble, the conclusion is that the lipids are not present. In this case, the Sudan III indicator will form small micelles/droplets and disperse throughout the solution.

A positive result indicates the lipid layers sitting on top of the water layer with a red-orange color. When using ethanol for testing lipids the presence of the color changes from colorless to milky (emulsion test).

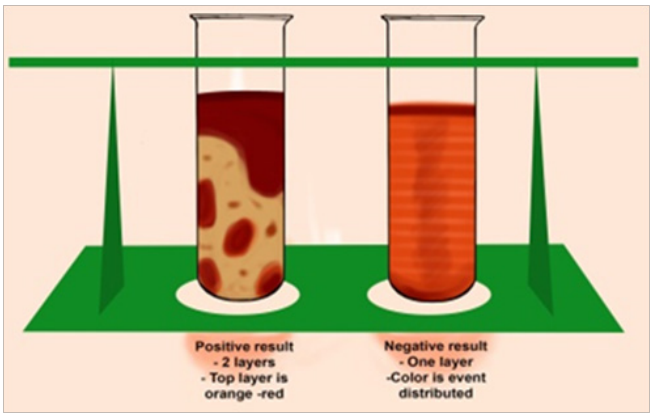


Figure 12.13: Color change during the test for lipids

12.2.4. Test for vitamin C (Ascorbic Acid)

Activity 12.2.4

Squeeze the orange fruits to extract the juice and carry out the following test.

Experiment	Observation	Conclusion
To 3cm3of DCPIP add drops of juice extracted from orange fruit Note the observation and draw a conclusion		
To $2cm^3$ of water add $2cm^3$ of DCPIP Note the observation and draw a conclusion		

Vitamin C is tested by using DCPIP (Dichlophenol Indophenol). Its positive (presence of vitamin C) test decolorizes DCPIP, while the negative (absence of vitamin C) test is indicated by the persistence blue color of DCPIP.

12.3. The structure of DNA

Activity 12.3

1. Discuss the following in the class.

   Nucleotide sequences make DNA, and DNA makes you. If DNA is a double stranded structure, how do the two strands in DNA join or stick together to form double stranded structure?

   Make a report on it and present to the class.

A basic unit nucleotide of DNA (Deoxyribo Nucleic Acid) is made up of pentose sugar(Deoxyribose) , a nitrogenous base, and a phosphate sugar. However, a combination of only a pentose sugar and nitrogenous base, without phosphate group, is called nucleoside.

Pentose sugar + Nitrogenous base + Phosphate group =Nucleotide

Pentose sugar + Nitrogenous base=Nucleoside

In DNA, bases are covalently bonded to the 1' carbon of the pentose sugar. The purine and pyrimidines bases attached to pentose sugar from different positions of their nitrogen bases. Purine bases use the 9th position of nitrogen to attach with 1' carbon of pentose sugar, while pyrimidine bases use the 1st position of nitrogen to attach with 1' carbon of pentose sugar.

In both DNA, the phosphate group (PO42–) attaches to the 5' carbon of pentose sugar. Thus, by attaching phosphate group to a nucleoside yields a nucleoside phosphate or nucleotide. The complex of deoxyribose, nitrogenous base and phosphate group is called DNA nucleotide (a deoxyribonucleotide).

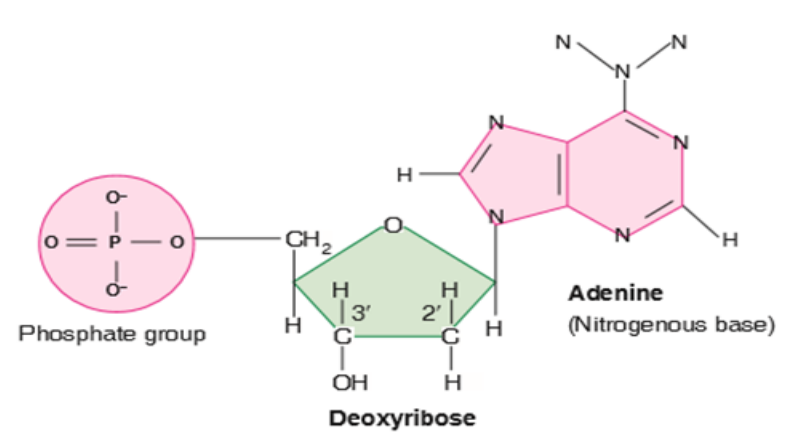


Figure12.25: DNA Nucleotide (Deoxyribonucleotide)

Phosphodiester Bond formation

Two nucleotides are covalently joined together by a bond called phosphodiester bond. In phosphodiester bond, the phosphate group, which is attached on 5' of one nucleotide, forms a bond with the 3' carbon of another nucleotide. In this way, many phosphodiester bonds are formed between sugar and phosphate groups. The repeated sugar-phosphate-sugar-phosphate backbone is a strong one. Because of this strong backbone, DNA is a stable structure.

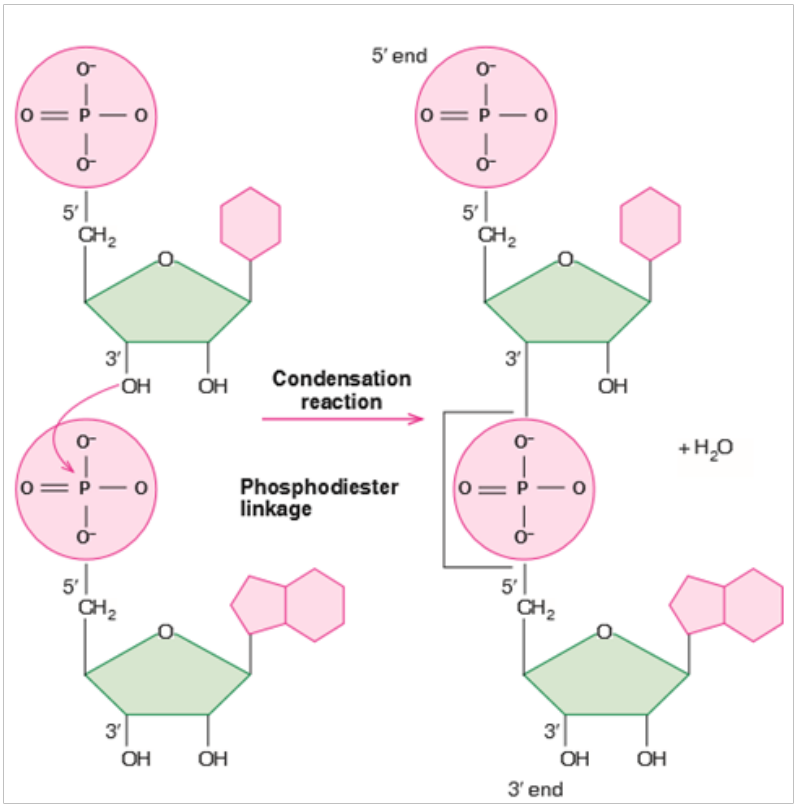


Figure 12.26: A figure showing phosphodiester bond formation (between two ribonucleosides)

In 1953, James D. Watson, an American molecular biologist, and Francis H.C.Crick, a British molecular biologist, proposed a model for the physical and chemical structure of the DNA molecule. Today, their model is known as double helix model of DNA or simply the Structure of DNA. The main features of Watson and Crick double helix model of DNA are:

1. Two polynucleotide chains wind around each other in a right-hand double helix.
2. The two polynucleotide chains run side-by-side in an antiparallel fashion. This means that one strand of DNA will orient itself in a 5' -3' direction, whereas, the other strand will orient itself alongside the first one in a 3'-5' direction. In this way, the two strands are oriented in opposite directions
3. On one hand, the sugar-phosphate backbones lie outside of the double helix. On the other hand, the bases orient themselves toward the central axis of the double helix structure. The bases of one strand are bonded with the bases of the other strand of double helix by hydrogen bonds. These bonds are weak chemical bonds. Since hydrogen bonds are relatively weak bonds, the two strands can be easily separated by heating

the DNA. The bonding of these bases in the double helical structure follows the Chargaff's base pairing rules. For example—Adenine (A) will form a hydrogen bond with Thymine (T). Similarly, Guanine (G) will form a hydrogen bond with Cytosine (C). This specific base paring is called complementary base pairing.

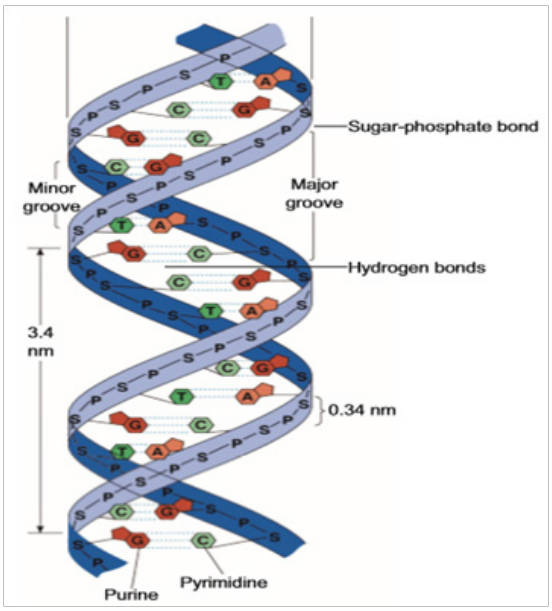


Figure12.27: Double helical structure of DNA(schematic)

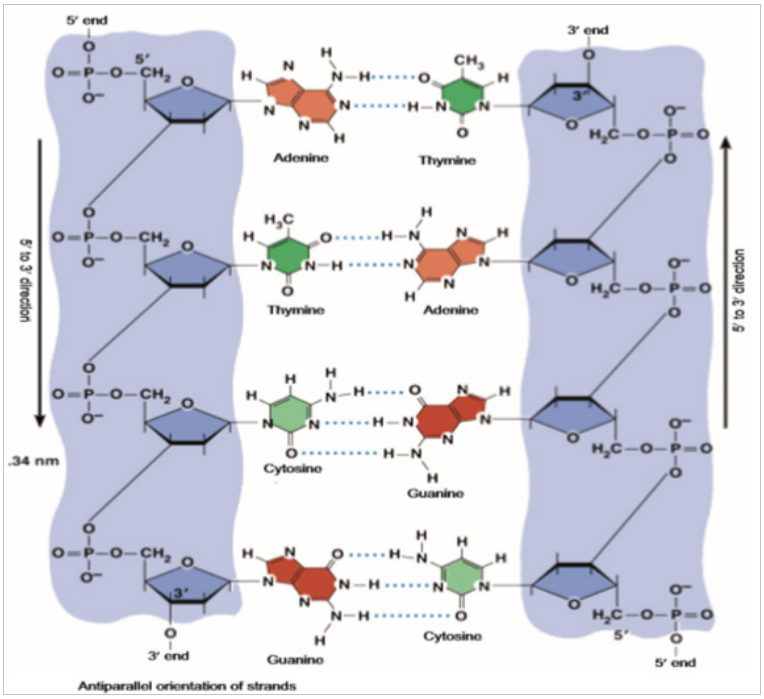


Figure 12.28: Chemical structure of DNA

4. The distance between adjacent bases is 0.34 nm in the DNA helix. A complete turn of the helix takes 3.4 nm. One complete turn, which is 360° turn, accommodates 10 base pairs (bp). And the diameter of the helix is 2 nm.
5. There are major and minor grooves in the double helix. The two sugar-phosphate backbones of the double helix are not equally spaced from one another along the helical axis, because of the way the base bind with each other. As a result, there is an unequal size of grooves between the backbones. The wider groove is called major groove; rich in chemical information. The narrower groove is called minor groove; less rich in chemical information.

DNA is also Described as a Twisted Ladder Structure

A typical ladder has two long wooden or metal side strands or pieces between which a series of rungs or bars are set in suitable distances. In the structure of DNA, the pentose sugars and phosphate groups make up the "long two side strands or pieces" of a typical ladder. And the A-T and G-C base pairs which are bonded by hydrogen bonds make up the

"rungs or bars"of a typical ladder. But unlike a typical ladder which is straight, the two strands of DNA are twisted into spiral. Scientists call this a double helix. DNA also folds and coils itself into more complex shapes. The coiled shape makes it very small. In fact, it is small enough to easily fit inside any of our cells. If a DNA from a cell is unfolded, it would stretch out to a length of about six feet. The structural twisted nature of DNA has been attributed to enhance its stability and strength. Thus, for these simple similarities with a typical ladder, DNA is also referred as a twisted ladder structure.

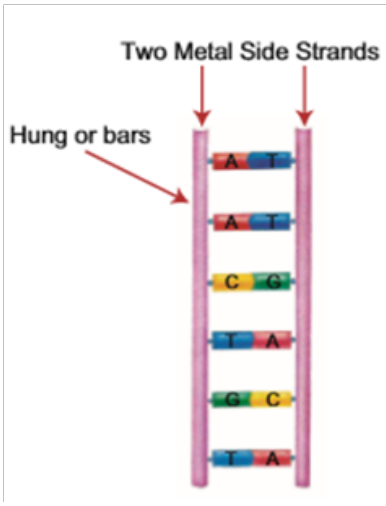


Figure 12.29: Typical ladder

Figure 12.30: A twisted ladder DNAStructure

Application activity 12.3

i. Watson and Crick proposed the model of .....
ii. Enzyme ..... maintains the length of telomere.
iii. ..... can be used to cure cancer.
iv. ..... bonds are seen in both DNA and RNA.
v. ..... directs synthesis of proteins in the body.

12.4. Water and Enzymes

Activity 12.4.a

You are provided with three groups of enzymes: Group A Group B Group C Enzymes Maltase and lactase Dehydrogenase and oxidase Pepsin and renin

	Group A	Group B	Group C
Enzymes	Maltase and lactase	Dehydrogenase and oxidase	Pepsin and renin

Make a research to find out:

a). Specific role of each of the six enzymes mentioned above

b). Criterion followed to name enzymes of group A, B and C respectively

Activity 12.4.b

1. **What is the medium of reaction in the organisms?**
2. **If two people are boiling the same quantity of cooking oil and water, which one could evaporate first? Explain your choice.**

12.4.1. Water

Living organisms contain between 60% and 90% of water, the remaining being the dry mass. The scientist accepts that life originates from water and most of animals live in water. The function of water is defined by its properties mainly: Its physical properties, solvent properties, heat capacity, surface tension and freezing points. The physical and chemical properties of water differ from those of most other liquids but make it uniquely effective in supporting living activities.

a). Physical properties of water

Water has the high boiling point (100°C) compared to other liquid due to the hydrogen bond that exists among molecules of water. This help the water to exist on the surface in a liquid state otherwise it would evaporate.

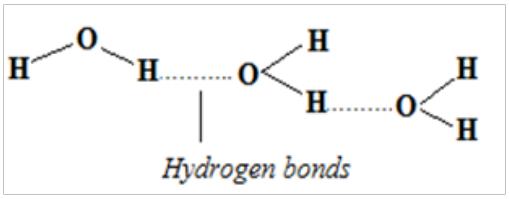


Figure 12.39: Molecular structure of water

Table 12.4: Biological significance of the physical properties of water

Property of water	Significance for living organism
Water is a liquid at room temperature	Provides a fluid environment inside the cells and aquatic environment for organisms to live in.
Universal solvent	The chemical reaction inside the cell happens in aqueous solution. Water is the main transport medium in organisms
Water has high surface tension	Water forms a surface film at an air water interface. This allows some aquatic organisms such as pond skaters to land on the surface of ponds and move over it.
Ice is less dense than liquid water	Ice forms on the surface the body of water and insulates the water below, allowing aquatic life to survive)
Water has adhesion forces	Along with low viscosity adhesion forces help capillarity so that for example, water can move upward through narrow channels in the soil against gravity.
High specific heat capacity	Water being a major component of internal fluids, organisms resist temperature changes and so remain relatively stable.
High latent heat of vaporization	Heat is lost from surface when water evaporates from it. This is used as cooling mechanism; sweating in animals and transpiration in plants
High latent heat of fusion	Cell content and aquatic habitats are slow to freeze in cold weather.
Water is colorless and transparent	Transmission of sunlight helps aquatic plants to photosynthesize.
Water is denser than air	Acts as a habitat for large organisms. It helps supports and disperses reproductive structures such as larvae and large floating fruits like coconuts
Water is difficult to compress	Water is an important structural agent acting as a hydrostastic skeleton in invertebrates(worms) and turgid cells in plants

Water has high tensile strength	Continuous columns of water are pulled up the xylem to the top of the plant during transpiration
Water combines with many organic molecules to from hydrated molecules	Most organic molecules occur in a hydrated form in a cell. If water is removed, their physical and chemical properties are affected; the use and storage of food.

b. Solvent properties of water

Water is a polar molecule due to its chemical arrangement of hydrogen and oxygen atom in asymmetric shape instead of being linear. Most of the substance that are transported in the blood is dissolved in the plasma, the fluid part of the blood. Water occupies around 92% of the constituents of plasma. Thus the oxygen atom has a positive charge and hydrogen atom net positive charges. This is of great importance because all negative and positive ions are attracted by water. Therefore, water is a good solvent because the ionic solids and polar molecules are dissolved in it.

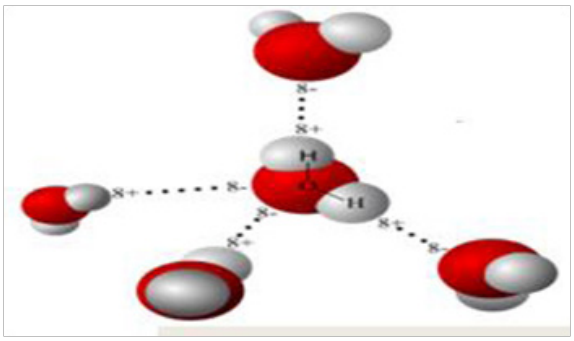


Fig 12.40: Asymmetric model of water

c. Heat capacity and latent heat of vaporization

Large changes of heat results in a comparatively small rise in water temperature this explain why water has a high heat capacity compared to other liquid. The high heat capacity is defined as the amount of heat required to raise the temperature of 1gram to 10C.The high thermal capacity of water make the ideal environment for life in plant and animals because it helps in maintaining the temperature even if there will be environmental fluctuations in temperature. The biological importance of this is that the range of temperatures in which biochemical processes can proceed is narrow.

The latent heat of vaporization is a measure of heat energy needed to cause the evaporation of a liquid, which means to change from water liquid to water vapor. During vaporization the energy transferred to water molecules correspond to the loss of energy in the surroundings which therefore cool down. During sweating and transpiration living organisms use vaporization to cool down.

i. Surface tension

The surface tension of water results from its polar nature, and is defined as the ability of the external surface of the liquid to resist to external force due to cohesive nature of its molecules. The high surface tension of water and the cohesion force play a vital role in capillarity thus help the transport of substance in vessels (tracheid of plant) to the stems and to fulfill the blood in the cardiovascular vessels. Water being the second liquid with high surface tension after mercury its surface tension is lowered by the dissolution of ions and molecules and tend to collect at the interface between its liquid phase and other.

ii. Freezing points

Oppositely to other liquid water expand as it freezes, under 40 C temperatures the hydrogen bond becomes more rigid but more open. This explains why the solid water (ice) is less dense than the liquid water and why the ice floats over water rather than sinking. When the bodies of water freeze the ice float over the liquid act as an insulator and prevent water below it from freezing. This protects the aquatic organisms so that they can survive the winter.

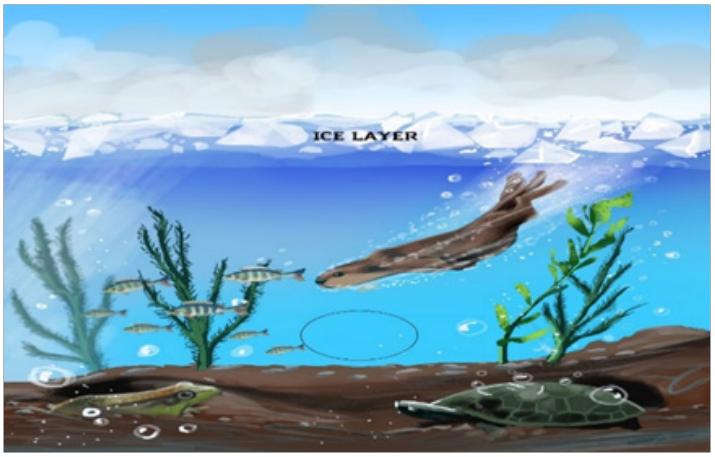


Figure 12.41: Aquatic organisms in winter

d. Functions of water

- Turgidity of plant cell which increase their size is due to the availability of water.
- The transport of substances (minerals, nutrients in plant and animals) is done in water.
- Excretion of waste product
- Support for hydrostatic skeleton.

- Temperature regulation in plant and animals(transpiration)
- Seed germination by breaking down the seed coat
- Medium for biochemical reaction.

12.4.2. Enzymes, their characteristics and actions

Activity 12.4.2a

You are provided with three groups of enzymes:

	Group A	Group B	Group C
Enzymes	Maltase and lactase	Deshydrogenase and oxidase	Pepsine and renine

Make a research from text book or internet to find out:

a). What is the specific role of each of the six enzymes mentioned above?

b). What criterion was followed to name enzymes of group A, B and C respectively?

a. Criteria for naming enzymes

Enzymes are biological catalyst produced by a living organism to control the speed of specific biochemical reactions (metabolism) by reducing its activation energy.

First of all, Individual enzymes are named by adding -ase to the name of the substrate with which they react. The enzyme that controls urea decomposition is called urease; those that control protein hydrolyses are known as proteases.

A second way of naming enzymes refers to the enzyme commission number (EC number) which is a numerical classification scheme for enzymes based on the chemical reactions they catalyse.

As a system of enzyme nomenclature, every EC number is associated with a recommended name for the respective enzyme catalysing a specific reaction. They include:

- Oxidoreductases catalyse redox reactions by the transfer of hydrogen, oxygen or electrons from one molecule to another. Example: Oxidase catalyses the addition of oxygen to hydrogen to form water.
- Glucose + oxygen glucose oxidase ($\rightarrow^{\text{Oxidase}}$)gluconic acid +water
- Hydrolase catalyses the hydrolysis of a substrate by the addition of water.

- Sucrose + water ($\rightarrow^{Hydrolase}$)glucose+ fructose
- Ligases catalyze reactions in which new chemical bonds are formed and use ATP as energy source.
- Amino acid + tRNA ($\rightarrow(^{ligase})$)amino acid-tRNA complex.
- Transferases catalyze group transfer reactions. The transfer occurs from one molecule that will be the donor to another molecule that will be the acceptor. Most of the time, the donor is a cofactor that is charged with the group about to be transferred. Example: Hexokinase used in glycolysis.
- Lyases catalyze reactions where functional groups are added to break double bonds in molecules or the reverse where double bonds are formed by the removal of functional groups. For example: Fructose bisphosphate aldolase used in converting fructose 1,6-bisphospate to G3P and DHAP by cutting C-C bond.
- Isomerases catalyze reactions that transfer functional groups within a molecule so that isomeric forms are produced. These enzymes allow for structural or geometric changes within a compound. Sometime the interconverstion is carried out by an intramolecular oxidoreduction. In this case, one molecule is both the hydrogen acceptor and donor, so there's no oxidized product. The lack of a oxidized product is the reason this enzyme falls under this classification. The subclasses are created under this category by the type of isomerism. For example: phosphoglucose isomerase for converting glucose 6-phosphate to fructose 6-phosphate. Moving chemical group inside same substrate.

A third way of naming enzymes is by their specific names e.g. trypsin and pepsin are proteases. Pepsin, trypsin, and some other enzymes possess, in addition, the peculiar property known as autocatalysis, which permits them to cause their own formation from an inert precursor called zymogen.

b. Characteristics of enzymes

Activity 12.4.2b

Requirement:

Three test tubes, match box, about 1g of liver, 1g of sands, 1% H_2O_2 and MnO2 powder.

Procedure:

- Label three test tubes A, B and C respectively.
- Put about 0.1 g of MnO2 powder in test tube A and 1g of liver in tube B and 0.1g of sand in tube C.

- Pour 5 ml of H2O2 (hydrogen peroxide) in each tube. What do you observe?
- Place a glowing splint in the mouth parts of each test tube. What do you observe?

Questions

1. **Explain your observations.**
2. **Write down the chemical equation of the reaction taking place in tube A and B**
3. **Carry out your further research to find out the characteristics of enzymes**

The following are main characteristics of enzymes

1. Enzymes are protein in nature: all enzymes are made up of proteins.
2. Enzymes are affected by temperature. They work best at specific temperatures; for example, enzymes found in human bodies work best at 37oC. This is called the optimum temperature.
   - Very low temperatures inactivate enzymes. Therefore enzymes are not able to catalyse reactions temperatures beyond the optimum temperature denature enzymes. The structure of the protein molecule is destroyed by heat.
3. Enzymes work best at specific pH. Different enzymes have a given specific pH at which they act best.

   This pH is called optimum pH. Some enzymes work best at low pH (acidic medium) while others work best at high pH (alkaline medium).

   Most enzymes in the human body for instance work best at neutral or slightly alkaline pH. Examples are: lipases, peptidases and amylase. A few enzymes like pepsin that digests proteins in the stomach works best at an acidic pH of 2.
4. Enzymes remain unchanged after catalysing a reaction. Enzymes are catalysts and can therefore be used over and over again in small amounts without being changed.
5. Enzymes catalyse reversible reactions. This means that they can change a substrate into products and the products back to the original substrate.

6. Enzymes are substrate-specific. This means that an enzyme can only catalyse one reaction involving aparticular substrate. This is because they have active sites which can only fit to a particular substrate whose shape complements the active site. For example, pepsin works on proteins but not on fats or starch.
7. Enzymes work rapidly. Enzymes work very fast in converting substrates into products. The fastest known enzyme is catalase, which is found in both animal and plant tissues.
8. Enzymes are efficient. This is best described by the fact that:
    - They are required in very small amounts.
    - They are not used up in a reaction and can therefore be used repeatedly.
9. Enzymes are globular proteins.
10. Enzymes lower the activation energy (Ea) required for reactions to take place. In many chemical reactions, the substrate will not be converted to a product unless it is temporarily given some extra energy. This energy is called activation energy i.e. the minimum energy required the make a reaction take place.

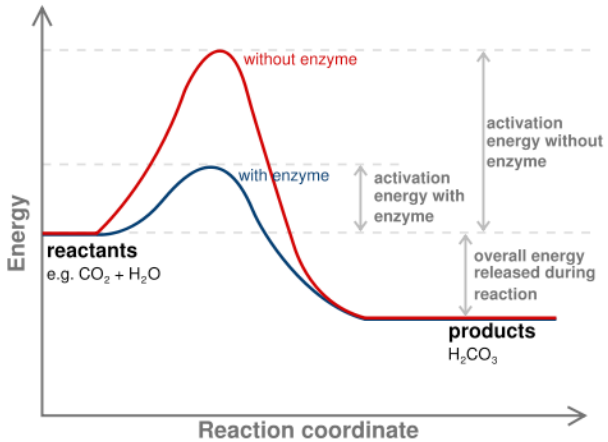


Figure 12.42: An enzyme lowers the activation energy

An enzyme provides a reaction surface for a reaction to take place. This is normally a hollow or cleft in the enzyme which is called the active site, but it is normally hydrophobic in nature rather than hydrophilic. An enzyme provides a reaction surface and a hydrophilic environment for the reaction to take place.

A very small amount of enzymes is needed to react with a large amount of substrate. The turnover number of an enzyme is the number or reactions an enzyme molecule can catalyse in **one second**. Enzymes have a high

turnover number e.g. the turnover number of catalase is 200,000 i.e. one molecule of enzyme catalase can catalyse the breakdown of about 200,000 molecules of hydrogen peroxide per second into water and oxygen at body temperature.

A cofactor is the best general term to describe the non-protein substances required by an enzyme to function properly. This term covers both organic molecules and metal ions. A co-enzyme is an organic molecule that acts as a cofactor. A prosthetic group is a cofactor that is covalently bound to the enzyme.

12.4.3. Factors affecting enzyme action

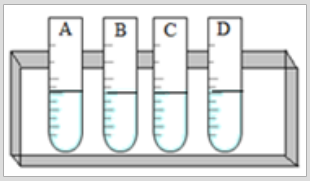
Activity 12.4.3

You will need

Eight test tubes containing 2 cm3 starch solution, amylase solution, and cold water (ice) water bath, iodine solution, HCl solution, and droppers

Procedure:

1. Label your test tubes A-D as follows:



2. Add 1 cm^3 of starch solution to each test tube
3. Keep tube A and B in cold (ice) and tube C and D in the water bath at 35^0C for 5 minutes.
4. Add 1 cm^3 of 1M HCl on test tubes B and D, then shake the mixture to stir.
5. Add 1 cm^3 of amylase solution on each test tube. Shake and therefore keep A and B in cold and C and D in water bath for 10 minutes.
6. Take a sample from each tube and mix it with one drop of iodine. Use a different tile for each test tube. Record and interpret your observation and then draw a conclusion.

Enzymes activities can be limited by a number of factors such as the temperature, the pH, the concentration of the substrate or the enzyme itself and the presence of inhibitors.

i. Temperature

At zero temperature, the enzyme cannot work because it is inactivated. At low temperatures, an enzyme-controlled reaction occurs very slowly.

The molecules in solution move slowly and take a longer time to bind to active sites. Increasing temperature increases the kinetic energy of the reactants. As the reactant molecules move faster, they increase the number of collisions of molecules to form enzyme-substrate complex.

At optimum temperature, the rate of reaction is at maximum. The enzyme is still in active state. The optimum temperature varies with different enzymes. The optimum temperature for enzymes in the human body is about 37°C. When the temperature exceeds the optimum level, the enzyme is denatured.

The effect is irreversible. However, some species are thermophilic that is they better work at high temperatures; others are thermophobic, that is they better work at low temperatures. For example, some thermophilic algae and bacteria can survive in hot springs of 60°C.

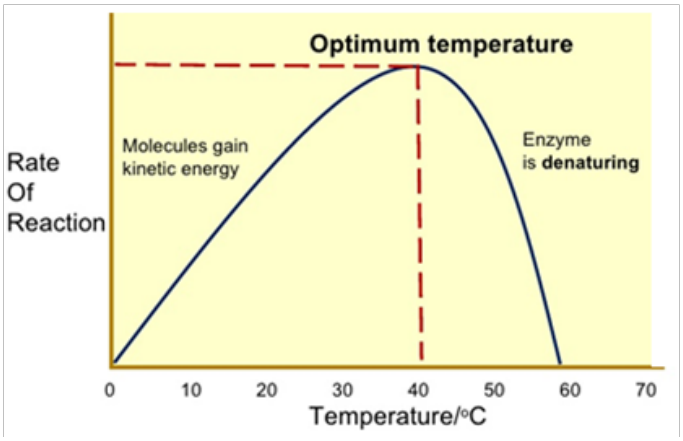


Figure 12.43: Effect of heat on the enzyme controlled reaction

The rate doubles for each 10°C rise in temperature between 0°C and 40°C. The temperature coefficient Q10 is the number which indicates the effect of rising the temperature by 10°C on the enzyme-controlled reaction. The Q10 is defined as the increase in the rate of a reaction or a physiological process for a 10°C rise in temperature. It is calculated as the ratio between rate of reaction occurring at (X + l0)°C and the rate of reaction at X °C. The Q10 at a given temperature x can be calculated from:

$$Q_{10} = \frac{\text{Rate of reaction at } (x+10)^{\circ}C}{\text{Rate of reaction at } x^{\circ}C}$$

Worked out example

The rate of an enzyme-controlled reaction has been recorded at different temperatures as follows:

Temperature/ °C	0.0	10.0	20.0	30.0	40.0	50.0	60.0	70.0
Rate/ mgs^{-1}	0.01	0.1	0.2	0.4	0.8	0.4	0.2	0.02

Calculate the Q10 of that reaction at 30 °C

Solution

$$Q_{10} \text{ at } 30^\circ C = \frac{\text{Rate of reaction at } (x+10)^\circ C}{\text{Rate of reaction at } x^\circ C} = \frac{\text{Rate at } (30+10)^\circ C}{\text{Rate at } 30^\circ C} = \frac{Rate at\ 40\ ℃}{Rate at\ 30\ ℃} = \frac{0.8}{0.4} = x\ 2$$

This means that the rate of the reaction doubles if the temperature is raised from 30°C to 40°C

Be aware that not all enzymes have an optimum temperature of 40°C. Some bacteria and algae living in hot springs (e.g. Amashyuza in Rubavu) are able to tolerate very high temperatures. Enzymes from such organisms are proving useful in various industrial applications.

ii. The pH

Most enzymes are effective only within a narrow pH range. The optimum pH is the pH at which the maximum rate of reaction occurs. Below or above the optimum pH the H+ or OH- ions react with functional groups of amino acids in the enzyme which loses its tertiary structure and become natured.

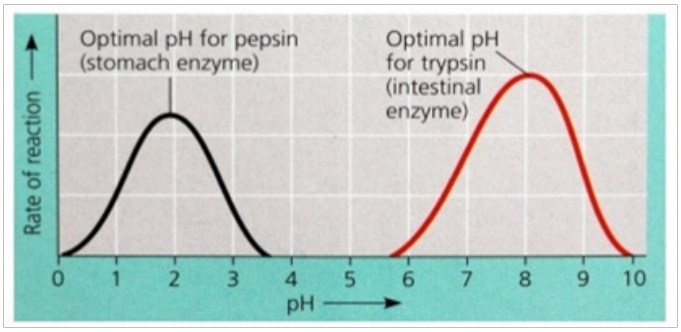


Figure 12.44: Optimum pH depends on nature of enzyme

Different enzymes have different pH optima (look in the table).

Table 12.5: Optimum pH of some digestive enzymes

Enzyme	Optimum pH
Pepsin and rennin	2.0
Salivary amylase	6.8
Trypsin	7.8
Lipase	9.0

iii. Enzyme concentration

The rate of an enzyme-catalyzed reaction is directly proportional to the concentration of the enzyme if substrates are present in excess concentration and no other factors are limiting.

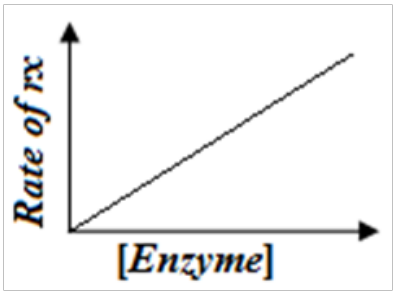


Figure 12.45: Effect of enzyme concentration on the rate of reaction

iv. Substrate concentration

At low substrate concentration, the rate of an enzyme reaction increases with increasing substrate concentration. The active site of an enzyme molecule can only bind with a certain number of substrate molecules at a given time. At high substrate concentration, there is saturation of active sites and the velocity of the reaction reaches the maximum rate.

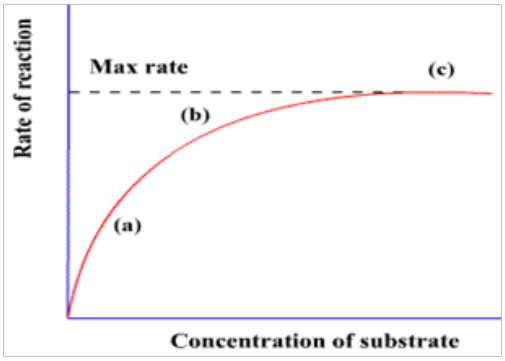


Figure 12.46: Effect of the concentration of substrate on the rate of reaction

v. Inhibitors

The inhibitors are chemicals or substances that prevent the action of an enzyme. An inhibitor binds to an enzyme and then decreases or stops its activity. There are three types of inhibitors:

a). Competitive inhibitors are molecules that have the similar shape as the substrate. They are competing with the substrate to the active site of the enzyme e.g. O_2 compete with CO_2 for the site of RuBP-carboxylase.

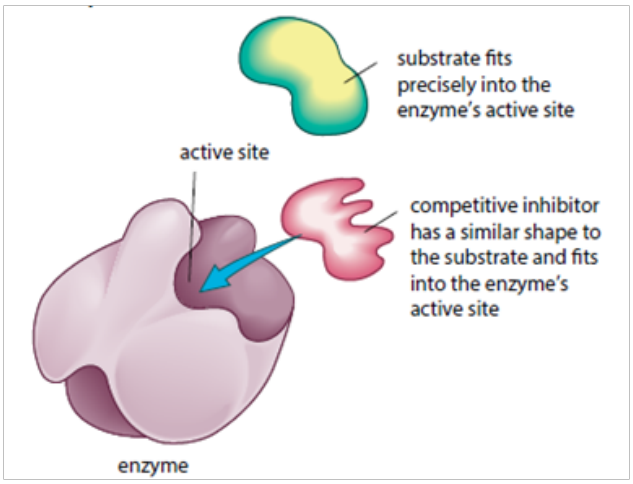


Figure 12.47: Competitive inhibition

b). Non-competitive inhibitors are molecules that can be fixed to the other part of enzyme (not to the active site) so that they change the shape of active site, due to this the substrate cannot bind to the active sit of the enzyme.

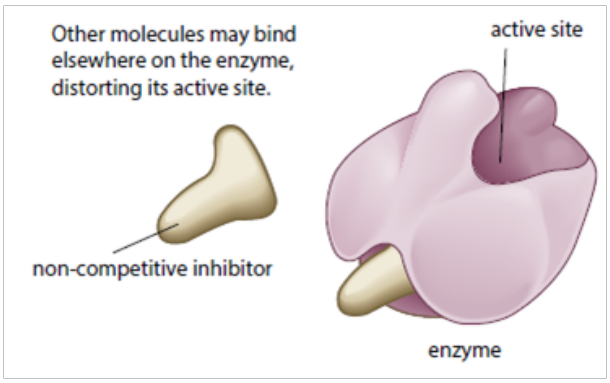


Figure 12.48: Non-competitive inhibition

c). End product inhibitor or Allosteric inhibitor or Allostery.

This is a chain enzymatic metabolic pathway where the final end product acts as an allosteric reversible inhibitor for the first, the second or the third step in the metabolic pathway. The shape of an allosteric enzyme is altered by the binding of the end product to an allosteric site. This decreases enzymatic activity. By acting as allosteric inhibitors of enzymes in an earlier metabolic pathway, the metabolites can help to regulate metabolism according to the needs of organisms. This is an example of negative feedback.

This often happen when few enzymes are working on a large number of substrate e.g. ATP is an end-product inhibitor of the enzyme PFK (Phosphofructokinase) in glycolysis during cell respiration. The end-product inhibitor leads to a negative feedback.

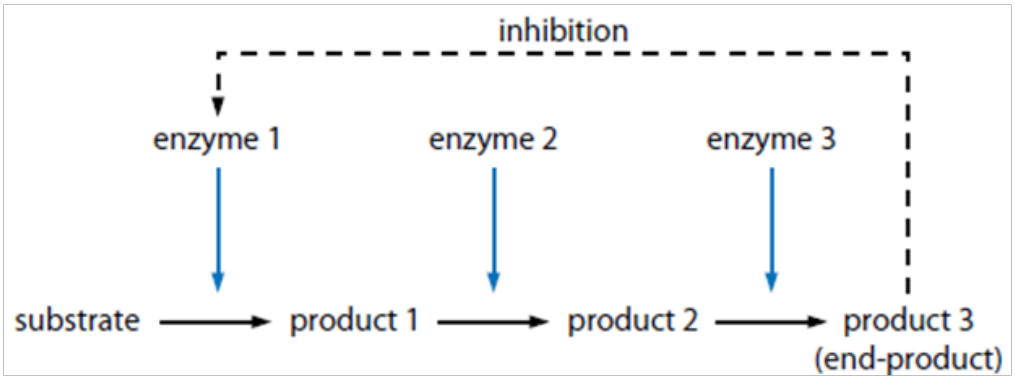


Figure12.49: Model of end – product inhibition

The products of enzyme-catalysed reactions are often involved in the feedback control of those enzymes. Glucose-1-phosphate is the product formed from this enzyme-catalysed reaction. As its concentration increases, it increasingly inhibits the enzyme.

Note: Reversible and irreversible inhibition

Competitive inhibitor is reversible inhibitor as it binds temporarily to the active site. It can be overcome by increasing the relative concentration of the substrate. Some non-competitive inhibitors are reversible, that is, if the inhibitor binds temporarily and loosely to the allosteric site. Some inhibitors have very tightly, often, by forming covalent bonds with enzyme.

The nerve gas DIPF (DiIsopropylPhosphoFluoridate) is an irreversible inhibitor. It binds permanently with enzyme acetylcholisterase, altering its shape. The enzyme cannot bind with and break down its substrate acetylcholine (neurotransmitter). Acetylcholine molecules accumulate in the synaptic cleft.

Nerve impulses cannot be stopped causing continuous muscle contraction. This leads to convulsions, paralysis and eventually death.

Many pesticides such as organophosphate pesticides act as irreversible enzyme inhibitors. Exposure to pesticides can produce harmful effects to the nervous and muscular systems of humans. Heavy metal ions such as Pb2+, Hg2+, Ag+, As+ and iodine-containing compounds which combine permanently with sulphydryl groups in the active site or other parts of the enzyme cause inactivation of enzyme.

This usually disrupts disulphide bridges and cause denaturation of the enzyme.

12.4.4. Importance of enzymes in living organisms

Activity 12.4.4

Elaborate your ideas about the need for different enzymes in living organisms.

Without enzymes, most of the biochemical reactions in living cells at body temperature would occur very slowly at not at all. Enzyme can only catalyze reactions in which the substrate shape fits that of its active site

There are thousands upon thousands of metabolic reactions that happen in the body that require enzymes to speed up their rate of reaction, or will never happen. Enzymes are very specific, so nearly each of these chemical reactions has its own enzyme to increase its rate of reaction. In addition, the organism has several areas that differ from one another by the pH. Therefore, the acid medium requires enzymes that work at low pH while other media are alkaline and therefore require enzymes that work at high pH. In addition to digestion, enzymes are known to catalyze about 4,000 other chemical reactions in your body. For example, enzymes are needed to copy genetic material before your cells divide.

Enzymes are also needed to generate energy molecules called ATP, move fluid and nutrients around the insides of cells and pump waste material out of cells. Most enzymes work best at normal body temperature – about 98 degrees Fahrenheit – and in an alkaline environment. As such, high fevers and over-acidity reduce the effectiveness of most enzymes. Some enzymes need co-factors or co-enzymes to work properly.

12.4.5. Mode of action of enzymes

Activity 12.4.5

There are two main hypotheses that explain the more of action of an enzyme on its substrate: the lock and key hypothesis and the induced-fit hypothesis. Carry out a research to find the relevance of each.

Enzymes do not change but substrates are converted to products. A substrate is a molecule upon which an enzyme acts. In the case of a single substrate, the substrate binds with the enzyme active site to form an enzyme-substrate complex. Thereafter the substrate is transformed into one or more products, which are then released from the active site.

This process is summarized as follows:

$$E+S \rightarrow ES \rightarrow EP \rightarrow E+P$$

Whereby: E = enzyme, S = substrate(s), ES = Complex Enzyme-Substrate and P= product (s). There are two main hypotheses explaining the mechanism of enzyme action:

a. The lock and key hypothesis by Emil Fischer

In this hypothesis the substrate is the key and enzyme is the lock. In otherwise the active site is exactly complementary to the shape of the substrate.

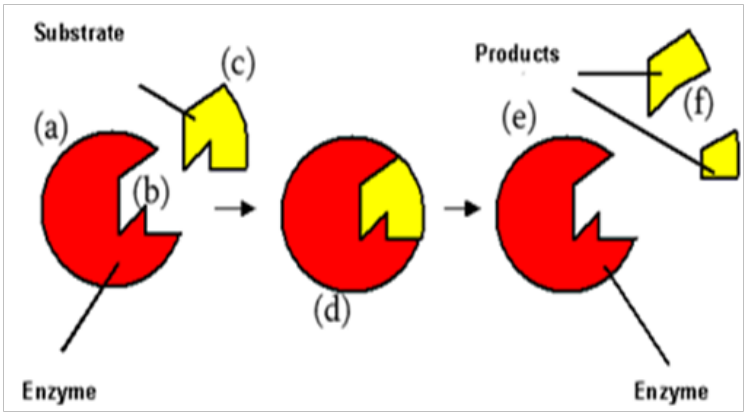


Figure 12.50: Lock and key hypothesis

b. The induced-fit hypothesis by Daniel Koshland

The induced-fit hypothesis is a modified version of the lock and key hypothesis and is more widely accepted hypothesis. In this hypothesis, the active site is flexible and is not exactly complementary to the shape of the substrate.

An enzyme collides with the substrate molecule.The substrate binds to the active site. The bindings induce a slight change in the shape of the enzyme to enclose the substrate making the fit more precise. The active site now becomes fully complementary with the substrate as the substrate binds to the enzyme.

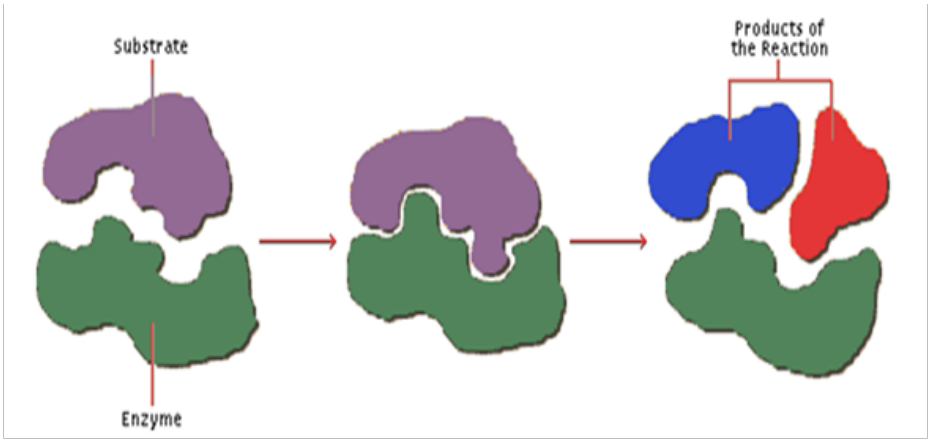


Figure 12.51: Induced fit hypothesis model

Note that the activity of a given enzyme in vitro (outside a living body) may not be necessarily the same in vivo (inside a living body).

Application activity 12.4

1. Fill the blank with appropriate terms: Enzymes are biological _____ produced by _____ cells. Enzymes reduce the amount of _____ energy required for reactions to occur. They consist of globular _____ with _____ structure.
2. Answer the following questions:
   a). What is the main role of enzymes?
   b). What would happen if there are no enzymes in the cell? State any four properties of enzymes.

Skills lab 12

Test the components of the food and beverages produced by different companies in our country. Here, you will need to use the knowledge of food test techniques.test techniques.

End unit assessment 12

1. Biological molecules are divided into (Choose the correct answer):
   a). Organic molecules and inorganic molecules
   b). Carbohydrates and starch
   c). Lipids, carbohydrates and water
   d). Carbohydrates, food and potatoes
2. Name the reagents that are used to test for the following food substances
   a). Lipids
   b). Starch
   c). Reducing sugar
3. Some drops of fresh pineapple fruit juice are added drop by drop to DCPIP solution. The deep blue color of the DCPIP quickly fades.
   a). Explain why the blue colour disappeared?
   b). What is the importance of this food substance to the human body?
4. Write the formula of a monosaccharide with 3 atoms of carbon
5. Compare the structure of fat(triglycerides)and the phospholipids
6. Give two examples of how carbohydrates are used in the body.
7. The formula for a hexose is $C_6H_{12}O_6$ or $(CH_2O)6$. What would be the formula of?
   a). Triose
   b). Pentose
8. Distinguish between:
   a). Alpha glucose and beta glucose
   b). Glycogen and cellulose
   c). Amylopectin and amylose
9. The drug can cleave the covalent bond between two sulfur atoms of non-adjacent amino acids. Which level of protein can be affected the most if the drug is mixed with primary, secondary, tertiary and quaternary structure of proteins.
10. State the property of water that allows each of the following to take place. In each case, explain its importance:

a). The cooling of skin during sweating

b). The transport of glucose and ions in a mammal

c). Much smaller temperature fluctuations in lakes and oceans than in terrestrial (land-based) habitats.

11. Construct the table that organize the following terms and label the columns and rows.

Phosphodiester linkages	Monosaccharide	polypeptides
Peptide bonds	Nucleotides	Triacylglycerol
Glycosidic linkages	Amino acids	Polynucleotides
Ester linkages	fatty acids	Polysaccharides

12. Explain what happen during protein denaturation?

13. Enzymes are biocatalysts.

a). What is the meaning of the following terms elated to enzyme activity?

i. Catalyst

ii. Activation energy

iii. Lock and key

iv. Q10

b). Why are there hundreds of different enzymes in a cell?

c). How do enzymes reduce the activation energy of a reaction?

14. Enzyme activity is related to a number of factors.

a). Explain why enzymes work faster at high temperatures

b). Describe what happens to the enzyme structure if the temperature is raised well above the optimum temperature.

c). How are enzymes affected by pH?

d). Why do different enzymes have a different optimum pH?

e). What is the difference between a reversible and irreversible enzyme inhibitor?

15. Some bacteria and algae can survive in the boiling waters of hot springs. Enzymes from these organisms are used in industrial processes. Why are these enzymes useful?

16. The following set data show the effect of temperature on the completion time of an enzyme reaction.

Temperature / ^{0}C	**0.0**	**15**	**25**	**35**	**45**	**55**	**65**
Rate of reaction / min^{-1}	0.00	0.07	0.12	0.25	0.50	0.28	0.00

a). Plot the data on a graph

b). What is the optimum temperature of this reaction?

c). Describe the shape of the graph between 10 and 40^0 C

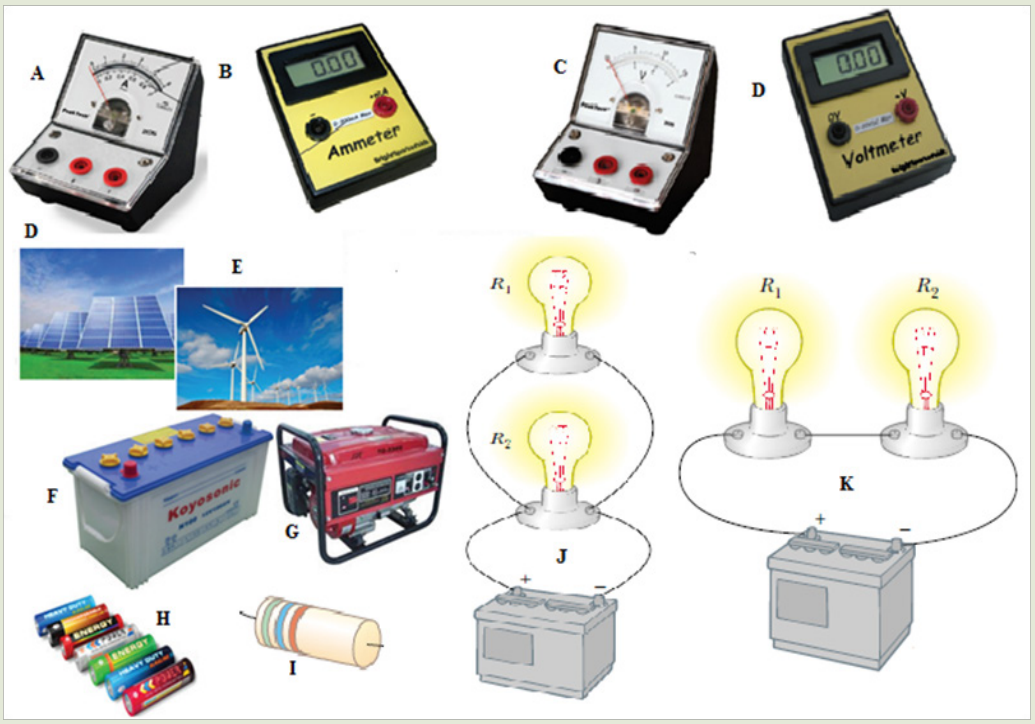
d). Calculate the rate of increase between 20 and 30^0 C.

UNIT 13:

KIRCHHOFF'S LAWS IN ELECTRIC CIRCUITS

Key unit competence: Interpret and solve problems in electric circuits using Kirchhoff's laws

Introductory Activity 13



Look at the illustration given above.

a). What type of devices available in the illustration above?

b). Can you suggest the names of the available devices in the illustration above?

c). Is there any complete circuit in the illustration above?

d). What kind of electrical circuits identified in the illustration above?

e). Have you ever used or connected these electrical components somewhere? If yes, what were the difficulties in handling these electrical components in circuit construction?

f). What can be considered to select the best electrical device(s) to be used in electrical circuit construction?

g). What can be put in recognition to minimize risks when connecting these electrical components in the circuit?

13.1. Simple electric circuits and its construction

Activity 13.1

Observe the following diagram and answer the next questions

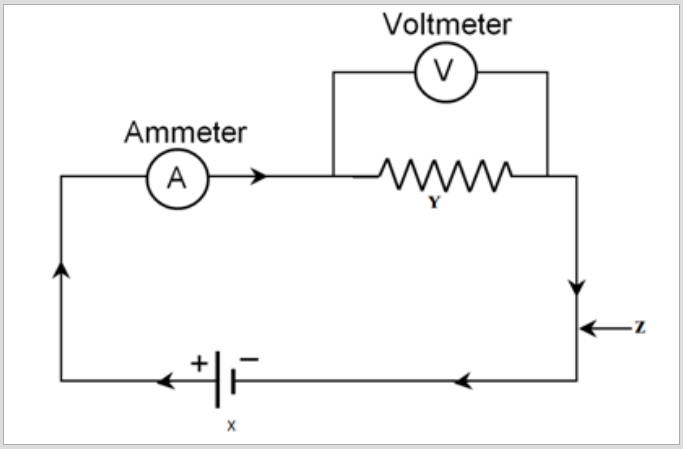


Figure 13.1 Electric circuit

a). Explain the role of part X, Y, Z

b). Give any example of device that is represented by X and Y.

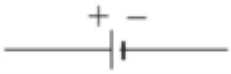
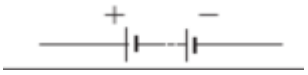
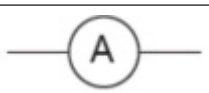
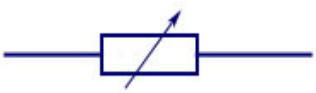
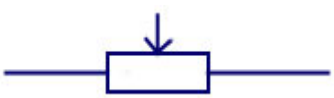
c). Are ammeters and voltmeters necessary in creation of simple electric circuit? Explain.

d). Give other necessary devices in simple electrical circuit.

13.1.1. Electric circuit components, devices and their symbols

In electric circuit diagrams, we represent the actual components with symbols. Table below shows some of the components, their symbols and definition that are used in electric circuit diagrams.

Table 13.1: Electric components and their symbols

Component	Circuit Symbol	Definition/ Description
Cell		Small source of electric energy
Battery		Large source of electric energy
Power supply		D.C mains
Open switch		Breaks the circuit.
Closed switch		Completes the circuit.
Wires joined		Junctions
Connecting wire		Joins two or more components in electric circuit.
Lamb/ bulb		Converts electric energy to heat and light. The best example is the warning light on a car dashboard.
Ammeter		measures electric current
Voltmeter		Measures p.d and voltage
Resistor		Is used to restrict the amount of current flow through a device.
Rheostat		It is used to control the current flow with two contacts. Applicable in controlling lamp brightness, capacitor charge rate, etc.
Potentiometer		It is used to control the voltage flow and has three contacts. Have applications in changing a mechanical angle change to an electrical parameter. Abbreviated as 'POT'.
Capacitor		Is a device that is used to store electrical energy. It consists of two metals plates that are separated by a dielectric. It is applicable as a filter, that is, to block DC signals and allow AC signals. Abbreviated with the letter 'C'.

Lighting Lamp		This is used to provide light for the output.
Heater		This transducer is used to change electrical energy into heat.
Inductor		It is used to produce a magnetic field when a certain current is passed through a coil of wire. The wire is coiled on a soft iron core. Have applications in motors, and tank circuits. Abbreviated as 'L'.

13.1.2. Sample arrangement of electric components in a simple electric circuit

Remember the cell provides electrical energy needed to light the bulb. The bulb converts electrical energy into light and heat energy.

A cell is a kind of a 'pump' which provides electrical energy needed to drive charges along a complete path formed by the wire through the bulb switch and back again to the cell.

When the switch is open, the bulb does not light. This is called an open circuit. When the bulb lights the circuit is called closed circuit.

In a series circuit, the current is the same at all points; it is not used up. In a parallel circuit the total current equals the sum of the currents in the separate branches.

Schematic diagram and its corresponding illustrations:

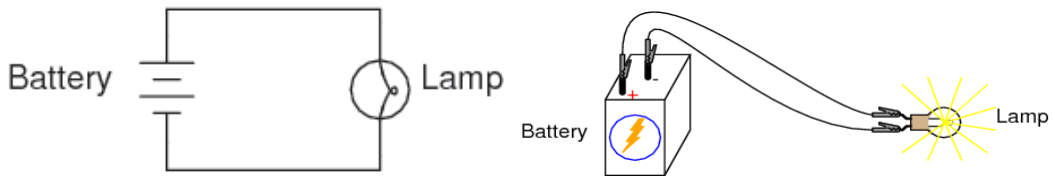


Figure 13.2: Diagram and illustration of simple electric circuit

Illustration of break in electric circuits

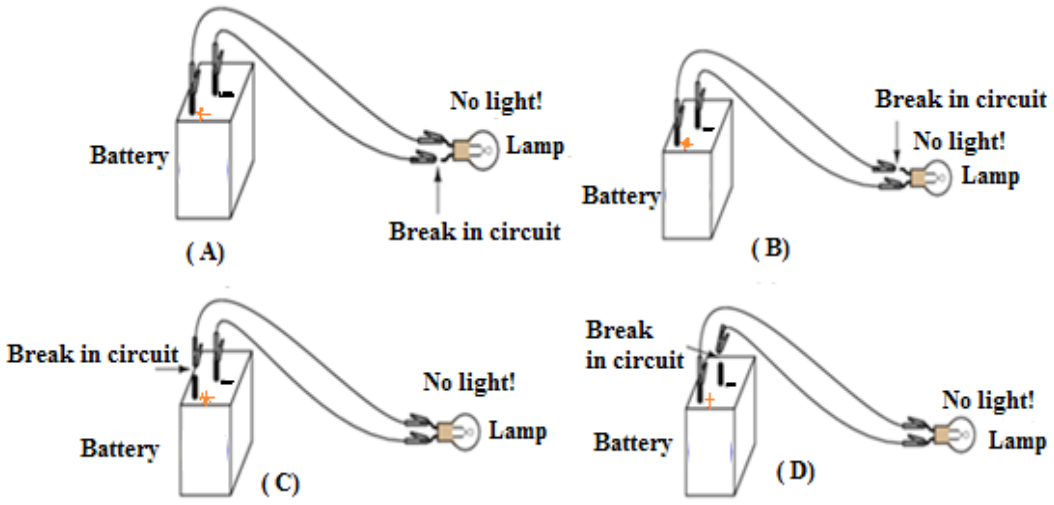


Fig13.3: Illustration of Break in electric circuit

Application activity 13.1

1. Define the term electric circuit.
2. Draw a diagram for a simple circuit using preferable electric components.
3. What is an open circuit?
4. The following are some symbols of electric components.

   a). Name the electric components represented by these symbol.

   b). Using these symbols, draw a simple circuit diagram.

13.2. Measurement of electric current and voltage in a simple electric circuit

Application activity 13.2

Interpret the scale on the figure below for an analogue ammeter that measures current in the range 0 – 1 A, or 0 – 5 A. Determine the reading of the Ammeter.

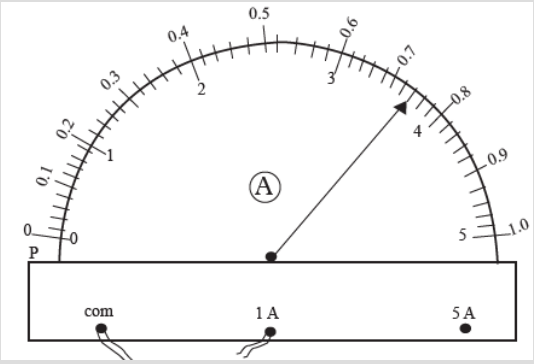


Figure 13.4 : An ammeter with a scale range

Measurement of electric current

Ammeter has two positive terminals and one negative terminal. Fig. (a) Shows an analogue ammeter and (b) shows a digital ammeter.

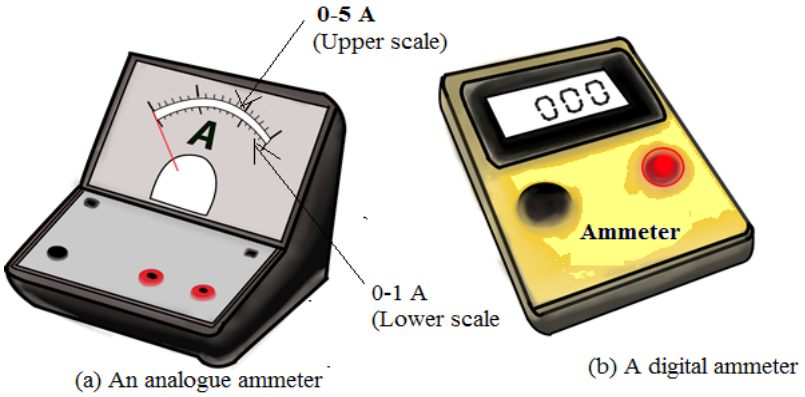


Fig13.5: Ammeters

An analogue ammeter may have more than one scale (Fig.(a)). The magnitude of the current determines the scale to be used. Smaller currents are measured in milliamperes (mA) and microamperes (A).

1 mA = 1/ 1 000 A = 1 x10^{-3} A, 1μ A = 1 10^{-6} A.

How to use an ammeter to measure current in a circuit

Reading an analogue ammeter

Figure below shows the scale on an analogue ammeter that measures current in the range 0 – 1 A, or 0 – 5 A.

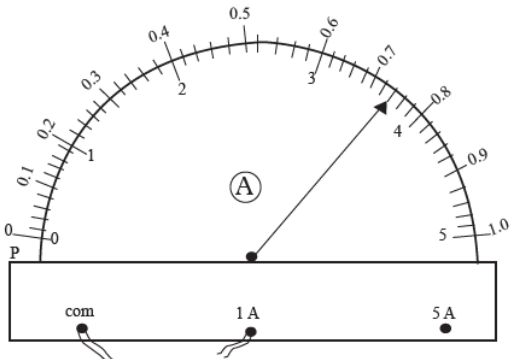


Figure13.6: An ammeter with a scale range of 0 – 1 A, 0 – 5 A

When connected to the 1 A terminal, the upper scale running from 0 - 1 A should be used.

We determine the current represented by each smallest division on the upper scale

as follows:

5 divisions correspond to..... 0.1 A

1 division corresponds to ..... 0.1/5A = 0.02 A

In Figure 13.6, the pointer is on the second mark after the 0.7 mark, hence the ammeter reading is:

0.7 A + (2 divisions × 0.02 A) = 0.7A + 0.04 A = 0.74 A

Potential difference (p.d)

Potential difference is defined as the work done in moving one coulomb of charge from one point to the other in an electrical circuit. The SI unit of potential difference is the volt (V).

In the electric circuit, the electrons move towards the positive terminal of the battery. The battery lifts the electrons up through an electrical height. This electrical height is called a potential.

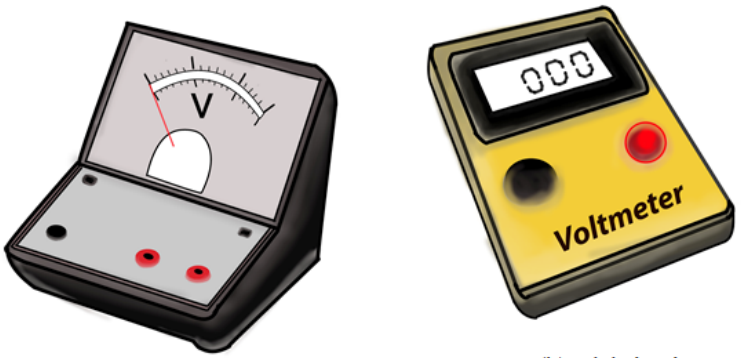
The positive and the negative terminals have a difference in potential. The potential difference is also known as the voltage.

$$Volt = \frac{Work\ done\ (joule)}{charge\ (coulombs)}$$

1 volt is therefore defined as the energy needed to move one coulomb of charge from one point to another.

Measurement of voltage

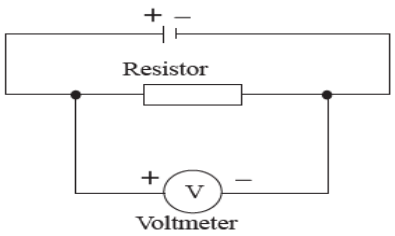
Figure 13.7(a) shows analog voltmeter and (b) shows digital voltmeter the symbol for a voltmeter. A voltmeter is used to measure voltage across a device in an electric circuit.

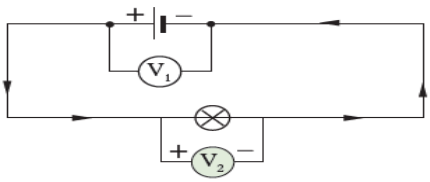


(a) Analog voltmeter (b) Digital voltmeter

Figure13.7: Voltmeters

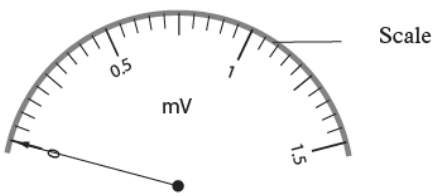
The positive terminal of voltmeter is connected to the wire from the positive terminal of the cells and the negative terminal to the wire leading to negative terminal. A voltmeter is always parallel to the device whose voltage is to be measured.





Connecting a voltmeter in an electric circuit

Voltmeters have uniform scales calibrated in volts or mill-volts. The most used scales have a range of 0 – 5 V and 0 – 1.5 V. Figure below shows a scale of a voltmeter.



Application activity 13.2

1. Define the term potential difference and state its SI units.
2. Name the instrument used to measure voltage.
3. Define a volt.
4. In a circuit, 5 joules are used to drive 2 coulombs of charge across a bulb in a simple circuit. Find the potential difference across the bulb?
5. Name the instrument used to measure potential difference.
6. Two cells, A and B connected in parallel are in series with a bulb as shown in Figure below.

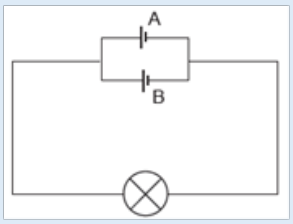


Figure13.8: Circuit with parallel cells

Copy the diagram and show where the:

a). Ammeter should be connected in order to measure the current through cell A.

b). Voltmeter should be connected to measure the potential difference across both the bulb and cell B.

13.3. Different sources of electric current

Activity 13.3

Search on internet about different sources of electric current and answer the following questions.

a). What is a source of electric current?

b). What is another name of electric source of energy?

c). List some of electric sources you have found.

d). On the picture below are some sources of electric energy:

Fig13.9: Circuit with parallel cells

- Name them and tell what the common role they have.
- Tell what energy is changed in electric energy for each device.

There are several different devices that can supply the voltage necessary to generate an electric current. The two most common sources are generators and electrolytic cells.

Generators use mechanical energy, such as water pouring through a dam or the motion of a turbine driven by steam, to produce electricity. The electric outlets on the walls of homes and other buildings, from which electricity to operate lights and appliances is drawn, are connected to giant generators located in electric power stations. Each outlet contains two terminals. The voltage between the terminals drives an electric current through the appliance that is plugged into the outlet.

Generator electrolytic cells use chemical energy to produce electricity. Chemical reactions within an electrolytic cell produce a potential difference between the cell's terminals. An electric battery consists of a cell or group of cells connected together.

There are many sources of electric current other than mechanical generators and electrolytic cells. Fuel cells or engines, for example, produce electricity through chemical reactions. Unlike electrolytic cells, however, fuel cells do not store chemicals and therefore must be constantly refilled.

Certain sources of electric current operate on the principle that some metals hold onto their electrons more strongly than other metals do. Platinum, for example, holds its electrons less strongly than aluminum does. If a strip of

platinum and a strip of aluminum are pressed together under the proper conditions, some electrons will flow from the platinum to the aluminum. As the aluminum gains electrons and becomes negative, the platinum loses electrons and becomes positive.

The strength with which a metal holds its electrons varies with temperature. If two strips of different metals are joined and the joint heated, electrons will pass from one strip to the other. Electricity produced directly by heating is called thermoelectricity.

Some substances emit electrons when they are struck by light. Electricity produced in this way is called photo-electricity. When pressure is applied to certain crystals, a potential difference develops across them. Electricity thus produced is called piezoelectricity. Some microphones work on this principle.

Notice: An electric generator is a device which is used to produce electric energy, which can be stored in batteries or can be directly supplied to the homes, shops, offices, etc. Electric generators work on the principle of electromagnetic induction. A conductor coil (a copper coil tightly wound onto a metal core) is rotated rapidly between the poles of a horseshoe type magnet. A conductor coil (a copper coil tightly wound onto a metal core) is rotated rapidly between the poles of a horseshoe type magnet. The conductor coil along with its core is known as an armature. The armature is connected to a shaft of a mechanical energy source such as a motor and rotated. The mechanical energy required can be provided by engines operating on fuels such as diesel, petrol, natural gas, etc. or via renewable energy sources such as a wind turbine, water turbine, solar-powered turbine. When the coil rotates, it cuts the magnetic field which lies between the two poles of the magnet. The magnetic field will interfere with the electrons in the conductor to induce a flow of electric current inside it.

Application activity 13.3

1. Sources of electric current are also known as electric generators. Explain electric generators.
2. What is an engine? Compare engine to batteries.
3. Compare and contrast battery from dry cells
4. Research on differences between piezoelectricity and thermoelectricity.
5. Research on differences between renewable and non- renewable form of energy sources.

13.4. Connection of electrical current source and resistors

Activity 13.4

Task1: Series circuits

Provided materials: Battery cells, three torch light bulbs and conducting wires

Technical procedure:

- Arrange the battery cells as shown in figure below.
- Connect all the two bulbs in series and switch on.
- Remove one bulb and discuss your observations.
- Arrange the circuit to have two bulbs, and then to have one bulb and discuss the observation.

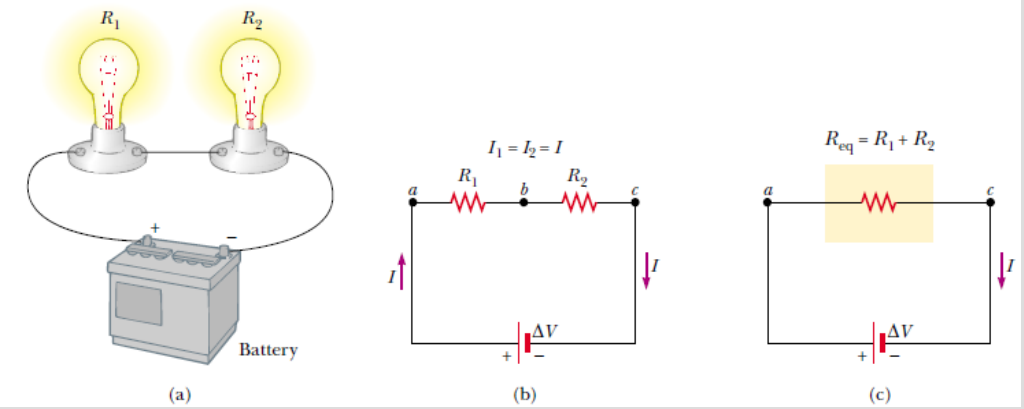


Fig13.10: Series circuit

Use your observations to answer the following questions:

1. What happens in the circuit with three bulbs when one bulb is removed?
2. What happens when the circuit has two bulbs?
3. What happens when the circuit has one bulb only?

Task 2: Parallel circuits

Materials: Battery cells, two torch light bulbs and conducting wires

Technical procedure:

- Arrange the battery cells as shown in figure 13.11.

- Connect all the three bulbs in parallel and switch on.
- Remove one bulb and discuss your observations.
- Remove the second bulb and discuss your observations.

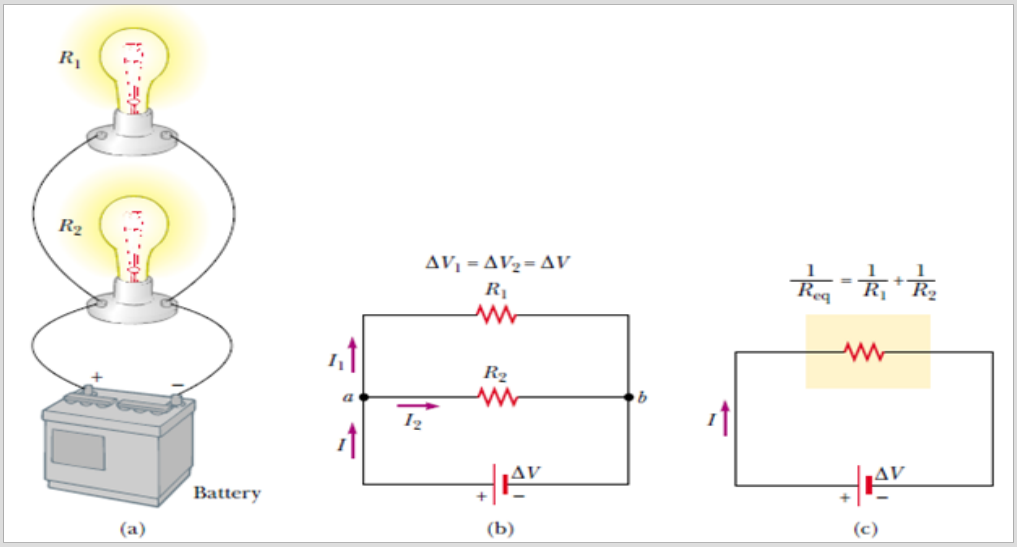


Figure 13.11: Parallel circuit

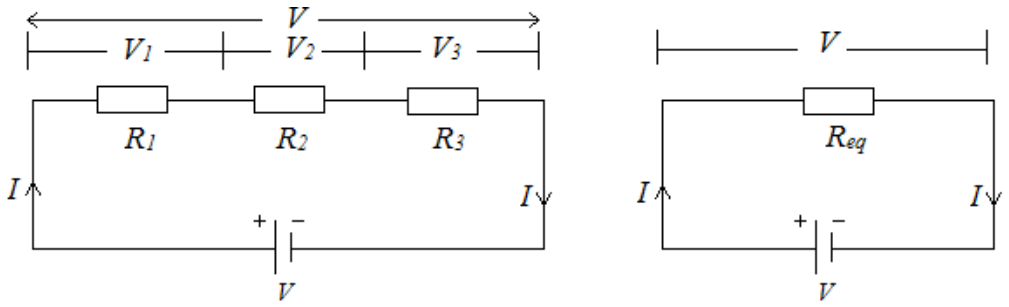
Use your observation to answer to questions below:

1. What happens in the circuit with two bulbs when one bulb is removed?
2. What happens when the circuit has two bulbs?
3. What happens when the circuit has one bulb only?

Circuits consisting of just one battery and one load resistance are very simple to analyze, but they are not often found in practical applications. Usually, we find circuits where more than two components are connected together. There are two basic ways in which to connect more than two circuit components: series and parallel.

13.4.1. Resistors in series

The defining characteristic of a series circuit is that there is only one path for electrons to flow. Consider three resistors R_1, R_2 and R_3 connected in series across a battery of potential difference V. Across the resistors the potential difference drops are V_1, V_2 and V_3 but the current flow I is constant due to the same amount of charges flowing across each resistor.



$$V_1 = IR_1, V_2 = IR_2 \text{ and } V_3 = IR_3$$

The total potential difference across the series combination is

$$V = V_1 + V_2 + V_3 = I(R_1 + R_2 + R_3) \qquad (1)$$

If R is the equivalent or effective resistance of the series combination, and has same charge Q, then $V = R_{eq}$

Substituting V inequation (1)

$R_{eq} = I(R1 + R2 + R3)$

$R_{eq} = R1 + R2 + R3 \qquad (2)$

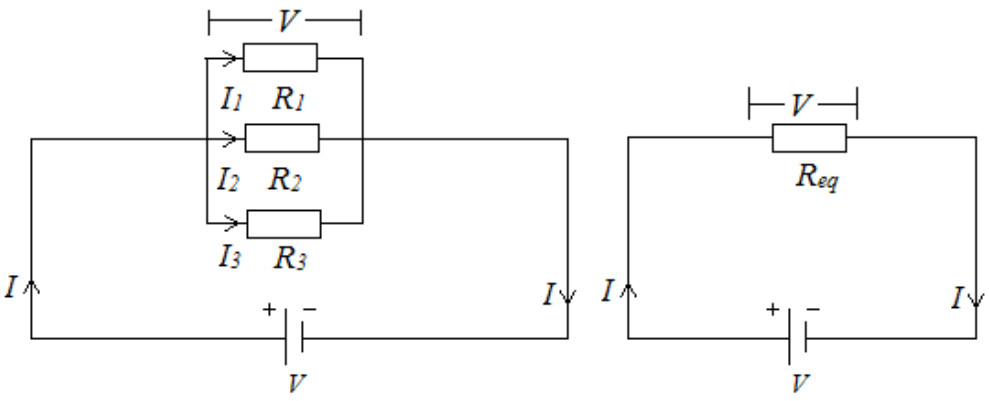
For any n resistors connected in series combination, the effective resistance is

$R_{eq} = R1 + R2 + R3 + ... + R_n$

13.4.2. Resistors in parallel

The defining characteristic of a parallel circuit is that all components are connected between the same set of electrically common points and the resistors form more than one continuous path for electrons to flow.

Assume three resistors of resistance R_1, R_2 and R_3 connected in parallel across a battery of potential difference V. The potential difference across each resistor is the same and is equal to the potential difference V across the battery, but the current flow splits into three parts I_1, I_2 and I_3 due to the separation of charges.



$$I_1 = \frac{V}{R_1}, I_2 = \frac{V}{R_2} \text{ and } I_3 = \frac{V}{R_3}$$

The total current $I = I_1 + I_2 + I_3$

$$I = V\left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_2}\right)$$

If R_{eq} is the resistance of the equivalent resistor,

$$I = \frac{V}{R_{eq}}$$

Hence $\frac{V}{R_{eq}} = V\left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_2}\right)$

$$\frac{1}{R_{eq}} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_2}$$

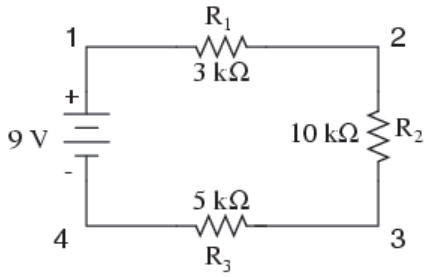
For n resistors connected in parallel combination, the effective resistance is

$$\frac{1}{R_{eq}} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_2} + \ldots + \frac{1}{R_n}$$

Household circuits are always wires so that the lights and appliances are connected in parallel. This way each device operated independently of the others, so if one is turned off the others remain on. This also had the advantage that the voltage supplied to each element is the same.

Example 1

Find the equivalent resistance in the circuit below:



Solution:

$$R = R_1 + R_2 + R_3$$
$$R = 3k\Omega + 10k\Omega + 5k\Omega$$
$$R = 18k\Omega$$

With a 9V battery, by V = I R the total current in the circuit is:

$$I = \frac{V}{R} = \frac{9V}{18k\Omega} = 0.5mA$$

The current through each resistor would be 0.5 A.

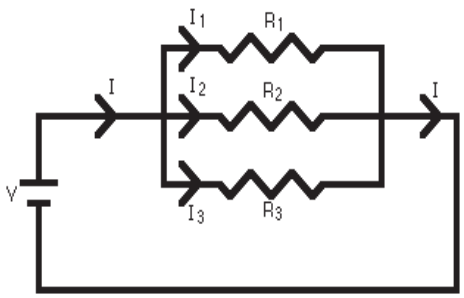
Example 2

A parallel circuit is shown in the figure below. In this case the current supplied by the battery splits up, and the amount going through each resistor depends on the resistance.

From the figure below, if the values of the three resistors are

R_1=8 Ω ,R_2=8 Ω and R_3= 4Ω

Determine the total resistance of the circuit.



Solution

The total resistance R is found by $\frac{1}{R}=\frac{1}{8}+\frac{1}{8}+\frac{1}{4}=\frac{1}{2}$

This gives that $R=2\Omega$

With a 10 V battery, by V = I R the total current in the circuit is: $I=\frac{V}{R}=\frac{10V}{2\Omega}=5\,A$

The individual currents can also be found using $I=\frac{V}{R}$. The voltage across each resistor is 10 V, so:

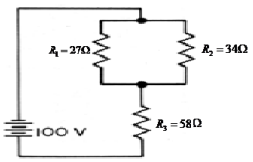
$$I_1=\frac{10}{8}=1.25\,A,\ \mathrm{I}_2=\frac{10}{8}=1.25A$$

$$I_3=\frac{10}{4}=2.5A$$

Note that the currents add together to 5A, the total current.

Example 3

Interpret the circuit below and determine the total resistance of the circuit.



Solution

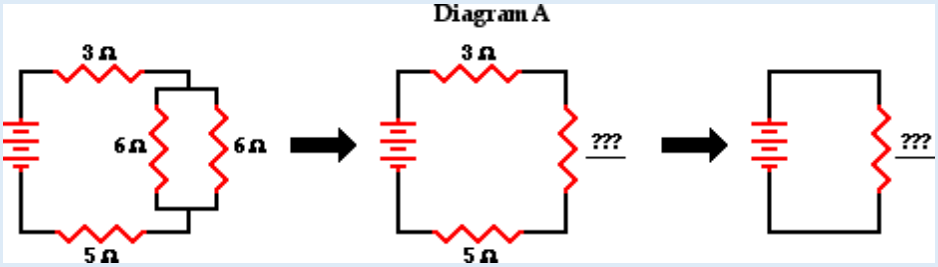
Here we can use the shorter product over sum equation as we only have two parallel resistors.

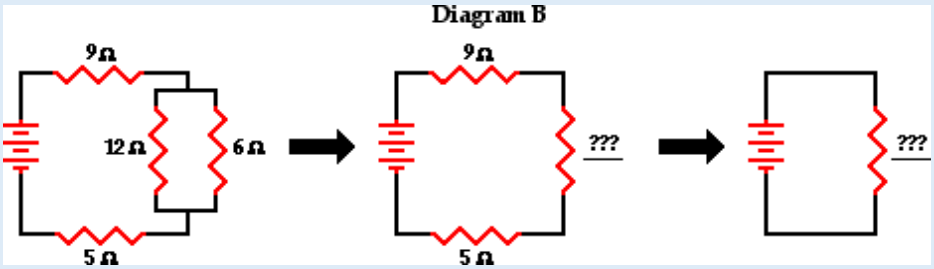
$$R_{1\|2}=\frac{(R_1\times R_2)}{(R_1+R_2)}=\frac{27\times34}{27+34}=\frac{918}{61}=15.049\,\Omega$$

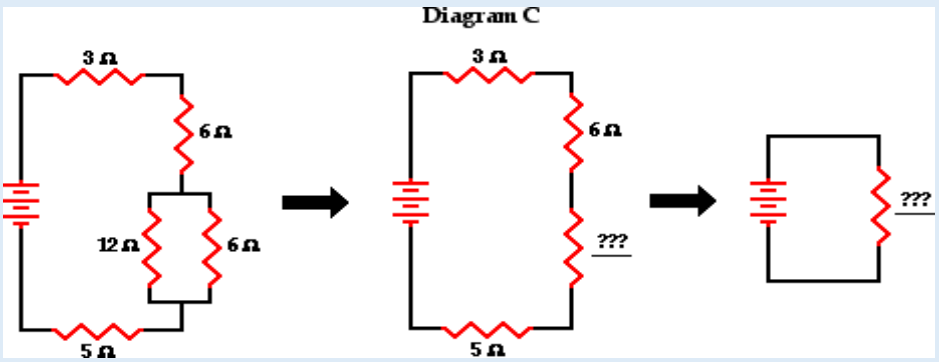
$$R_T=R_{1\|2}+R_3=15.049\,\Omega+58\Omega=73.049\,\Omega$$

Application activity 13.4

1. Use the concept of equivalent resistance to determine the unknown resistance of the identified resistor that would make the circuit's equivalent.

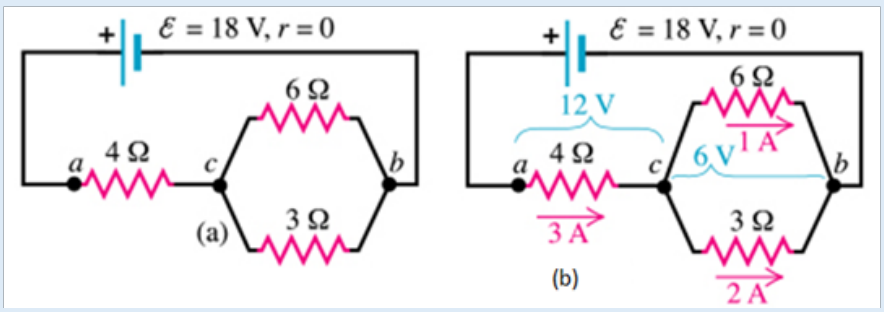




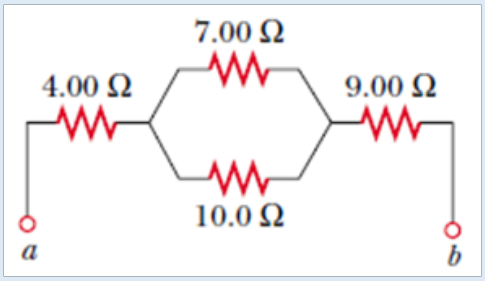


2. A parallel pair of resistance of value of 3Ω and 6Ω are together connected in series with another resistor of value $4\ \Omega$ and a battery of e.m.f. 18 V as shown on the fig. (a) below. Calculate the current through each resistor.

Calculate the current through each resistor.



3. (a) Find the equivalent resistance between points a and b in Figure below. (b) A potential difference of 34.0 V is applied between points a and b. Calculate the current in each resistor.



4. Use your understanding of equivalent resistance to complete the following statements:
   i. Two 3Ω resistors placed in series would provide a resistance which is equivalent to one ___ Ω resistor.
   ii. Three 3Ω resistors placed in series would provide a resistance which is equivalent to one ___ Ω resistor.
   iii. Three 5Ω resistors placed in series would provide a resistance which is equivalent to one ___ Ω resistor.
   iv. Three resistors with resistance values of 2Ω , 4Ω and 6Ω are placed in series. These would provide a resistance which is equivalent to one ___ Ω resistor.
   v. Three resistors with resistance values of 5Ω , 6Ω and 7Ω are placed in series. These would provide a resistance which is equivalent to one _____ Ω resistor.
   vi. Three resistors with resistance values of 12Ω, 3Ω and 21Ω are placed in series. These would provide a resistance which is equivalent to one _____ Ω resistor.
5. As the number of resistors in a series circuit increases, the overall resistance _____ (increases, decreases, remains the same) and the current in the circuit _____ (increases, decreases, remains the same).
6. Imagine that we add a third resistor in series with the first two. Does the current in the battery (a) increase, (b) decrease, or (c) remain the same? Does the terminal voltage of the battery (d) increase, (e) decrease, or (f) remain the same?
7. Imagine that we add a third resistor in parallel with the first two. Does the current in the battery (a) increase, (b) decrease, or (c) remain the same? Does the terminal voltage of the battery (d) increase, (e) decrease, or (f) remain the same?

13.5. Kirchhoff's laws and its applications in solving problems in electric circuits

Activity 13.5

A single-loop circuit contains two resistors and two batteries, as shown in the following figure (neglect the internal resistances of the batteries).

a). Find the current in the circuit.

b). What power is delivered to each resistor? What power is delivered by the 12V battery?

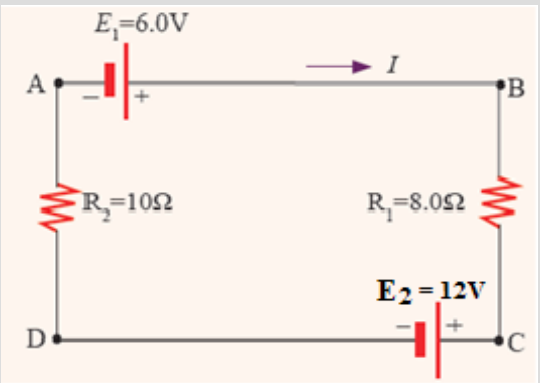


Figure13.12: Electric circuit with more emf

13.5.1. Kirchhoff's laws

Simple circuits can be analyzed using the expression V = IR and the rules for series and parallel combinations of resistors. Very often, however, it is not possible to reduce a circuit to a single loop.

The procedure for analyzing more complex circuits is greatly simplified if we use two principles called Kirchhoff's rules developed by the German Physicist Gustav Robert Kirchhoff (1824-1887).

First, here are two terms that we will use often. A junction in a circuit is a point where three or more conductors meet. Junctions are also called nodes of branch points. A loop is any closed conducting path.

In figure 13.13 below, the points *a* and *b* are junctions, but points c and d are not. The curved lines show some possible loops in this circuits.

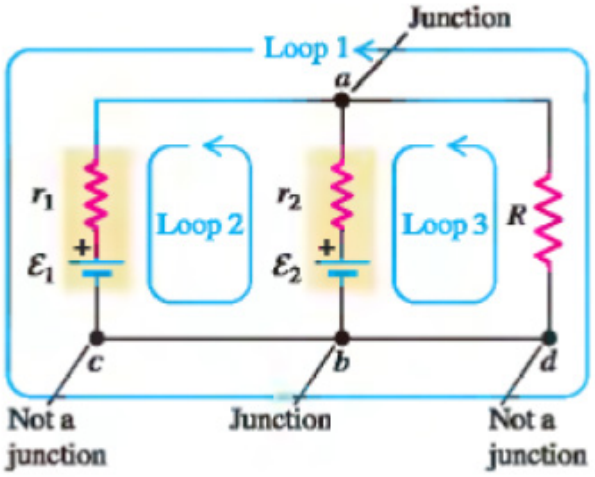


Figure 13.13: A circuit with loops

Kirchhoff's junction rule: The algebraic sum of the currents into any junction is zero. That is, $\sum I = 0$

i.e The sum of the currents entering the junction must equal the sum of the currents leaving the junction.

Kirchhoff's loop rule: The algebraic sum of the potential differences in any loop, including those associated with emfs and those of resistive elements, must equal zero. That is, $\sum V = 0$

Kirchhoff's first rule is a statement of conservation of electric charge. All charges that enter a given point in a circuit must leave that point because charge cannot build up at a point.

Kirchhoff's second rule follows from the law of conservation of energy.

The sum of the increases in energy as the charge passes through some circuit elements must equal the sum of the decreases in energy as it passes through other elements. The potential energy decreases whenever the charge moves through a potential drop $-IR$ across a resistor or whenever it moves in the reverse direction through a sourceof emf. The potential energy increases whenever the charge passes through abattery from the negative terminal to the positive terminal.

When applying Kirchhoff's second rule in practice, we imagine traveling around the loop and consider changes in electric potential, rather than the changes in potential energy.

Problem solving strategy

Junction rule: Assign symbols and directions to the currents in the various junctions. If you guess the wrong direction for a current it does not matter. The end result will be a negative answer for that current and the magnitude will be correct.

Loop rule: You must choose a direction for moving around the loop. As you move around the loop the voltage drops and increases should be recorded according to the rules (a-d) below.

a). If a resistor is traversed in the direction of the current, the change in potential across the resistor is $-IR$.

b). If a resistor is traversed in the direction opposite the current, the change in potential across the resistor is $+IR$.

c). If a source of emf is traversed in the direction of the emf (from – to +) the change in potential is $+\mathcal{E}$.

d). If a source of emf is traversed opposite the direction of the emf (from + to –) the change in potential is $-\mathcal{E}$.

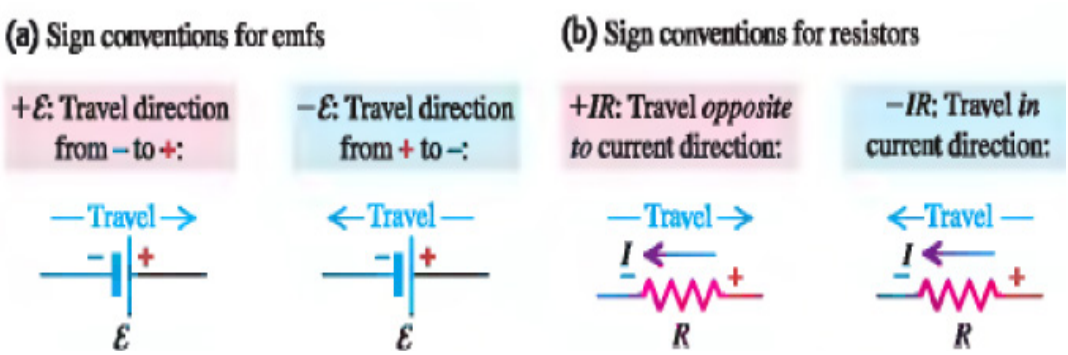


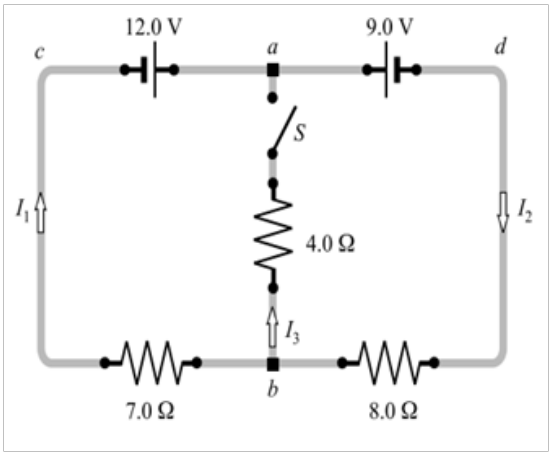
Figure 13.14: Sign conventions

Example

1. In figure below, find I_1, I_2, and I_3 if S is

   a). Open

   b). Closed



a). When S is open, $I_3=0$, because no current can flow through the open switch. Applying the node rule to point a gives

$I_1+I_3=I_2$ or $I_2=I_1+0=I_1$

Applying the loop rule to loop acbda gives

$-12.0+7.0I_1+8.0I_2+9.0=0$

Because $I_2=I_1$,

$15.0I_1=3.0$ *or* $I_1=0.20\ A$

Also $I_2=I_1=0.20\ A$

b). With S close, I_3 is no longer known to be zero. Applying the node rule to point a gives

$I_1+I_3=I_2$ (1)

Applying the loop rule to loop acba gives

$-12.0+7.0I_1-4.0I_3=0$ (2)

And to loop adba gives

$-9.0-8.0I_2-4.0I_3=0$ (3)

We must solve (1), (2) , and (3) for I_1, I_2, and I_3. From (3)

$I_3=-2.0I_2-2.25$

Substituting tis in (2) also gives

$-12.0+7.0I_1+9.0+8.0I_2=0$ or $7.0I_1+8.0I_2=3.0$

Substituting for I_3 in (1) also gives

I_1-2.0I_2-2.25=I_2 or I_1=3.0I_2+2.25

Substituting this value in the previous equation finally gives

21.0I_2+15.75+8.0I_2=3.0 or I_2=-0.44 A

Using this in the equation for I_1 gives

I_1=3.0(-0.44)+2.25=0.93 A

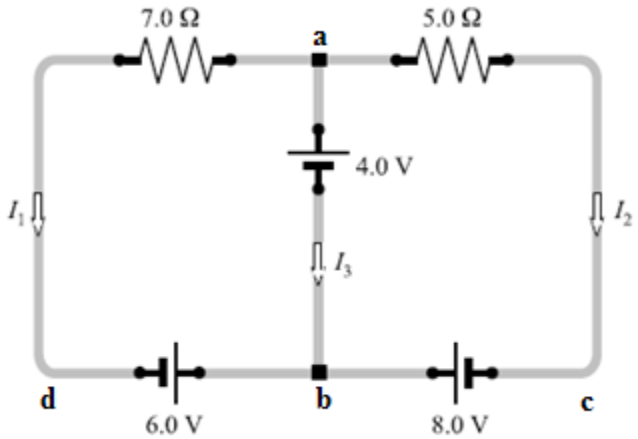
From (1)

I_3=I_2-I_1=(-0.44)-0.93=-1.37 A

13.5.2. Application of Kirchhoff's laws

Kirchhoff's laws are mainly applied in solving problems in complex electric circuits.

Example 1: Find the currents in the circuit given below:



Solution

This circuit cannot be reduced further because it contains no resistors in simple series or parallel combinations. We therefore revert to Kirchhoff's rules. If the currents had not been labeled and shown by arrows, we would do that first. No special care needed to be taken in assigning the current directions, since those chosen incorrectly will simply give negative numerical values.

We apply the node rule to node b in the figure above.

Current into b = Current out of b.

$$I_1 + I_2 + I_3 = 0 \qquad (1)$$

Next we apply the loop rule to loop adba. In volts,

$$-7.0I_1 + 6.0 + 4.0 = 0 \quad or \quad I_1 = \frac{10.0}{7.0} A$$

(why must the term $7.0I_1$ have a negative sign?) we then apply the loop rule to loop abca. In volts,

$$-4.0 - 8.0 + 5.0I_2 = 0 \; or \; I_2 = \frac{12.0}{5.0} A$$

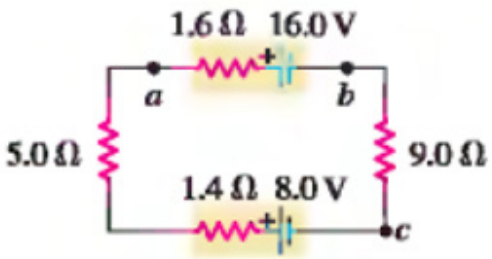
(why must the signs be as written?)

Now we return to equation (1) to find

$$I_3 = -I_1 - I_2 \; , \; I_3 = \left(\frac{10.0}{7.0} - \frac{12.0}{5.0}\right) A, \; I_3 = \frac{-50-84}{35} A \; and \; I_3 = -3.8A$$

The minus sign tells us that I_3 is opposite in direction to that shown in the figure.

Example 2: The circuit shown in figure below contains two batteries, each with an emf and an internal resistance, and two resistors. Find (a) the current in the circuit (magnitude and direction); (b) the terminal voltage V_{ab} of the 16.0 V battery; (c) the potential difference V_{ac} of a with respect to point c.



Solution

1. a) The current is counterclockwise, because the 16 V battery determines the direction of current flow.

$$+16.0V - 8.0V - I\left(1.6\Omega + 5.0\Omega + 1.4\Omega + 9.0\Omega\right) = 0$$

$$I = \frac{+16.0V - 8.0V}{\left(1.6\Omega + 5.0\Omega + 1.4\Omega + 9.0\Omega\right)} = 0.47\,A$$

b).

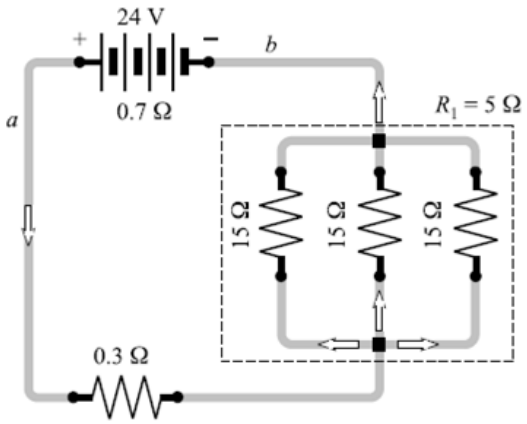
$$V_b + 16.0V - I(1.6\Omega) = V_a, so,$$
$$V_a - V_b = V_{ab} = 16.0V - (0.47\,A)(1.6\Omega) = 15.2V$$

c).

$$V_c + 8.0V - I(1.4\Omega + 5.0\Omega) = V_a, so,$$
$$V_{ac} = (5.0\Omega)(0.47\,A) + (1.4\Omega)(0.47\,A) + 8.0V = 11.0V$$

Example 3: In figure below, the battery has an internal resistance of 0.7Ω. Find:

i. The current drawn from battery,
ii. The current in each 15 Ω resistor,
iii. The terminal voltage of the battery.



Solution

i. For parallel group resistance R_1 we have

$$\frac{1}{R_1} = \frac{1}{15\Omega} + \frac{1}{15\Omega} + \frac{1}{15\Omega} = \frac{3}{15\Omega}$$
$$R_1 = 5.0\Omega$$

Then $R_{eq} = 5.0\Omega + 0.3\Omega + 0.7\Omega = 6.0\Omega$

And $I = \frac{\varepsilon}{R_{eq}} = \frac{24V}{6.0\Omega} = 4.0\,A$

ii. The three resistor combination is equivalent to R_1=5.0 Ω. A current of 4.0 A flows through it. Hence, the p.d across the combination is

IR_1=(4.0 A)(5.0 Ω)=20 V

This is also the p.d across each **15** Ω resistor. Therefore, the current through each **15** Ω resistor is

$$I_{15} = \frac{V}{R} = \frac{20.0V}{15\Omega} = 1.3\,A$$

In this special case, we know that one –third of the current will go through each resistor. Hence

$$I_{15} = \frac{V}{R} = \frac{4.0V}{3\Omega} = 1.3\,A$$

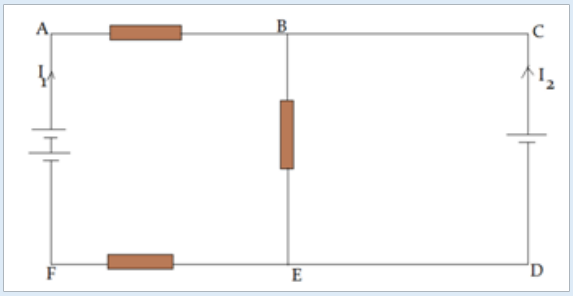
iii. We start at a and go to b outside the battery:

$$V\ from\ a\ to\ b = -(4.0A)(0.3\Omega) - (4.0A)(5.0\Omega) = -21.2V$$

The terminal p.d of the battery is $21.2V$. Or, we could write for this case of a discharging battery, $ter\min al\ p.d = \varepsilon - Ir = 24V - (4.0A)(0.7\Omega) = 21.2V$

Application activity 13.5

In the circuit below, each cell has e.m.f of 1.5 V and zero internal resistance. Each resistor has a resistance of 10 Ω. There are currents I_1 and I_2 in the branches as shown.



a). Use Kirchhoff's first law to write down an expression for the current in BE, in terms of I_1 and I_2

b). Use Kirchhoff's first law to write down equations for the circuit loops i) ABEFA ; ii) CBEDC

Skills lab 13

Conduct a survey to find out how people construct electric circuits and apply Kirchhoff's laws in analysis of complex electric circuits before installation process.

Collect and analyze data about when, where, and why people use Kirchhoff's laws in dealing with complex electric circuits.

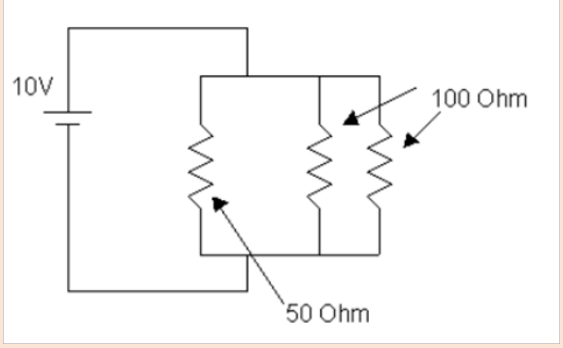
To complete this project you must

- Develop a survey sheet about electric components, demonstration of Kirchhoff's laws and complex electric circuit.
- Distribute your survey sheet to other student-teachers, family members and neighbors.
- Compile and analyze your data.
- Create a report to display your findings in your sheet.

Plan it! To get started, think about the format and content of your survey sheet. Brainstorm what kinds of questions you will ask. Develop a plan for involving student-teachers in your class or other classes to gather more data.

End unit assessment 13

1. State Kirchhoff Current Law and Kirchhoff Voltage Law.
2. Distinguish between electric sources of currents from receptors.
3. Find the current across the 10V battery



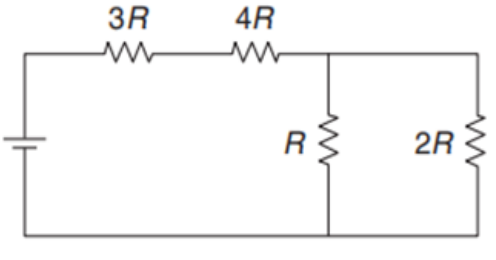
4. The figure below shows four resistors connected in a circuit with a battery.Which of the following correctly ranks the potential difference, across the four resistors?

(A) $\Delta V_{4R} > \Delta V_{3R} > \Delta V_{2R} > \Delta V_R$

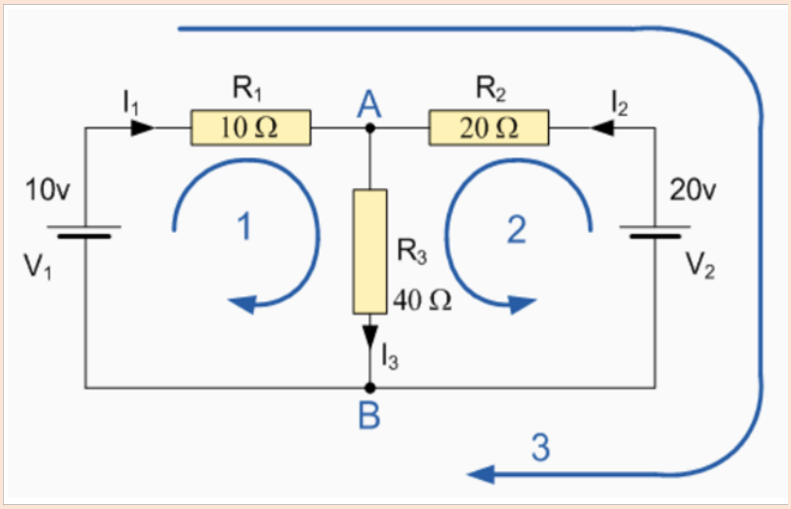
(B) $\Delta V_{4R} > \Delta V_{3R} > \Delta V_{2R} = \Delta V_R$

(C) $\Delta V_{4R} = \Delta V_{3R} > \Delta V_R > \Delta V_{2R}$

(D) $\Delta V_{2R} = \Delta V_R > \Delta V_{3R} > \Delta V_{4R}$



5. Determine the values of the the current flowing through each of the resistors.



UNIT 14:

ASEXUAL AND SEXUAL REPRODUCTION IN PLANTS

Key Unit competence: Describe modes of reproduction in plants and apply various methods of asexual and Sexual reproduction as means of increasing crop yield.

Introductory Activity 14

A.

The kingdom Plantae comprises about 260,000 known species including flowering and non-flowering plants. All plants have a general organization which includes vegetative and reproductive organs. Plants reproduce through different ways. Use the books and other source of information to do the following:

1. Write on how lower organisms such unicellular plant and another like cassava, sugar cane and apple reproduce.
2. Describe the techniques used by people to grow Irish potatoes, cassava and bananas.
3. Describe each of the following methods of asexual reproduction: fragmentation, budding and spore formation.

B.

1. Observe the following pictures and suggest what is going on.
2. How are the pictures below related to reproduction in flowering plants.

14.1. Comparison between sexual and asexual reproduction

Application activity 14.1

Observe the figures below taken from environment and answer the following questions.

A

B

a). Name the figure A and B.

b). Where the smallest (young ones) of banana come from?

c). How the young bananas have been reproduced?

d). How the young one of goat has been reproduced?

e). Compare the modes of reproduction of banana plant and a goat.

Asexual Reproduction

Prokaryote like bacteria and unicellular organisms reproduce asexually by cell division or binary fission of the parent cell. This is a type of reproduction done by a single organism without production of gametes. It usually results in the production of identical offspring, the only genetic variation arising as a result of random mutation among the individuals.

Sexual Reproduction

Sexual reproduction is a type of reproduction in which two parents are involved, each capable of producing gametes. It is essentially cellular in nature which involves the fertilization of one haploid gamete with another, producing a new diploid cell called the zygote. The union of structurally similar physiologically different gametes is called isogamy. This is found only in lower forms such as Protozoa. The following is summary table of difference between asexual and sexual reproduction.

Table 14.1: Comparison between asexual reproduction and sexual reproduction

Asexual reproduction	Sexual reproduction
One parent only	Usually two parents
No gametes are produced	Gametes are produced. These are haploid and nuclei of two gametes fuse (fertilization) to form a diploid zygote.
Depends on mitosis	Depends on meiosis being present at some stage in life cycle to prevent chromosome doubling in every generation.
Offspring identical to parent	Offspring are not identical to parents. They show genetic variation as a result of genetic recombination
Commonly occurs in plants, simple animals and microorganisms; absent in more complex animals	Occurs in almost all plant and animal species
Often results in rapid production of large numbers of offspring	Less rapid increase in numbers

Application activity 14.1

1. **Distinguish sexual and asexual reproduction.**

14.2. Asexual reproduction in plants

14.2.1. Asexual reproduction and its methods

Activity 14.2.1.a

Brainstorm on asexual reproduction in lower organisms and write reports as an out-of-class activity.

There are five common modes of asexual reproduction: **fission, budding, vegetative reproduction, spore formation and fragmentation.**

1. Fission

An important form of fission is binary fission. In binary fission, the parent organism is replaced by two daughter organisms, because it literally divides in two. Organisms, prokaryotes, and eukaryotes (such as unicellular fungi), reproduce asexually through binary fission; most of these are also capable of sexual reproduction. Another type of fission is multiple fission that is advantageous to the plant life cycle. Multiple fission at the cellular level occurs in many algae. The nucleus of the parent cell divides several times by mitosis, producing several nuclei. The cytoplasm then separates, creating multiple daughter cells.

2. Budding

Some cells split via budding resulting in a 'mother' and 'daughter' cell. The offspring organism is smaller than the parent. Budding is also known on a multicellular level.

3. Vegetative reproduction

Vegetative reproduction is a type of asexual reproduction found in plants where new individuals are formed without the production of seeds or spores by meiosis or syngamy. Examples of vegetative reproduction include the formation of miniaturized plants called plantlets on specialized leaves (for example in kalanchoe) and some produce new plants out of rhizomes or stolon (for example in strawberry). Other plants reproduce by forming bulbs or tubers (for example tulip bulbs and dahlia tubers). Some plants produce adventitious shoots and suckers that form along their lateral roots.

Plants that reproduce vegetative may form a clonal colony, where all the individuals are clones, and the clones may cover a large area.

4. Spore formation

Many multicellular organisms form spores during their biological life cycle in a process called sporogenesis. Plants and many algae on the other hand undergo sporic meiosis where meiosis leads to the formation of haploid spores rather than gametes. These spores grow into multicellular individuals (called gametophytes in the case of plants) without a fertilization event. Then in the plant life cycle. Fungi and some algae can also utilize true asexual spore formation, which involves mitosis giving rise to reproductive cells called mitospores that develop into a new organism after dispersal.

Activity 14.2.1.b

Demonstration of asexual reproduction by fragmentation in algae

Requirements

Glass beakers of 500ml, Scalpel, Forceps, Pins, Spatula, Weighing balance, Labels, Artificial, fertilizers, Clear river water and Spirogyra (algae).

Procedure

1. Label five beakers of the same size as A, B, C, D and E. Pour water in each beaker. Weigh several measures of artificial fertilizers of 1 g each.
2. Transfer 1g of fertilizer to beaker A, then 2g to beaker B, 3g to beaker C, 4g to beaker D Then lastly put 5g of fertilizers to beaker E. Note the concentration of fertilizers is increasing from A- E.
3. Using forceps pick spirogyra and put it on a tile. Add several drops of water to avoid drying. Tease off a piece of spirogyra using a pin. Cut that piece into 5 fragments of the same length and transfer each piece into the beaker.
4. Stand the beakers in a place where they can receive adequate sunlight for the seven days. On the next day start to examine the fragments in each beaker every day and record any observable changes such the increase in size of the spirogyra.

   Draw a table as this shown here and record your observation.

Table16.2: Rate of growth

Beaker	Rate of growth	Reason
A		
B		
C		
D		
E		

5. Fragmentation

Fragmentation is a form of asexual reproduction where a new organism grows from a fragment of the parent. Each fragment develops into a mature, fully grown individual. Fragmentation is seen in many organisms such as fungi, and plants. Some plants have specialized structures for reproduction via fragmentation, such as gemma in liverworts. Most lichens, which are a

symbiotic union of a fungus and photosynthetic algae, reproduce through fragmentation to ensure that new individuals contain both symbionts. These fragments can take the form of soredia, dust-like particles consisting of fungal hyphen wrapped around photobiont cells.

Figure 14.1: Bryophyllum, a plant that reproduces asexually via new shoots from the leaves

Application activity 14.2.1

1. Write on the types of asexual reproduction.
2. Explain the term fragmentation and give one example of plant which reproduces by using this type of asexual reproduction.

14.2.2. Advantages and disadvantages of asexual reproduction

Activity 14.2.2

Make discussion on asexual reproduction in lower organisms and higher plants, outlining advantages and disadvantages.

a. Advantages of asexual reproduction

- As it occurs by mitosis,the offspring are genetically identical not only to their parents but also amongst themselves. This is advantageous, for example in agriculture where successive generations of plant retain the desired qualities or high yield of flowers, fruits and cereals.
- It is faster as it does not involve stages such development of gonads, gamete formation, fertilization, fruit and seed formation and their dispersal. So plant can be rapidly propagated vegetatively using cuttings, bulbs, offsets and other parts.

b. Disadvantages of asexual reproduction

- It is suited to the environments that are generally stable in terms of humidity, soil, temperature and light. If there is a drastic change in the environment, the individuals may not be able to cope with the change and so gradually die off.
- It can lead to overcrowding of the individuals plants around the parent plant and this can lead to excessive competition for soil nutrients, light and moisture. The new individuals will therefore become weak and only a few may survive.

Application activity 14.2.2

Explain the advantages and disadvantages of asexual reproduction.

14.2.3. Vegetative and artificial propagation in flowering plants

Activity 14.2.3

Demonstration of asexual reproduction in plants by cuttings

Requirements

Growth medium or moist soil, sweet potatoes vines, elephant grass, sugarcane or cassava stems, secateurs/sharp knife and rooting hormone.

Procedure

1. Collect clean and healthy stems from cassava, sugarcane or potato plants.
2. Using a secateurs/sharp knife, cut the stem of either cassava, sugarcane or sweet potato stems into suitable sizes.
3. Place them in either suitable medium of growth or apply rooting hormone if available or plant them in moist soil in the school garden.
4. Leave the set up for about 13 days, and then observe the development of roots and leaves at nodes.

Draw and record what you will observe after 13 days on the development of roots and leaves at nodes.

Artificial vegetative propagation is the deliberate production of new plants from parts of old plants by humans. This can be done by following three methods: Cutting, layering, and grafting.

a. Cutting

This is simple procedure in which part of the plant is removed by cutting and placed in a suitable medium for grow. The part of the plant which is removed by cutting it from the parent plant is called a 'cutting'. In this method one-year-old stem of root is cut from a distance of 20 to 30 cm. and is buried in the moist soil in natural position. After sometime, roots develop from this cutting and it grows into a new plant. This method is commonly used in rose and sugar cane. Care is taken that nodes which were lower in parent plant (morphologically) are put in the soil, while the morphologically higher nodes are kept up. Adventitious roots are given off at the lower nodes.

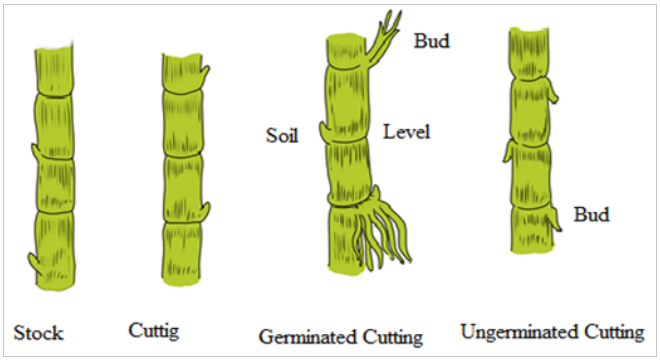


Figure 14.2. cutting technique

b. Layering

This method of vegetative propagation is used in those plants whose soft branches occur near the ground such as jasmine plant. In this method, a branch of the plant which is near to the ground is pulled towards the ground and a part of this branch is covered with moist soil leaving the tip of this branch above the ground. After sometime, roots develop from that part of the branch which was buried in the soil. This branch is then cut of along with the roots from the parent plant and develops into a new plant. This method of asexual reproduction is also used in the production of plants such as Bougainvillea, jasmine, guava, strawberries, lemon, China rose etc.

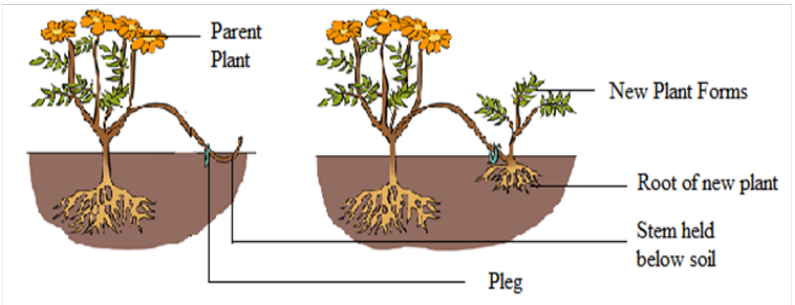


Figure 14.3: Vegetative propagation by layering.

c. Grafting

In this method of vegetative propagation the stems of two different plants are joined together so as to produce a new plant containing the characters of both plants. Out of the two plants one plant has a strong root system while the other has a better flower or fruit yield. The plant of which the root system is taken is called 'stock', while the other plant of which the shoot is selected is known as 'scion' or 'graft'. These two stems i.e. the stock and the scion are fitted together by making slanting cuts in them and bound tightly with a piece of cloth and is covered with a polythene sheet.

While joining the scion with the stock care should be taken that the diameter of the stock and scion chosen for grafting should be equal. Scion gets the mineral and water from the soil through the stock and develops branches and produce fruits. This method of propagation is used in mango, apple, banana, pear, grape, pineapple and peach.

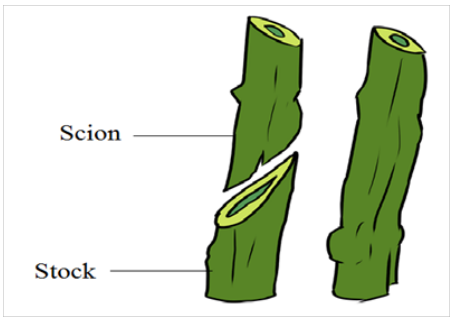


Figure 14.4: Vegetative propagation by grafting

Vegetative and artificial reproduction in flowering plants

The reproductive part of the plant is a flower. The union of male and female gametes to form a zygote is called fertilization. The transfer of pollen grains from the anther to the stigma of the same flower or the different flower is called pollination. In nature, plants reproduce asexually in a variety of ways. The vegetative reproductive parts in flowering plant are stem, branches, and leaves and they have the following characteristics:

Characteristics of the stem

- Stem develops from the plumule of embryo.
- Stem is generally the ascending part of the plant axis.
- It bears a terminal bud for growth in length.
- The stem is differentiated into nodes and internodes. The stem nodes possess dissimilar appendages called leaves.

- The young stem is green and capable of performing photosynthesis.
- In the mature state it bears flowers and fruits. Leaves and stem branches develop exogenously.
- Stem exposes leaves, flowers and fruits to their most suitable position in the aerial environment for optimum function, Hair, if present, is commonly multicellula.

- Stems are usually positively phototropic, negatively geotropic and negatively hydrotropic.

Characteristics of the leaf

- It is dissimilar lateral flattened outgrowth of the stem.
- The leaf is exogenous in origin.
- It is borne on the stem in the region of a node.
- An axillary bud is often present in the axil of the leaf.
- Leaf has limited growth. An apical bud or a regular growing point is absent.
- The leaf base may possess two lateral outgrowths called stipules.
- A leaf is differentiated into three parts: leaf base, petiole and lamina.
- The lamina possesses prominent vascular strands called veins.
- It is green and specialized to perform photosynthesis.
- Leaf bears abundant stomata for exchange of gases and it is the major seat of transpiration.

Characteristics of the branches

A branch or tree branch is a woody structural member connected to but not part of the central trunk of a tree. Large branches are known as boughs and small branches are known as twigs. Due to a broad range of species of trees, branches and twigs can be found in many different shapes and sizes.

Application activity 14.2.3

1. Discuss on the methods of artificial vegetative propagation.
2. Cassava produces flowers, fruits and seeds. Why people prefer to grow cassava by cutting rather than using seed?

14.2.4. Application of artificial propagation in growing improved varieties of plants

Activity 14.2.4

Using addition resources to your textbook available in your school such as the books from the school library and search further information from the internet. Discuss on application of artificial propagation in growing improved varieties of plants.

Artificial vegetative propagation is usually used in agriculture for the propagation (or reproduction) of those plants which produce either very few seeds or do not produce viable seeds. Some examples of such plants which are reproduced by artificial vegetative propagation methods are: Banana, Pineapple, Orange, Grape, Rose, etc.

Vegetative propagation of particular cultivars that have desirable characteristics is very common practice. Reasons for preferring vegetative rather than sexual means of reproduction vary, but commonly include greater ease and speed of propagation of certain plants, such as many perennial root crops and vines. Another major attraction is that the resulting plant amounts to a clone of the parent plant and accordingly is of a more predictable quality than most seedlings. However, as can be seen in many variegated plants, this does not always apply, because many plants actually are chimeras and cuttings might reflect the attributes of only one or some of the parent cell lines.

Man-made methods of vegetative reproduction are usually enhancements of natural processes, but they range from rooting cuttings to grafting and artificial propagation by laboratory tissue culture. In horticulture, a “cutting” is a piece that has been cut off from a mother plant and then caused to grow into a whole plant. A popular use of grafting is to produce fruit trees, sometimes with more than one variety of the same fruit species growing from the same stem. Rootstocks for fruit trees are either seedlings or propagated by layering.

Figure 14.5: Mass propagation of eucalyptus seedlings

Application activity 14.2.4

1. **Describe the methods of artificial vegetative propagation.**
2. **Elaborate on the characteristics of vegetative reproductive parts in a flowering plant**
3. **Explain the application of artificial propagation in growing improved varieties of plants.**

14.3. Sexual reproduction in plants

14.3.1. Alternation of generation in bryophytes and pteridophytes

Activity 14.3.1

Visit library and computer lab and search the comparison of the lifecycles of mosses and ferns and present them on Manilla paper.

The life cycle of an organism is the progressive sequence of changes which an organism goes through from the moment of fertilization to death. During its life cycle, the organism produces new generations of individuals which repeat continuously the process. New generations are produced by reproduction, which may be sexual or asexual. The life cycle involves the mitosis and meiosis. In this lesson, we will make concern on how meiosis can affect the life cycle of living organisms.

The life cycle is seen in seaweeds, mosses ferns and their relatives. Their life cycles start with the sexually mature adult plants. Since they produce gametes, they are called the gametophytes (gamete plants). These are haploid, as they produce eggs and sperms. The egg and sperm fuse to produce a diploid zygote, but this does not develop directly into a new gametophyte. Instead it grows (by mitotic divisions) into another plant which is quite distinct from the gametophyte called sporophyte (spore plant).The function of sporophyte is to produce spores.In bryophytes, the sporophyte depends on gametophyte for nourishment. As spores are formed by meiosis, they are haploid. When the later are dispersed by wind on suitable soil, they germinate and grow by mitosis into gametophytes, which then repeat the sequence of life cycle. Ferns and mosses consist of two distinct plants: haploid gametophyte and diploid sporophyte, which alternate each other within the life cycle. This phenomenon is known as alternation of generation.

So, Alternation of generation is a phenomenon in the plant life cycle in which a diploid stage a sporophyte alternates with a haploid stage of gametophyte.

The importance of alternation of generation to organisms

- Spores produced can survive hash conditions and only germinate when conditions are favorable.
- It ensures rapid multiplication of the plant species as spores are usually produced in vast numbers.
- Interdependence between the gametophyte and sporophyte generations ensure that both generations exist at any given time. This prevents extinction of the plant species.

a. Alternation of generations in bryophytes (mosses)

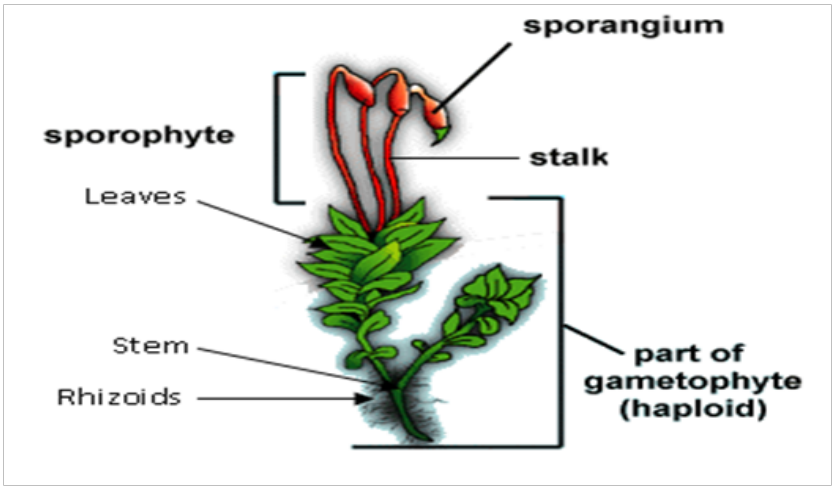


Figure 14. 6: Diagram of a moss

The life cycle in mosses involves alternation of generations between the diploid (2n)sporophyte and haploid (n) gametophyte. In bryophytes, the gametophyte is adominant stage of the life cycle, while the sporophyte is dependent on gametophyte for supplying water and nutrients. Gametes are formed in special reproductive organs at the tips of gametophytes.Sperms are produced in antheridia (singular: antheridium), and eggs in archegonia (singular: archegonium). Some species produce both sperms and eggs on the same plant, whereas others produce sperms and eggs on separate plants.

During fertilization which requires water, the sperms released from antheridia fuse with the eggs and form a diploid (2n) zygote. The zygote grows into the sporophyte. Sporophyte is a long stalk ending in a capsule in which haploid(n) spores are formed by meiosis. When spores become mature

enough, the capsule bursts and spores are scattered by wind. Under suitable conditions, spores then germinate by forming underground filaments called protonemata (singular: protonema). Small buds produced by protonemata give rise to new gametophyte plants which can start the cycle again.

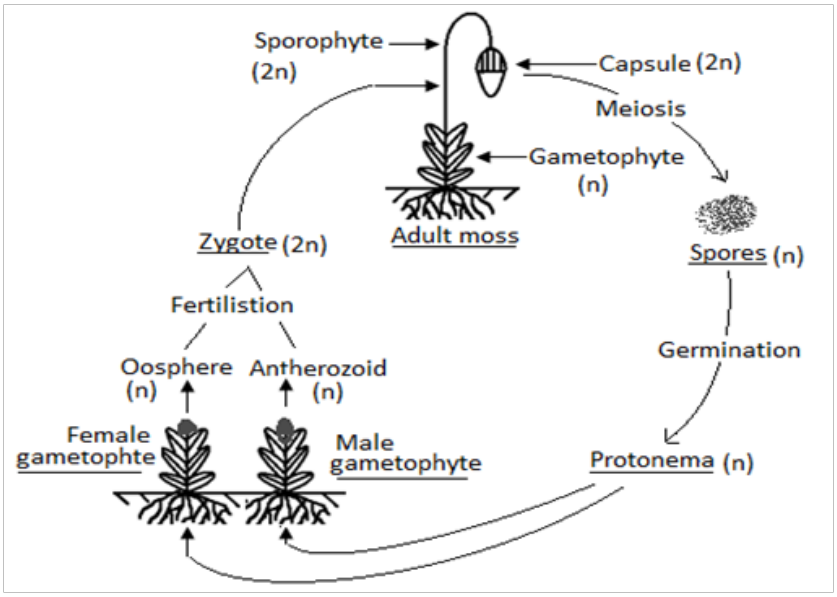


Figure 14.7: Alternation of generations in bryophytes (mosses).

Bryophytes grow in habitats where water is available constantly because for fertilization to occur, the sperm of bryophyte must swim to an egg. Without water, this movement is impossible. Because of this dependence on water for reproduction, bryophytes must live in habitats where water is available at least part of the year.

b. Alternation of generations in pteridophytes (ferns)

Pteridophytes also exhibit alternation of generations. Ferns are formed of true roots, stem and leaves (fronds) with vascular tissues. They have lignified tissues. The horizontal underground stem is called rhizome which bears adventitious roots. The leafy plant is a sporophyte. Mature leaves commonly called fronds bearing yellow or orange masses of sporangia which are grouped into the structures called sori (singular: sorus) on their lower side.

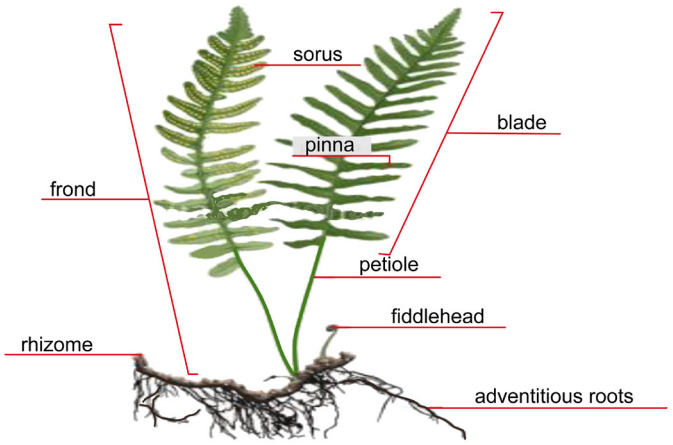


Figure 14.8. Diagram of ferns

Ferns have a life cycle in which the diploid (2n) sporophyte is the dominant generation, large (some fern trees can have 7 m of height), and differentiated into leaf, stem and roots with vascular tissues, while the haploid gametophyte (prothallus) is very simple with few millimeters. In ferns, the sporophyte produces haploid spores by meiosis. This is done on the underside of the leaf called frond, in sporangia (singular: sporangium). Sporangia are grouped into sori (singular: sorus).

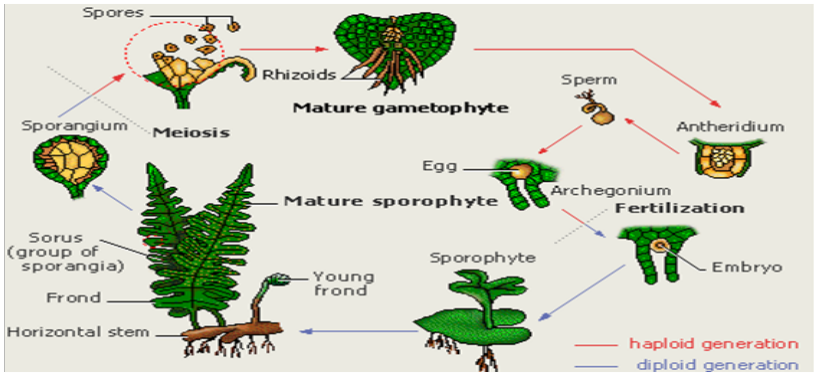


Figure 14.9: Alternation of generations in pteridophytes (ferns)

When spores are mature enough, the sporangia burst and spores are released on ground. If conditions are favorable, spores germinate and grow into the haploid heart-shaped gametophytes (n), which grows independently of sporophyte. Antheridia and archegonia found underside of gametophyte produce sperms and eggs respectively. During fertilization, sperms swim towards eggs and fuse together to form diploid zygotes, which grow and develop into new sporophytes. As sporophyte grows, the gametophyte dries and dies.

Differences between the mosses and ferns

Table 14.2: Comparison between moss and fern.

Mosses	Ferns
Have no vascular system	Have vascular system
No leaves, stems and roots	True leaves (fronds), stems, and roots
Gametophyte is dominant generation	Sporophyte is dominant generation
Spores in sporangia do not form clusters	Sporangia form clusters called sori
Transfer of sperm from antheridia to the archegonia rosette occurs by rain-splash	Sperm swim in water from the antheridia to an egg cell at the base of the archegonium

Application activity 14.3

1. Explain the meaning of the term alternation of generation.
2. How is water essential in the life cycle of a bryophyte?
3. What is the archegonium and antheridium?
4. How are these structures important in the life cycle of a moss plant?
5. What is the dominant stage of the fern life cycle?
6. Explain the relationship of the fern gametophyte and sporophyte.
7. Compare the gametophyte and sporophyte stages of the plant cycle. Which is haploid? Which is diploid?
8. How do bryophytes reproduce asexually?

14.3.2. Types, structure and functions of flowers

Activity 14.3.2

Collect different forms of flowers from the school compound or around the school, such as hibiscus, morning glory, sweet potato, or maize flower (or any type of flower in your community not necessarily the ones mentioned here)

1. Observe and describe the structures of collected flowers.
2. How do collected flowers differ externally?
3. Cut one of the flowers into two halves, draw and label one half of flower.

a. Structure of a typical complete flower

A flower is a reproductive organ of a plant, which produces fruits and seeds.

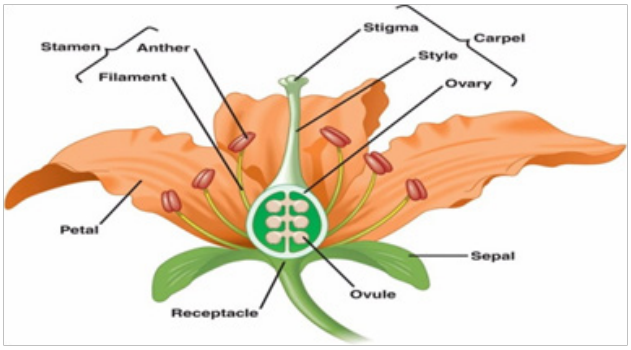


Figure 14.10: Diagram of a complete flower

A typical hermaphrodite or bisexual flower contains the following parts:

- **Pedicel**: It is the stalk which attaches the flower on the main floral axis.
- **Receptacle**: It is the swollen part at the end of the stalk where other parts of the flower are attached.
- **The calyx**: It is the set of sepals, generally having green colour. They protect the internal parts of the flower. In some plants, the sepals are coloured and are called petaloids.
- **The corolla**: It is the set of petals, with different colours and nectar glands that produce sugary substances which participate in attraction of pollinating agents. In some plants, the petals are green and are called sepaloids. Both calyx and corolla are collectively called perianth. They are called floral envelope or accessory organs as they do not participate directly in reproduction, or in formation of fruits and seeds, they all insure the protection of internal parts of the flower.
- **Androecium**: It is the male reproductive organs of the flower. It consists of many stamens. A stamen consists of: the filament which supports anther, and anther which contains the pollen grains or male gametes.
- **Gynoecium/pistil**: It is the female reproductive organ. It consists of many carpels, and each carpel is made of: stigma (plural: stigmata), style and ovary with ovules.

  a). The stigma: Receive pollen grains from anther during pollination.

  b). Style: Supports the stigma in a good position to receive pollen grains.

  c). Ovary: A sac where ovules are produced. Ovules become seeds after fertilisation.

b. Types of flowers

1. According to absence of some reproductive parts of the flower, we can distinguish:
   - i. **Unisexual flower**: A flower that consists of one type of reproductive organ. This can be staminate: unisexual male (with androecium only), or carpellate: unisexual female (with gynoecium only). E.g. flower of papaya.
   - ii. **Bisexual or hermaphrodite flower**: A flower with the two reproductive organs. It contains both male and female reproductive organs (androecium and gynoecium). E.g. flowers of beans.

   Dioecious plants are plants that have male flowers and female flowers on separate plants (e.g. papaya/pawpaw) while monoecious plants are plants that have both male and female flowers on the same plant (e.g. maize).

2. According to the position of ovary in the point of insertion of calyx, corolla and stamen, we can distinguish:
   - i. **A flower with inferior ovary:** It is when the ovary is located below the point of insertion of calyx, corolla and stamens.
   - ii. **A flower with superior ovary**: It is when the ovary is located over the point of insertion of calyx, corolla and stamens.
   - iii. **The semi-infer or semi-super flower**: When ovary is neither infer nor super but in the middle of receptacle which is hollowed.
     - When sepals are joined together, the flowers are called gamosepal, and where are not joined together, the flower is called dialysepal.
     - When petals are joined together, the flowers are called gamopetal, and when are not joined together, the flower is called dialypetal. When they are absent, the flower is called apetal.
3. According to the shape and symmetry of the flower, we can distinguish:
   - i. **Zygomorphic or irregular flower**: A flower with a bilateral symmetry. The flower cannot be divided into two similar halves. E.g. flowers of beans, cassia.
   - ii. **Actinomorphic or regular flower:** A flower with a radial symmetry. The flower can be divided into two or more planes to produce similar halves. E.g. flowers of coffee, orange.

4. **Dichogany**: It is when male and female organs of the flower mature at different times. We can distinguish:
   - **Protandry**: When stamens mature before pistil.
   - **Protogyny**: When pistil matures before stamen.
5. **Inflorescence** is when two or more flowers borne on a common stalk.

c. Representation of the number and characteristics of a flower

There are two ways by which we can present the number and characteristics of different parts of the flower. These ways include: floral diagram and floral formula.

Floral formula

The floral formula indicates the number and characteristics of different floral organs. It varies from one flower to another. By convention, there are standard symbols that are used to represent different parts of the flower and their characteristics:

K: for calyx, K5: calyx with five free sepals, K(5): calyx with five fused sepals.

C: for corolla, C5: corolla with five free petals, C(5): corolla with five fused petals.

P: for perianth, P4: four free tepals, P(4): four fused tepals, P2+2: four tepals in two whorls of three each.

A: for Androecium or stamen, A5: androecium with five free stamens, A(5): five fused stamens, A5+5: ten stamens in two whorls of five each, A0: stamens absent, A∞: stamens indefinite in number, A(9)+1: androecium of 10 stamens nine fused together and one free.

G: for pistil or gynoecium, G_2: two free carpels, G_2: two fused carpels, G0: carpels absent, G_2: Bicarpellary, syncarpous semi-inferior ovary.

Representation of the symmetry of flower:

Φ: zygomorphic or irregular

⊕: Actinomorphic or regular flower.

Representation of sex of the flower: ♂: staminate flower, ♀: pistillate flower,

Bisexual flower

The floral formula is specific to each species of plant.

Examples

Write a floral formula of coffee having

- The bilateral symmetry
- Hermaphrodite flower
- 1 calyx with 5 fused sepals
- 1 corolla with 5 petals
- 5 fused stamens
- 1 pistil with 2 carpels each one with infer ovary of 2 chambers

Answer: (5S)+ (5P) +(5A)+ 2C-2 or K(5) + C(5) + A(5) + G-2(2)

Write a floral formula of Irish potato having

- 5 free sepals
- 5 free petals
- 5 free stamens
- 2 fused carpels with 2 chambers having many ovules
- 1 infer ovary
- Bisexual flower
- Radial symmetry

Answer: $K_{(5)} + C_{(5)} + A_{5} + G^{-2}{}_{(2)}\infty$

Application activity 14.3.2

1. What are the male and female structures of a flower?
2. How might be an advantage for a plant to have many flowers together in a single structure?
3. Where does the female gametophyte develop?
4. Describe the flower and how it is involved in reproduction.

14.3.3. Pollination and double fertilisation in flowering plants

Activity 14.3.3

Visit a computer lab, download and describe the videos of the process of double fertilization in flowering plants, and search more information to identify different pollinating agents

1. Pollination

Pollination is transfer of pollen grains from anther to the stigma.

a). Types of pollination: There are two types of pollination such as: self-pollination and cross-pollination.

i. **Self-pollination**: It is the transfer of pollen grains from anther to the stigma of the same flower, or of different flowers but of the same plants. It involves one plant. E.g. flowers of maize and beans.

ii. **Cross-pollination**: It is the transfer of pollen grains from anther to the stigma of the flower of another plants. It involves two plants. E.g. flowers of pawpaw.

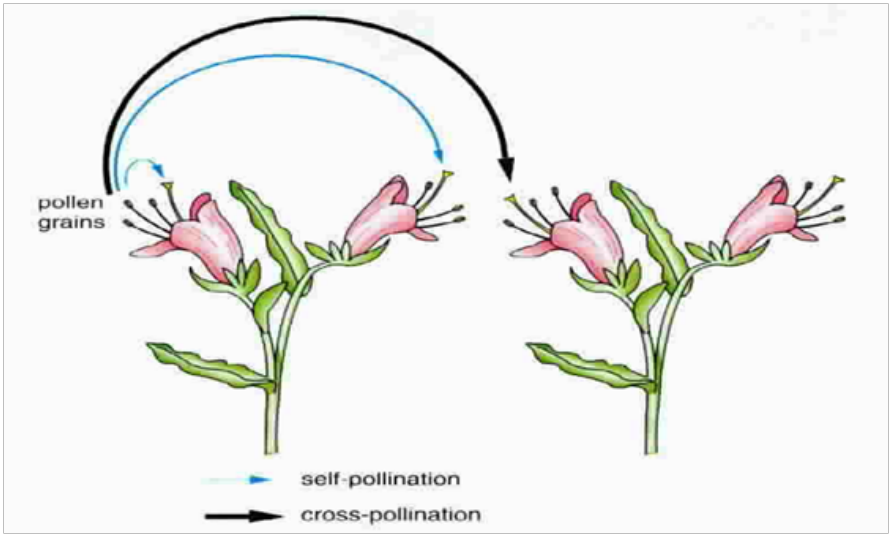


Figure 14.11: Self and cross-pollination

b). Main Pollinating agents

Flower structure is closely related with the way they are pollinated. This means that flowers are adapted to specific agents or mode of pollination. The common agents of pollination are: insects (entomophily), wind (anemophily), water (hydrophily), humans (anthropophily), and birds (ornithophily).

Characteristics of insect-pollinated flowers (entomophilous flowers)

- Flowers produce the nectar to attract pollinators.
- Flowers have large brightly coloured corolla to attract pollinators.
- Production of scents to attract pollinators.
- The surface of the stigma should be sticky to hold pollen grains.
- Pollen grains are sticky and rough enough to remain on the surface of stigma.

Characteristics of wind-pollinated flowers (enemophilous flowers)

- The flowers have large stigma to hold pollen grains.
- The surface of the stigma should be sticky to hold pollen grains.
- Pollen grains are rough enough to remain on the surface of stigma.
- The flowers are or are not brightly-colored.
- They have or do not have scent.
- They do or do not secrete nectar.
- They produce large quantities of pollen grains, as much of them never reach the stigmas.

2. Double fertilization and events after fertilization in flowering plants

Double fertilization is a complex fertilization mechanism of flowering plants (angiosperms). This process involves the joining of a female gametophyte (megagametophyte, also called the embryo sac) with two male gametes (sperm).

a. Development of pollen grains and plant ovules

i. Development of pollen grains

The pollen grains are produced in the anthers while the ovules are produced in the ovary.

Pollen grains

Each anther has four pollen sacs which contain many diploid microspore mother cells that undergo meiosis to form four microspores each. At first, the four microspores remain together as tetrads. The nucleus of each microspore then divides by mitosis, forming a generative nucleus and a tube or vegetative nucleus. At this point, the content of the pollen grain may be considered as the male gametophyte. A two-layered wall forms

around each pollen grain. The outer wall, the exine is thick and sculptured. The inner wall, the intine is thin and smooth. There are many pores or apertures in the wall through which a pollen tube may emerge.

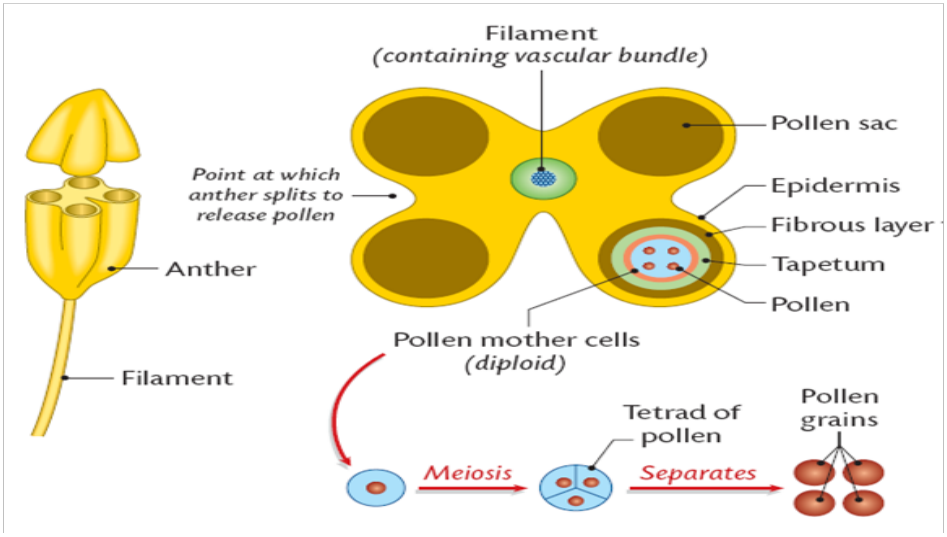


Figure 14.12: Development of pollen grains

ii. Development of Plant ovule

Each ovule is attached to the ovary wall by a short stalk called funicle. The main tissue in the ovule is the nucellus which is enclosed and protected by the integuments.

At one end of the ovule, there is a small pore called micropyle. A single diploid megaspore mother cell in the nucellus undergoes meiosis, producing four megaspores. Three of the four megaspores degenerate, while the remaining cell, called the embryo sac, grows to many times its original size. The nucleus of the embryo sac divides mitotically three times, resulting in eight haploid nuclei which are arranged in groups of four nuclei at the two poles. At this point, the contents of the embryo sac may be regarded at the female gametophyte.

One nucleus from each pole migrates to the centre of the embryo sac. These two nuclei are called polar nuclei, and they fuse to form a single diploid nucleus. Meanwhile, cell walls form around the remaining six nuclei and they form the synergids, antipodals and the egg (ovum). Only the egg functions as the female gamete.

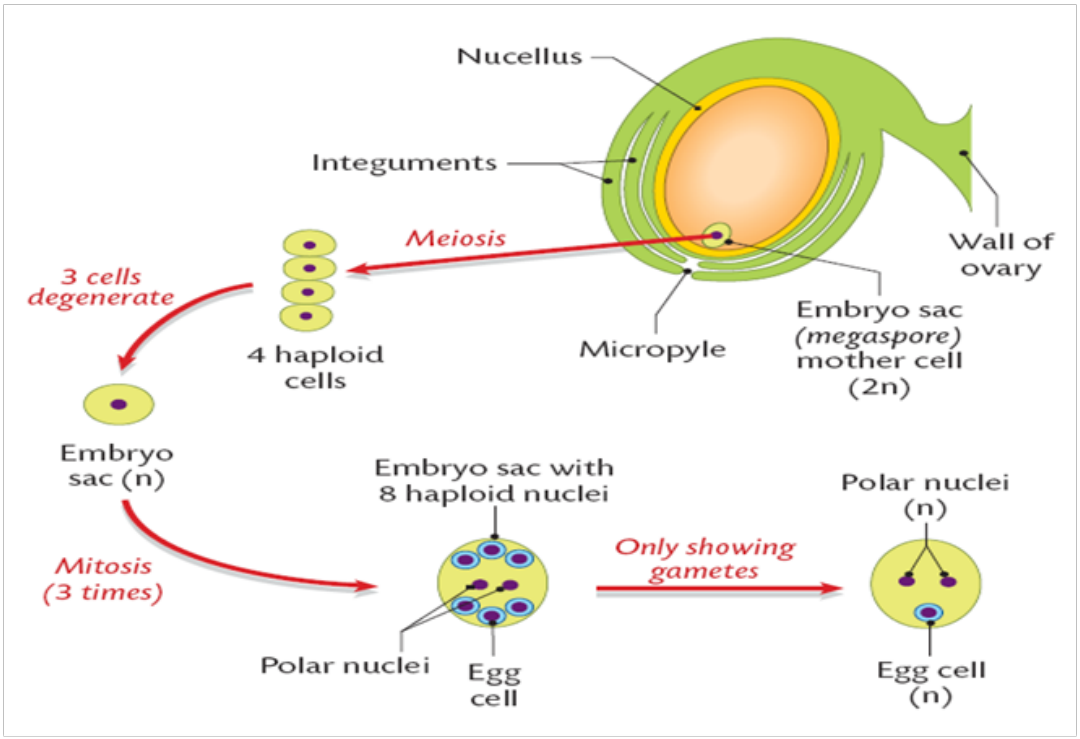


Figure 14.13: Development of plant female gamete

In summary, the pollen grain contains two haploid nuclei: one called generative nucleus, and the other the tube nucleus.

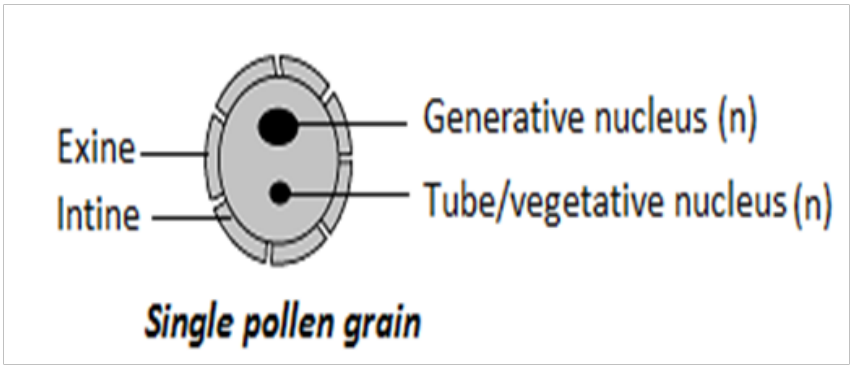


Figure 14.14: A pollen grain

On the other hand, the ovule or embryonic sac contains eight nuclei:

- Three antipodal nuclei/cells at one end
- Two polar nuclei/cells in the middle of ovule
- Two synergids (non-functional nuclei)
- One big egg cell.

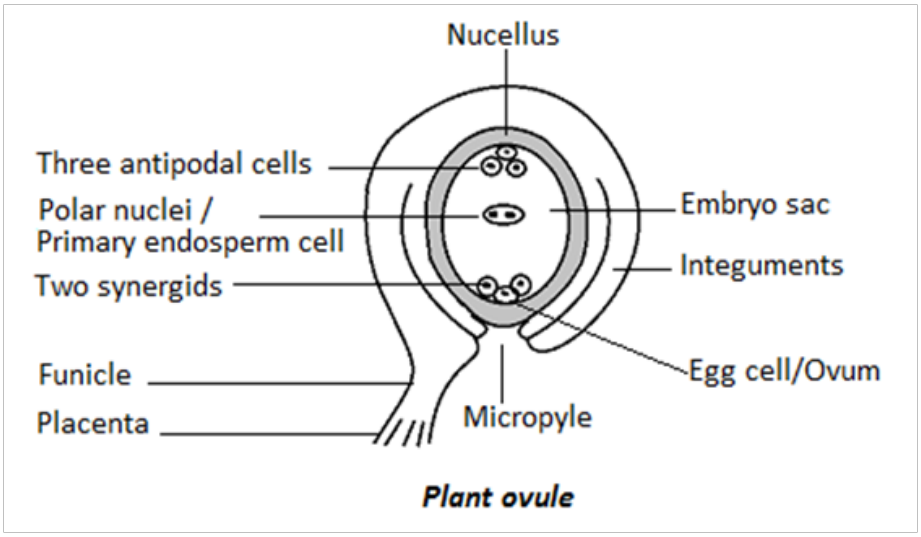


Figure 14.15: Structure of an ovule

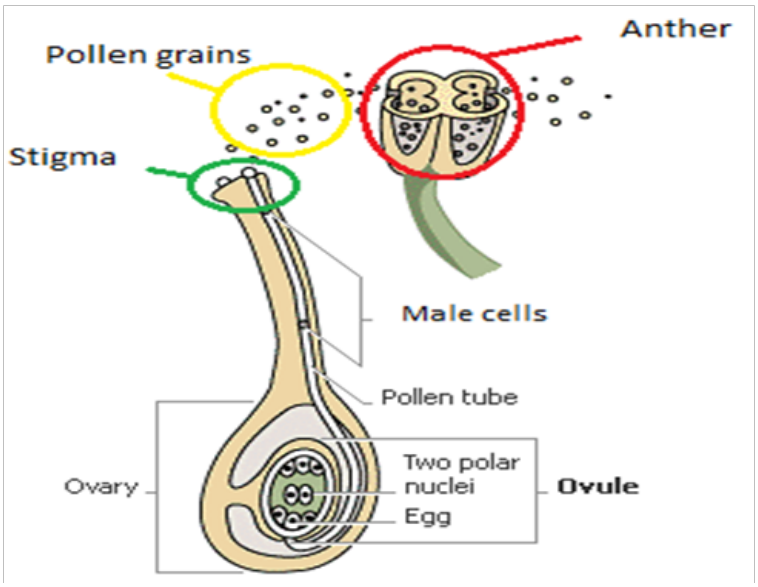
The process of double fertilization

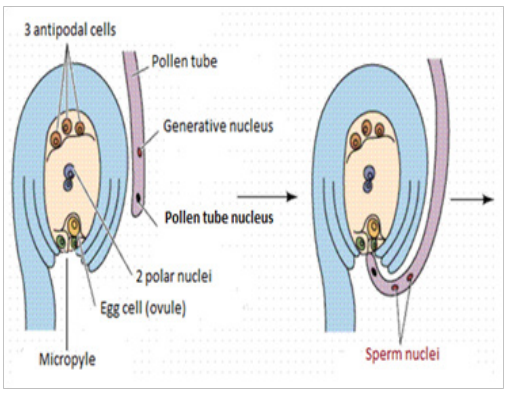
Double fertilization begins when a pollen grain adheres to the stigma of the carpel, the female reproductive structure of a flower.

The pollen grain then takes in moisture and begins to germinate, forming a pollen tube that extends down toward the ovary through the style.

The growth of the pollen tube is controlled by the pollen tube nucleus. In the pollen tube, the generative nucleus divides mitotically into two haploid nuclei which are the male gamete nuclei. These follow behind the tube nucleus as the pollen tube grows down the style towards the ovule. The tip of the pollen tube then enters the ovary and penetrates through the micropyle opening, releasing the two sperms in the megagametophyte or ovule.

The tube nucleus degenerates, leaving a clear passage for the entry of male nuclei. One nucleus fertilizes the egg cell to form a diploid zygote (2N), which will grow into a new plant embryo; the other fuses with polar nuclei to form a triploid nucleus (3N), which will grow into a food-rich tissue known as endosperm, which nourishes the seedling as it grows.





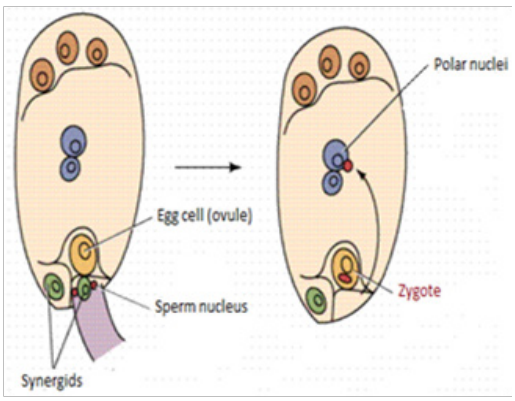


Figure 14.16: The process of double fertilization.

This process is described as double fertilisation and is typical of angiosperms. If there is more than one ovule in the ovary, each must be fertilised by separate pollen grain and hence the fruit will have many seeds genetically different from each other.

3. Events in a flower after fertilization

After fertilisation, the calyx, corolla, stamens and style may wither gradually and fall off, but in some flowers the calyx may persist. The ovule forms the seed, the two integuments of the ovule will form the seed coat, and the ovary will develop into fruit, with the ovary wall forming the pericarp (fruit wall). The diploid zygote undergoes cell division to form the embryo, the triploid primary endosperm nucleus develops into endosperm, a store used by the developing embryo. This persists in endospermic seeds of monocotyledons.

The micropyle persists as a small hole in the seed coat through which water is absorbed during germination.

Table 14.2: Floral parts and their fate after fertilization

Floral Part	Fate after fertilization
(a) Sepals, petals & stamens	All wither and drop off
(b) Ovary	Becomes the fruit
i) Ovary wall	¾® fruit wall
ii) Ovules	¾® seeds
iii) Integuments	¾® seed coat (testa)
iv) Fertilized egg	¾® embryo

Application activity 14.3.3

1. Are angiosperms typically wind or animal pollinated? How does this process occur?
2. What is meant by the term endosperm?
3. How are brightly coloured petals advantageous to the plant?
4. What do you understand by the term double fertilization?
5. What happens to the antipodal cells and synergids cells after fertilization?

14.3.4. Structures and types of fruits and seeds

Activity 14.3.4

Observe slides containing micrographs of different fruits and seeds. According to their characteristics:

1. Differentiate fruits.
2. Draw and show a structure of seed as seen on microscope

Below are some examples of fruits:

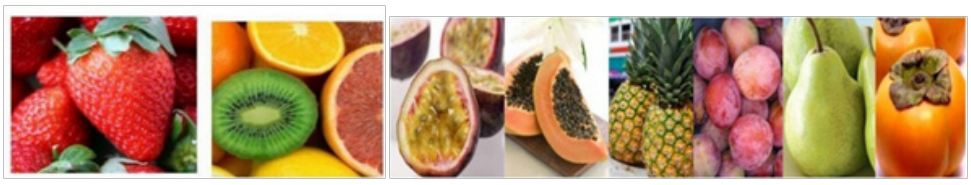


Figure 14.17: Different fruits

A fruit is a structure formed from the ovary of a flower, usually after the ovules have been fertilized. In nature, a fruit is normally produced only after fertilization of ovules has taken place, but in many plants, largely cultivated varieties such as seedless citrus fruits, grapes, bananas, and cucumbers, fruit matures without fertilization, a process known as parthenocarpy. Ovules within fertilized ovaries develop to produce seeds. In unfertilized varieties, seeds fail to develop, and the ovules remain with their original size.

A fruit consists of two main parts; pericarp (fruit wall) and the seed. The pericarp has three layers: epicarp or exocarp (outermost), mesocarpe (middle) and endocarp (inner).

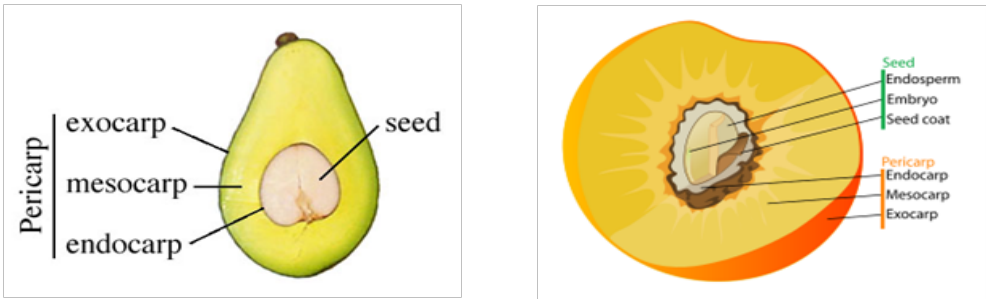


Figure 14.18: Structure of a fruit

The fruit can have a dry pericarp or fleshy pericarp. The fruits with fleshy pericarp include: berry and drupe. Drupe is a fleshy fruit with only one seed, E. g. avocado.

Figure 14.19: Structure of a drupe

Berry is a fleshy fruit having many seeds inside of it. E.g. tomatoes, orange, and pawpaw.

Figure 14.20: Structure of berry

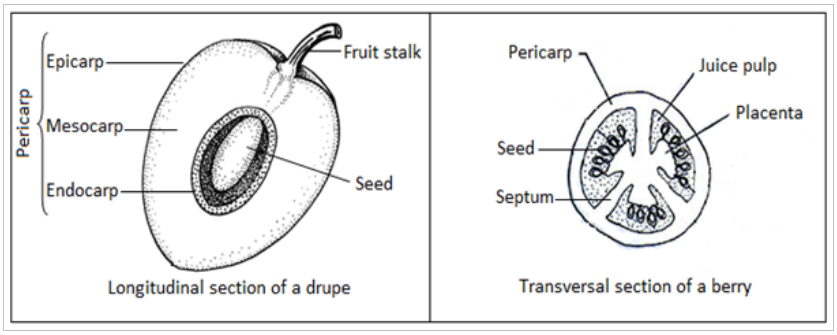


Figure 14.21: Structures of drupe and berry

The fruits with dry pericarp include indehiscent fruit or dehiscent fruit. Indehiscentfruits do not open. Seeds remain inside of the fruits. E.g. fruits of coconuts. Dehiscent fruits open and release seeds. These include: dehiscent fruits with one carpel, and those with many carpels. Dehiscent fruits with one carpel include; those which open along one side, e.g. follicle; and those which open along both sides, e.g. legume (beans).

Fruits of eucalyptus are exam of dehiscent fruits with many carpels.

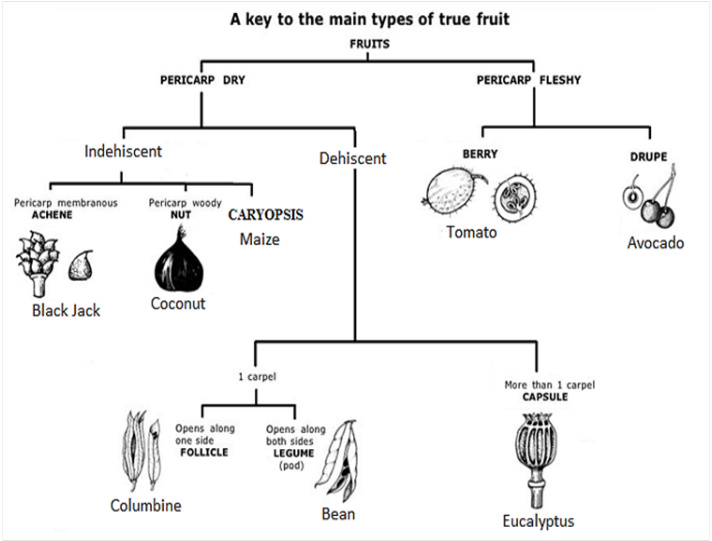


Figure 14.22: Types of fruits

The major function of a fruit is the protection of developing seeds. In many plants, the fruit also aids in seed distribution (dispersal).

Food value

Fruits are eaten raw or cooked, dried, canned, or preserved. Carbohydrates, including starches and sugars, constitute the principal nutritional material. Citrus fruits, tomatoes, and strawberries are primary sources of **vitamin C**, and most fruits contain considerable quantities of **vitamin A** and **vitamin B**. In general, fruits contain little protein or fat. Exceptions are avocados, nuts, and olives, which contain large quantities of fat, and grains and legumes, which contain considerable protein.

A seed is an embryonic plant enclosed in a protective outer covering. The formation of the seed is part of the process of reproduction in seed plants, the spermatophytes, including the gymnosperm and angiosperm plants. Seeds are the product of the ripened ovule, after fertilization by pollen and some growth within the mother plant. The embryo is developed from the zygote and the seed coat from the integuments of the ovule.

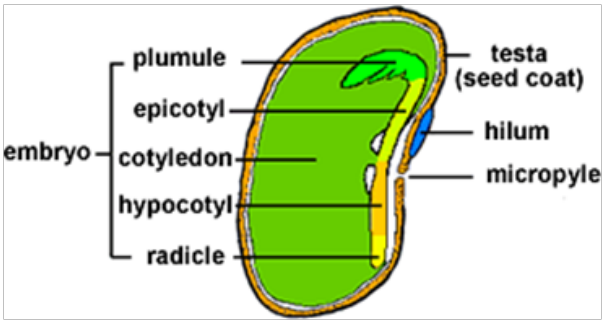


Figure 14.23: Structure of a seed

The main components of the embryo are:A seed is made up of a seed coat (testa), one or two cotyledons and an embryonic axis. The embryonic axis is made up of a plumule, an epicotyl, a hypocotyl and a radical. A seed which has one seed-leaf is described as monocotyledonous, and one which has two, as dicotyledonous. Maize is monocotyledonous seed while bean is a dicotyledonous seed.

- **The cotyledons**, the seed leaves, attached to the embryonic axis. There may be one (Monocotyledons), or two (Dicotyledons). The cotyledons are also the source of nutrients in the non-endospermic Dicotyledons, in this case they replace the endosperm, and are thick and leathery. In endospermic seeds the cotyledons are thin and papery.
- **The epicotyl**, the embryonic axis above the point of attachment of the cotyledon(s).

- **The plumule**, the tip of the epicotyl, and has a feathery appearance due to the presence of young leaf primordia at the apex, and will become the shoot upon germination.
- **The hypocotyl**, the embryonic axis below the point of attachment of the cotyledon(s), connecting the epicotyl and the radicle, being the stem-root transition zone.
- **The radicle**, the basal tip of the hypocotyl, grows into the primary root.

Monocotyledonous plants have two additional structures in the form of sheaths. The plumule is covered with a coleoptile that forms the first leaf while the radicle is covered with a coleorhiza that connects to the primary root and adventitiousroots form from the sides. Here the hypocotyl is a rudimentary axis between radicle and plumule.

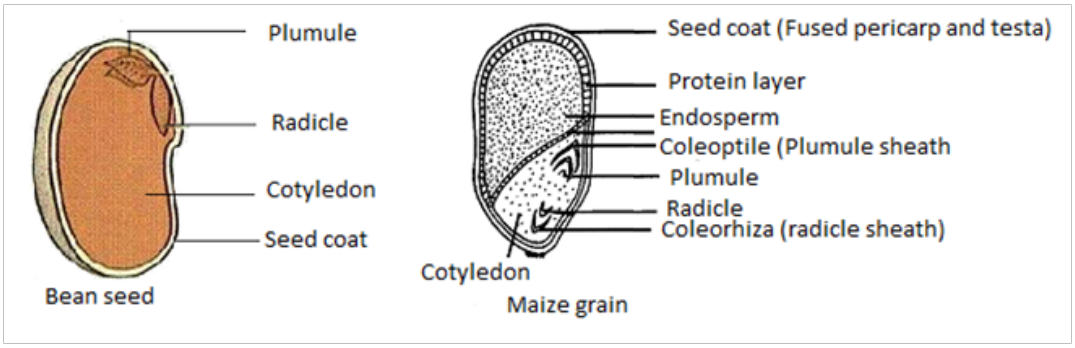


Figure 14.24: Structure of dicotyledonous and monocotyledonous seeds

Application activity 14.3.4

1. Describe the structure of a drupe
2. Differentiate between a drupe and a berry
3. What would happen to the fruit if ovules in the flower did not develop?
4. Compare the typical structure of seeds that are dispersed by animals to those dispersed by wind and water.

14.3.5. Fruits and seeds dispersal with their adaptations

Activity 14.3.5

A man visited a lake and he found around the lake some grown plants of maize, beans and other variety plants meanwhile nobody has cultivated any seed or fruit. Suggest ways of those variety of (plants) have been reached on that lake.

Dispersal of fruits and seeds is the scattering of fruits and seeds from their mother plants. They are four methods of seeds and fruits dispersal such as:

1. Dispersal by Wind
2. Dispersal by Water
3. Dispersal by Animals and
4. Mechanical Dispersal.

Seeds disperses by wind or water are typically **lightweight**, allowing them to be carried in air or to float on the surface of water. The wind carries also small seeds that have wing-like structure. Seeds dispersed by animals are typically contained in sweet, nutritious flesh fruits. They can be carried externally on their feet, fur, feathers, or beaks.

Those seeds with hooks or sticky substances rely on the chance that they will attach themselves to a passing animal. Other seeds are eaten by **animals and passed out in the faeces.**

With mechanical Dispersal: All dehiscent fruits scatter the seeds when they burst. This dehiscence is accompanied by the expression of great force in many fruits so that seeds are jerked a considerable distance away from the mother plant. Such fruits are called explosive fruits.

These seeds will germinate where the faeces will be deposited. The dispersal of seeds is important for the survival of the plant species because:

- It minimises overcrowding of plants growing around the parent plant that could then result in too much competition for nutrients and light;
- It allows the plant species to colonise new habitats which can offer suitable conditions.

Application activity 14.3.5

1. Why is it adaptive for some seeds to remain dormant before they germinate?
2. The seeds of a bishop pine germinate only after they have undergone a forest fire. Evaluate the significance of this structural adaptation.
3. Evaluate the importance of seed dispersal.

Skills lab 14

After studies, and completion of this unit 14, student-teachers will use the acquired knowledge to increase the crop productivity using different modes of vegetative and artificial propagation. They will also improve the quality of fruit plants by using forexample grafting method.

End unit assessment 14

PART A

A. Multiple choice questions (choose the best answers)

1. In cutting method of vegetative propagation, cuttings are mainly taken from
   a). Leaves of parent plant.
   b). Roots or stems of parent plant.
   c). Shoots of parent plant.
   d). Buds of parent plant.
2. Artificial methods of vegetative propagation includes
   a). Cloning
   b). Grafting
   c). Cuttings
   d). Both b and c
3. Example of plant in which vegetative propagation is occurred by leaves is called _____
   a). Cannabis
   b). Chrysanthemum
   c). Bryophyllum
   d). Brassica
4. Which of the following is NOT an advantage of asexual reproduction?
   a). Rapid reproduction.
   b). High genetic diversity.
   c). No need for a mate.
   d). Low resource investment in offspring.

B. PART B

1. Answer by true or false.
   a). Mosses have life cycles that depend on water for reproduction.
   b). In ferns the gametophyte depends on sporophyte.
   c). In mosses the sporophyte dominates over the gametophyte.
   d). Seeds that are dispersed by animals are not contained in a flesh-sweet tissue.
   e). During pollination, pollen grains move from stigma to anther.

2. Chose the letter that best answers the question or complete the statement.
   a). Which of the following is not part of a flower?
      i. Stamens
      ii. Petals
      iii. Carpels
      iv. Stem
      v. Sepals.
   b). Which flower structure that includes all the others listed below?
      i. Stigma
      ii. Carpel
      iii. Ovary
      iv. Style
      v. Ovule
   c). The thickened ovary wall of a plant may join with other parts of the flower stem to become the _____
      i. Cotyledon
      ii. Fruit
      iii. Endosperm
      iv. Seed
   d). In angiosperms the structures that produce the male gametophyte are called the _____
      i. Pollen tubes
      ii. Stigma
      iii. Anthers
      iv. Sepals

e). The small, multicellular structures by which liverworts reproduce asexually are _____

i. Archegonia
ii. Gemmae
iii. Protonema
iv. Rhizoids

f). The small, multicellular structures by which liverworts reproduce asexually are _____

i. Archegonia
ii. Gemmae
iii. Protonema
iv. Rhizoids

g). In angiosperms, the mature seed is surrounded by a _____

i. Flower
ii. Fruit
iii. Cotyledon
iv. Cone

h). The leaves of ferns are called _____

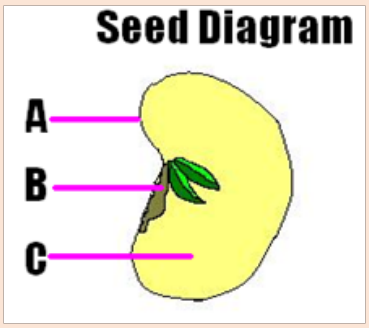
i. Spores
ii. Fronds
iii. Sori
iv. Rhizomes

i). The most recognizable stage of a moss is the _____

i. Gametophyte
ii. Archegonium
iii. Protonema
iv. Sporophyte

3. Which are more likely to be dispersed by animals- the seeds of angiosperms or the spores of a fern? Explain your reasoning.
4. Pollination is a process that occurs only in seed plants. What process in seedless plants is analogous to pollination?
5. What is the dominant stage of the ferns life cycle? Explain the relationship of the fern gametophyte and sporophyte.

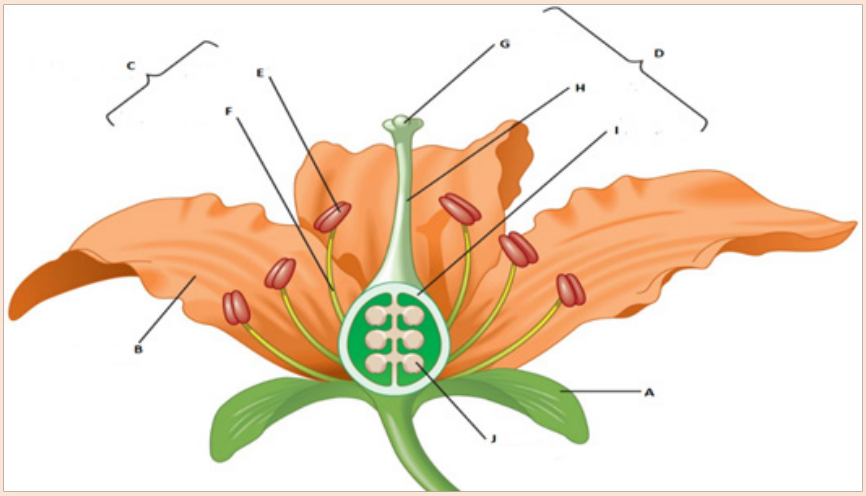
6. How is water essential in the life cycle of a bryophyte?
7. What characteristic of bryophytes is responsible for their small size? Explain.
8. Briefly explain why a seed may remain dormant even when the environmental conditions are favorable for germination.
9. Describe the relationship between the gametophyte and sporophyte in mosses.
10. During the lifecycle of a moss, what environmental conditions are necessary for fertilization to occur?
11. Describe the dominant stage in the lifecycle of a fern.
12. Propose a hypothesis to explain why angiosperms have become the dominant type of plant on the earth.
13. Moss plants are small. Ferns can grow as tall as small tree. Explain why this is so. How does your answer illustrate a major characteristic of the plant kingdom?
14. Study the structure of the seed below.



a). Name the parts labelled by: A, B and C

b). What is the importance of the part C for a growing seedling?

15. Many flowers have bright patterns of coloration that directly surround the reproductive structures. Evaluate the importance of those bright-coloured patterns to plants.
16. What is the function of endosperm?
17. Some plants form flowers that produce stamens but no carpels. Could fruit form on one of these flowers? Explain your answer.
18. Distinguish between pollination and fertilization.
19. Give names of letters from A to J, and explain the function of the parts represented by: B, G, and E.



20. Explain why the relationship between bees and flowers is described as mutually beneficial.
21. What is the main advantage of cross-pollination?
22. Why are the stamens of wind-pollinated plants and insect-pollinated plants different?
23. Differentiate wind-pollinated flowers from insect-pollinated flowers.
24. Give one example of a plant that uses each of the following dispersal mechanism:

    a). An explosive device which works by being inflated with water.

    b). A winged seed lifted by air currents.

    c). A buoyant seed carried by sea currents.

    d). A gluey substance which sticks the seed to an animal.

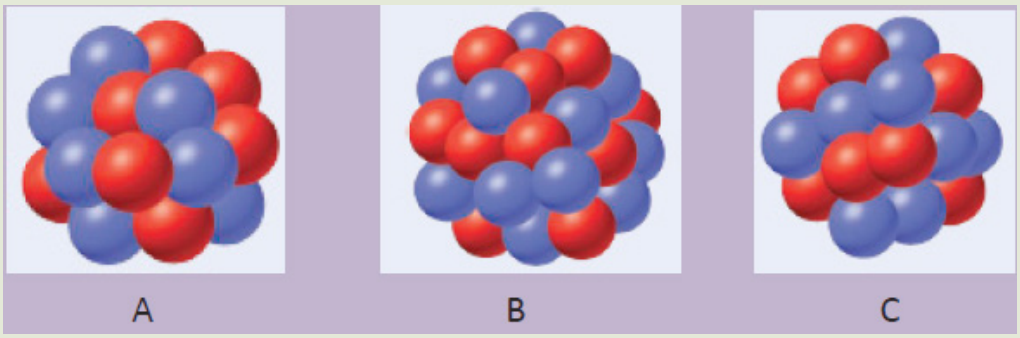
UNIT 15:

STRUCTURE OF AN ATOM AND MASS SPECTRUM

Key Unit competence: Interpret simple mass spectra and use them to calculate R.A.M. of different elements.

Introductory Activity

The presentations A, B, and C below show three atomic nuclei of different elements. Study the presentations carefully and answer the questions below.



1. How many blue and red spheres do you see in each of the diagrams above?
2. What do the three diagrams A, B, and C have in common?
3. Based on your knowledge concerning atomic structure, what do you think that

   d) the blue spheres represent?

   e) the red spheres represent? Provide explanations.

4. Using the information obtained in question (3) write the atomic symbol for each of the diagrams.
5. Are there some other particle(s) missing from the above diagrams? If yes name the particle(s).
6. What could you obtain if the atom is broken down?

15.1. The constituents of an atom, their properties and the outline of their discovery

Activity 15.1

1. Regardless of some exceptions, all atoms are composed of the same components. True or False? If this statement is true why do different atoms have different chemical properties?
2. The contributions of Joseph John Thomson and Ernest Rutherford led the way to today's understanding of the structure of the atom. Use available resources to figure out:
   c) the contribution of each of them.
   d) the modern view of the structure of the atom?
   e) the role of each particle in an atom.

Atoms are the basic units of elements and compounds. In ordinary chemical reactions, atoms retain their identity. An *atom* is the smallest identifiable unit of an *element*. There are about 91 different naturally occurring elements. In addition, scientists have succeeded in making over 20 synthetic elements (elements not found in nature but produced in Laboratories of Research Centres).

An element is defined as a substance that cannot be broken down by ordinary chemical methods in simpler substances. Some examples of elements include hydrogen (H), helium (He), potassium (K), carbon (C), and mercury (Hg). In an element, all atoms have the same number of protons or electrons although the number of neutrons can vary. A substance made of only one type of atom is also called element or elemental substance, for example: hydrogen (H_2), chlorine (Cl_2), sodium (Na). Elements are the basic building blocks of more complex matter.

A compound is a matter or substance formed by the combination of two or more different elements in fixed ratios. Consider, Hydrogen peroxide (H_2O_2) is a compound composed of two elements, hydrogen and oxygen, in a fixed ratio (2:2).

1. Discovery of the atom constituents

The oldest description of matter in science was advanced by the Greek philosopher Democritus in 400BC.

He suggested that matter can be divided into small particles up to an ultimate particle that cannot any more be divided, and called that particle atom. Atoms came from the Greek word *atomos* meaning indivisible.

The work of Dalton and other scientists such as Avogadro, etc., contributed more so that chemistry was beginning to be understood. They proposed new concept of atom, and from that moment scientists started to think about the nature of the atom.

In 1808 Dalton published *A New System of Chemical Philosophy,* in which he presented his theory of atoms:

Dalton's Atomic Theory

a). Each element is made up of tiny particles called atoms.

b). The atoms of a given element are identical; the atoms of different elements are different in some fundamental way(s).

c). Chemical compounds are formed when atoms of different elements combine with each other. A given compound always has the same relative numbers and types of atoms.

d). Chemical reactions involve reorganization of the atoms—changes in the way they are bound together. The atoms themselves are not changed in a chemical reaction.

Figure 15.1: John Dalton's Atomic Model

2. Discovery of Electrons and Thomson's Atomic Model

In 1897 J. J. Thomson (1856–1940) and other scientists conducted several experiments, and found that atoms are divisible. They conducted experiments with gas discharge tubes. A gas discharge tube is shown in Figure 15.2.

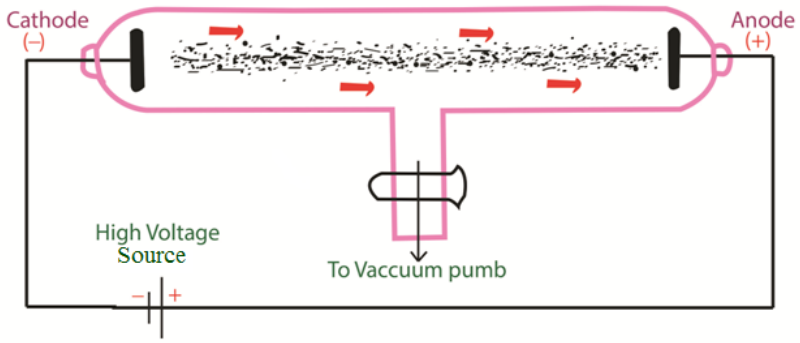


Figure 15.2: Gas discharge tube showing cathode rays originating from the cathode

The gas discharge tube is an evacuated glass tube and has two electrodes, a cathode (negative electrode) and an anode (positive electrode). The electrodes are connected to a high voltage source. Inside the tube, an electric discharge occurs between the electrodes.

The discharge or 'rays' originate from the cathode and move toward the anode, and hence are called cathode rays. Using luminescent techniques, the cathode rays are made visible and it was found that these rays are deflected away from negatively charged plates. The scientist J. J. Thomson concluded that the cathode rays consist of negatively charged particles, and he called them electrons.

Thomson postulated that an atom consisted of a diffuse cloud of positive charge with the negative electrons embedded randomly in it. This model, shown in Figure 15.3, is often called the plum pudding model because the electrons are like raisins dispersed in a pudding (the positive charge cloud), as in plum pudding.

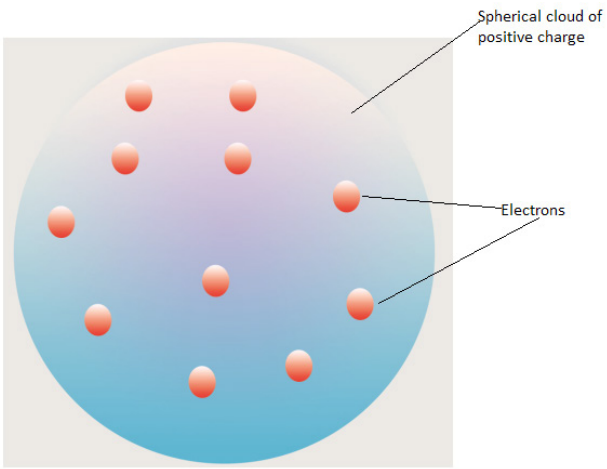


Figure 15.3.The plum pudding model of the atom.

In 1909 Robert Millikan (1868–1953) conducted the famous charged oil drop experiment and came to several conclusions: He found the magnitude of the charge of an electron equal to $-1.602\text{x}10^{-19}C$. From the charge-to-mass ratio(e/m) determined by Thomson, the mass of an electron was also calculated.

$$\frac{\textbf{Charge}}{\textbf{Mass}} = \textbf{-1.76x10}^{\textbf{8}}\ \textbf{Coulomb/gram}$$

$$\text{Mass} = \frac{\textbf{-1.6 x 10}^{\textbf{-19}}}{\textbf{-1.76 x 10}^{\textbf{8}}} = 9.11\ \text{x}\ \textbf{10}^{\textbf{-29}}\text{g} = 9.11\ \text{x}\ \textbf{10}^{\textbf{-31}}\text{kg}$$

3. Discovery of Protons and Rutherford's Atomic Model

In 1886 Eugene Goldstein (1850–1930) observed that a cathode-ray tube also generates a stream of positively charged particles that move towards the cathode. These were called canal rays because they were observed occasionally to pass through a channel, or "canal," drilled in the negative electrode (Figure 15.4). These *positive rays*, or *positive ions*, are created when the gaseous atoms in the tube lose electrons. Positive ions are formed by the process

$$\text{Atom} \longrightarrow \text{cation} + e^- \text{ (energy is absorbed)}$$

Different elements give positive ions with different e/m ratios. The regularity of the e/m values for different ions led to the idea that there is a subatomic particle with one unit of positive charge, called the proton. The proton is a fundamental particle with a charge equal in magnitude but opposite in sign to the charge on the electron. Its mass is almost 1836 times that of the electron.

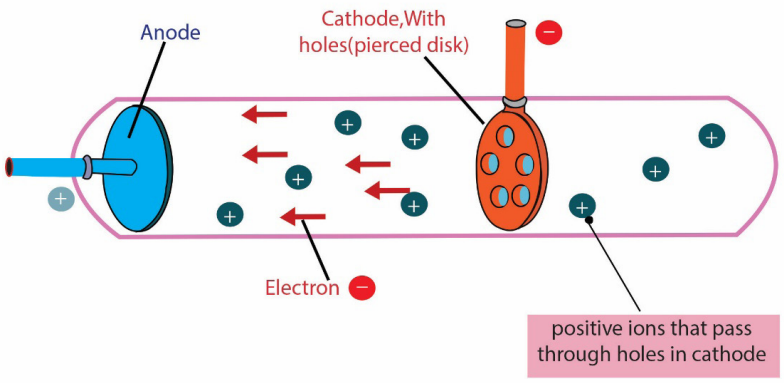


Figure 15.4: A cathode-ray tube with a different design and with a perforated cathode

The proton was observed by Ernest Rutherford and James Chadwick in 1919 as a particle that is emitted by bombardment of certain atoms with α-particles.

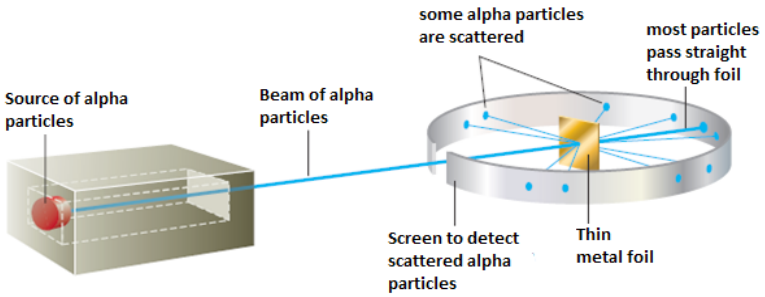


*Figure 15.5: Rutherford's experiment on **α**-particle bombardment of metal foil*

Rutherford reasoned that if Thomson's model were accurate, the massive α-particles should crash through the thin foil like cannonballs through gauze, as shown in Figure 15.6(a). He expected α-particles to travel through the foil with, at the most, very minor deflections in their paths. The results of the experiment were very different from those Rutherford anticipated. Although most of the α- particles passed straight through, many of the particles were deflected at large angles, as shown in Figure 15.6(b), and some were reflected, never hitting the detector. This outcome was a great surprise to Rutherford. Rutherford knew from these results that the plum pudding model for the atom could not be correct. The large deflections of the α-particles could be caused only by a centre of concentrated positive charge that contains most of the atom's mass, as illustrated in Figure 15.6(b). Most of the α-particles pass directly through the foil because the atom is mostly empty space. The deflected α-particles are those that had a "close encounter" with the massive positive centre of the atom, and the few reflected α-particles are those that made a "direct hit" on the much more massive positive centre.

In Rutherford's mind these results could be explained only in terms of a nuclear atom—an atom with a dense centre of positive charge (the nucleus) with electrons moving around the nucleus at a distance that is large relative to the nuclear radius.

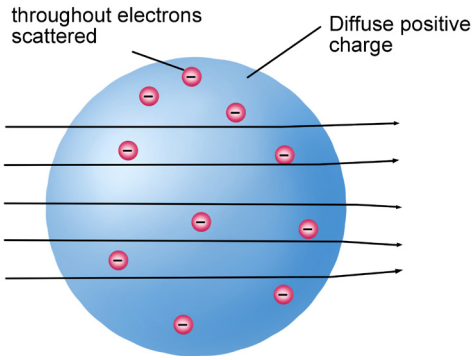


Figure 15.6 (a): The expected results of the metal foil experiment of Thomson's model were correct

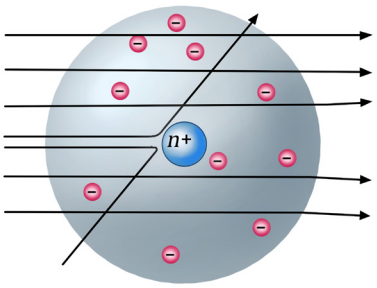


Figure 15.6 (b) : Actual results

4. Discovery of Neutrons

In spite of the success of Rutherford and his co-workers in explaining atomic structure, one major problem remained unsolved.

If the hydrogen contains one proton and the helium atom contains two protons, the relative atomic mass of helium should be twice that of hydrogen. However, the relative atomic mass of helium is four and not two.

This question was answered by the discovery of James Chadwick, English physicist who showed the origin of the extra mass of helium by bombarding a beryllium foil with alpha particles. (See figure 15.7).

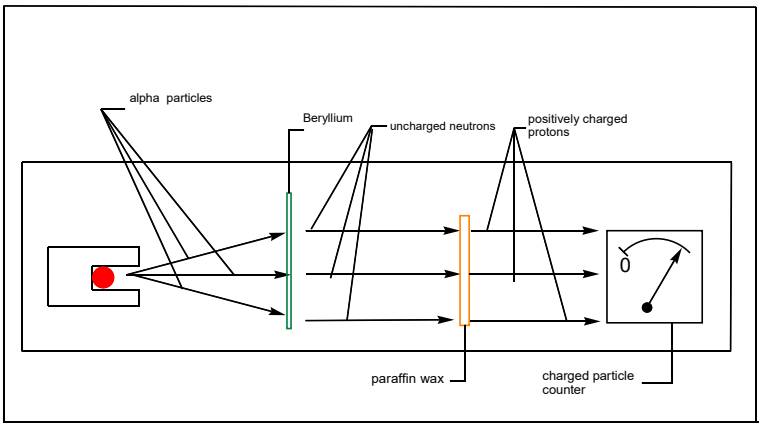


Figure 15.7: Chadwick's experiment

In the presence of beryllium, the alpha particles are not detected; but they displace uncharged particles from the nuclei of beryllium atoms. These uncharged particles cannot be detected by a charged counter of particles.

However, those uncharged particles can displace positively charged particles from another substance. They were called neutrons. The mass of the neutron is slightly greater than that of proton.

Figure 15.8 shows the location of the elementary particles (protons, neutrons, and electrons) in an atom. There are other subatomic particles, but the electron, the proton, and the neutron are the three fundamental components of the atom that are important in chemistry.

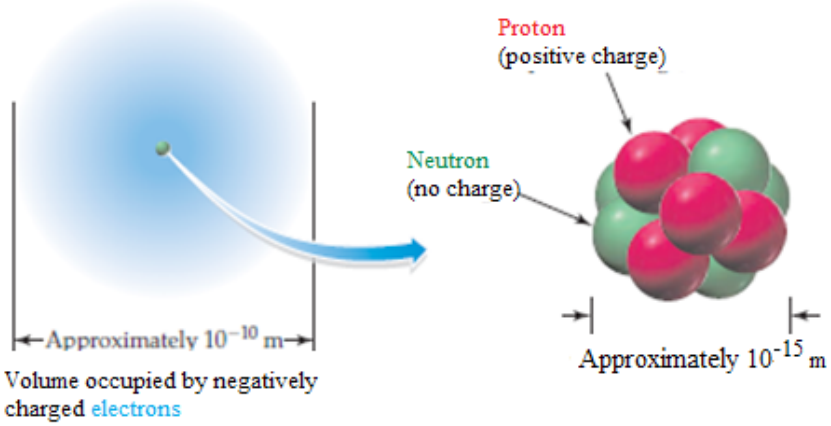


Figure 15.8: The protons and neutrons in an atom are contained in the tiny volume of the nucleus. Electrons are distributed within the sphere surrounding the nucleus.

As a result of the experiments described above, it is found that atoms consist of very small, very dense nuclei surrounded by clouds of electrons at relatively great distances from the nuclei. All nuclei contain protons; nuclei of all atoms except the common form of hydrogen also contain neutrons.

5. Properties of sub-atomic particles

Protons and neutrons are collectively known as nucleons. Both protons and neutrons have a mass almost equal to that of hydrogen atom. The neutron has no charge whereas the proton carries one positive charge. The electron with one negative charge occupies the space outside the nucleus.

The following table summarizes the relative masses, the relative charges and the position within the atom of these sub-atomic particles.

Table 15.1: Subatomic particles and some of their properties

Particle	Absolute charge(Coulomb)	Relative charge	Mass (kg)	Relative masse
Neutron(n)	0	0	1.675×10^{-27}	1.0087 amu*
Proton (p or P^+)	$+1.6\times10^{-19}$	+1	1.673×10^{-27}	1.0073 amu
Electron (e-)	-1.6×10^{-19}	-1	9.11×10^{-31}	0.00054858 amu

(*amu: atomic mass unity, 1 amu =1.67×10^{-27}kg)

Application activity 15.1

1. In an experiment, it was found that the total charge on an oil drop was 5.93×10^{-18} C. How many negative charges does the drop contain?
2. All atoms of the elements contain three fundamental particles. True or false? Give an example to support your answer.
3. Compare the atom constituents
   d) in terms of their relative masses
   e) in terms of their relative charges
4. Using the periodic table as a guide, specify the number of protons and electrons in a neutral atom of each of these elements.
   a). Carbon (C) b). Calcium (Ca) c).Chlorine (Cl) d). Chromium (Cr)

15.2. Concept of atomic number, mass number, isotopic mass and relative atomic mass

Activity 15.2

Atoms of the same element may exhibit some physical properties like mass, density, velocity. Is the statement true or false? Use available textbooks and suggest an appropriate explanation?

The atomic number or proton number, Z, denotes the number of protons in an atom's nucleus. It corresponds to the order of the element in the periodic table.

The mass number or nucleon number, A, denotes the total number of protons and neutrons in an atom.

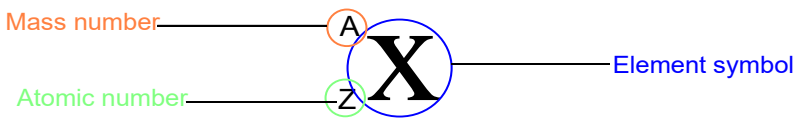
Mass number = number of protons + number of neutrons

= atomic number + neutron number

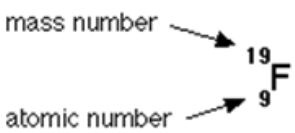
The number of neutrons can be obtained by subtracting the atomic number from the mass number.

Chemists use the following shorthand to represent an atom. The mass number is shown as a superscript (top number) and the atomic number is shown as a subscript (bottom number) beside the symbol of the element.

By convention, the atomic number is usually written to the left subscript of the elemental notation, and the mass number to the left superscript of the elemental notation as represented by the example below, where X represents any element symbol.



Example:



Some atoms of the same element have the same atomic number, but different mass numbers. This means a different number of neutrons. Such atoms are called isotopes of the element. They are nuclides of the same element.

Example

	Protons	Neutrons	mass number
carbon-12	6	6	12
carbon-13	6	7	13
carbon-14	6	8	14

Isotopes of an element have the same chemical properties because they have the same number of electrons.

When elements react, it is the electrons that are involved in the reactions.

This means that the isotopes of an element cannot be differentiated by chemical reactions.

Because isotopes of an element have different numbers of neutrons, they have different masses, and isotopes have slightly different physical properties.

The mass of a single isotope is its isotopic mass. The relative isotopic mass of an isotope is the relative mass of that isotope compared with the isotope $^{12}_{6}\mathbf{C}$, which is given a mass of 12.00 units (12.00 atomic mass units). That is, relative isotopic mass relates to the relative atomic mass scale on which one isotope of the carbon element, carbon-12 is taken as the reference standard for atomic masses and is given a relative mass of 12 units, precisely 12 atomic mass units (a.m.u).

$$\textbf{Relative isotopic mass} = \frac{\text{mass of 1 isotope of the element}}{\frac{1}{12}\text{ x mass of 1 atom of } {}^{12}_{6}\text{C}}$$

Atomic mass unit (a.m.u) is a unit of mass used to express “relative atomic masses”. It is 1/12 of the mass of the isotope of carbon-12 and is equal to 1.66054×10^{-27}kg.

The relative isotopic masses of all other atoms are obtained by comparison with the mass of a carbon-12 atom.

On that scale, the relative atomic mass of a proton and that of a neutron are both very close to one unit (1.0074 and 1.0089 units respectively). Since the relative mass of an electron is negligible (0.0005units), it follows that all isotopic masses will be close to whole numbers.

However relative atomic masses of elements are not close to whole numbers because natural occurring elements are often mixtures of isotopes.

Application activity 15.2

1. How do you call the members of each of the following pairs? Explain.
   a). $^{35}_{17}Cl$ and $^{37}_{17}Cl$
   b). $^{63}_{29}Cu$ and $^{65}_{29}Cu$

$^{35}_{17}Cl$

2. Write, using the periodic table, the correct symbols to identify an atom that contains
   a). 4 protons, 4 electrons, and 5 neutrons;
   b). 23 protons, 23 electrons, and 28 neutrons;
   c). 54 protons, 54 electrons, and 70 neutrons; and
   d). 31 protons, 31 electrons, and 38 neutrons.
3. Use the list of the words given below to fill in the blank spaces. Each word will be used once.

   Atomic number, Mass number, protons, Electrons, Isotopes, neutron

   a). The atomic number tells you how many _____ and _____ are in an atom.
   b). _____ is the number written as subscript on the left of the atomic symbol.
   c). The total number of protons and neutrons in an atom is called the _____.
   d). Atoms with the same number of protons but different number of neutrons are called _____.
   e). The subatomic particle that has no charge is called a _____.

15.3. Calculations of the relative atomic masses of elements

Activity 15.3

Using textbooks and internet connection, explain the concept of relative mass and attempt to solve the problems below.

1. Argon has three naturally occurring isotopes: argon-36, argon-38, and argon-40. Based on argon's reported relative atomic mass from the periodic table, which isotope do you think is the most abundant in nature? Explain.

2. Calculate the average atomic mass of an element with two naturally occurring isotopes: ^{85}X (72.15%, 84.9118 amu) and ^{87}X (27.85%, 86.9092 amu). Identify this element?
3. Boron has two naturally occurring isotopes. Find the percent abundances of ^{10}B and ^{11}B given the isotopic mass of ^{10}B = 10.0129 amu and the isotopic mass of ^{11}B = 11.0093 amu.

The **Relative atomic mass,** symbolized as R.A.M (Ar), is defined as the average of the relative isotopic masses of the different isotopes weighted in the proportions in which they naturally occur.

Thus, the different isotopic masses of the same elements and the percentage abundance of each isotope of an element must be known in order to accurately calculate the relative atomic mass of an element.

Notice: Remember that mass number is not the same as the relative atomic mass or isotopic mass! The mass number is the total number of protons and neutrons while relative atomic mass is the average of the isotopic masses.

Let $\mathbf{A_1, A_2, A_3,\ldots, A_n}$ be an abundance of n isotopes of the same chemical element with atomic mass $\mathbf{M_1, M_2, M_3,\ldots, M_n}$ respectively, the relative atomic mass (R.A.M) is given by the following equation:

$$\mathbf{RAM} = \frac{\mathbf{A_1M_1 + A_2M_2 + A_3M_3 + \cdots + A_nM_n}}{\mathbf{A_1 + A_2 + A_3 + \cdots + A_n}}$$

Example 1

Oxygen contains three isotopes ^{16}O, ^{17}O, and ^{18}O. Their respective relative abundances are 99.76%, 0.04%, and 0.20%. Calculate the relative atomic mass of oxygen.

Solution:

Relative isotopic mass of ^{16}O is 16 and its relative abundance is 99.76%;

Relative isotopic mass of ^{17}O is 17, abundance 0.04%;

Relative isotopic mass of ^{18}O is 18, abundance 0.20%.

$$\text{Relative atomic mass of oxygen} = \frac{(99.76 \times 16) + (0.04 \times 17) + (0.20 \times 18)}{99.76 + 0.04 + 0.20} = 16.0044$$

By applying the same formula, the relative abundance of the isotopes may be calculated knowing the relative atomic mass of the element and the atomic masses of the respective isotopes.

Example 2

Chlorine contains two isotopes ^{35}Cl and ^{37}Cl, what is the relative abundance of each isotope in a sample of chlorine if its relative atomic mass is 35.5?

Solution:

$$35.5 = \left(\frac{A \times 35}{100}\right) + \left(\frac{100 - A}{100}\right) \times 37$$

Note that if the abundance of the isotope of atomic mass 35 is A%, the abundance of the isotope of mass 37 will be (100 – A) %.

0.35A + (100 – A) x 0.37 = 35.5

0.35A– 0.37A =35.5 – 37

– 0.02A = – 1.5

A= 1.5/0.02 = 75

Therefore, the abundance of the isotope of relative atomic mass 35.5 is 75% while that for the isotope ^{17}Cl is 100 – 75 = 25%.

Application activity 15.3

1. Three isotopes of magnesium occur in nature. Their abundances and masses, determined by mass spectrometry, are listed in the following table. Use this information to calculate the relative atomic mass of magnesium.

Isotope	% abundance	Mass (amu)
^{24}Mg	78.99	23.98504
^{25}Mg	10.00	24.98584
^{26}Mg	11.01	25.98259

2. The atomic weight of gallium is 69.72 amu. The masses of the naturally occurring Isotopes are 68.9257 amu for ^{69}Ga and 70.9249 amu for ^{71}Ga. Calculate the percent abundance of each isotope.
3. Natural occurring lead contains 1.55% lead-204, 23.6% lead-206, 22.6% lead-205 and 52.3% lead -208. Calculate the relative atomic mass of lead.
4. The table below shows the mass number and the percentage abundances of the isotopes of an element X.

Mass number	54	56	57	58
Percentage abundance	5.84	91.68	2.17	0.31

e) Calculate the relative atomic mass of X

f) Identify the element X.

15.4. Mass spectrometer

15.4.1. Description and functioning of its components

Activity 15.4.1

1. Recall the concept of relative atomic mass of an element.
2. Do research and identify the name of an instrument used to estimate the mass and proportions of the isotopes that make up a given element. How does it work?

The mass spectrometer is an instrument that separates positive gaseous atoms and molecules according to their mass-charge ratio and that records the resulting mass spectrum.

In the mass spectrometer, atoms and molecules are converted into ions. The ions are separated as a result of the deflection which occurs in a magnetic and electric field.

The basic components of a mass spectrometer are: vaporisation chamber (to produce gaseous atoms or molecules), ionization chamber (to produce positive ions), accelerating chamber (to accelerate the positive ions to a high and constant velocity), magnetic field (to separate positive ions of different m/z ratios), detector (to detect the number and m/z ratio of the positive ions) and the recorder (to plot the mass spectrum of the sample).

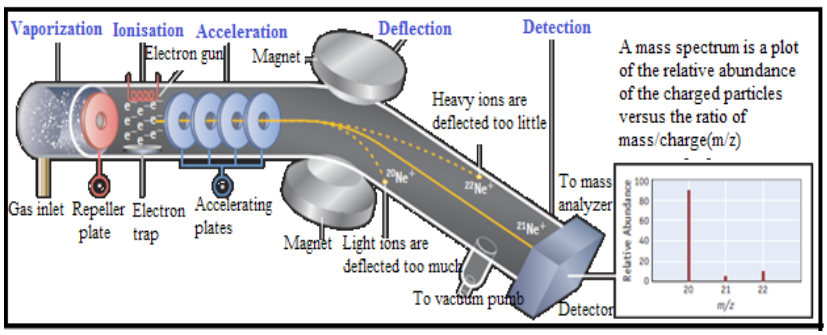


Figure15.9: Diagram of Mass spectrometer and its functioning
(Source: Kotz, et al., (2012). ***Chemistry & Chemical Reactivity,*** *Eighth Edition)*

A mass spectrometer works in five main stages, namely vaporization, ionization, acceleration, deflection, and detection to produce the mass spectrum.

Stage 1: Vaporization

At the beginning, the test sample is heated until it becomes vapour and is introduced as a vapour into the ionization chamber. When a sample is a solid with low vapour pressure, it can directly be introduced into the ionization chamber.

Stage 2: Ionisation

The vaporized sample passes into the ionization chamber (with a positive voltage of about 10,000 volts). The electrically heated metal coil gives off electrons which are attracted to the electron trap which is a positively charged plate.

The particles in the sample (atoms or molecules) are therefore bombarded with a stream of electrons (electrons gun), and some of the collisions are energetic enough to knock one or more electrons out of the sample particles to make positive ions. Mass spectrometers always work with positive ions.

Most of the positive ions formed will carry a charge of +1 because it is much more difficult to remove further electrons from an already positive ion.

Most of the sample molecules are not ionized at all but are continuously drawn off by vacuum pumps which are connected to the ionization chamber. Some of the molecules are converted to negative ions through the absorption of electrons.

Stage 3: Acceleration

The positive ions are accelerated by an electric field so that they move rapidly through the machine at high and constant velocity.

Stage 4: Deflection

The ions are then *deflected by a magnetic field* according to their masses and charges ratio. Different ions are deflected by the magnetic field at different extents. The extent to which the beam of ions is deflected depends on four factors:

1. *The magnitude of the accelerating voltage (electric field strength).* Higher voltages result in beams of more rapidly moving particles to be deflected less than the beams of the more slowly moving particles produced by lower voltages.

2. *Magnetic field strength.* Stronger fields deflect a given beam more than weaker fields.
3. *Masses of the particles.* Because of their inertia, heavier particles are deflected less than lighter particles that carry the same charge.
4. *Charges on the particles.* Particles with higher charges interact more strongly with magnetic fields and are thus deflected more than particles of equal mass with smaller charges

The two last factors (mass of the ion and charge on the ion) are combined into the *mass/charge ratio.* Mass/charge ratio is given the symbol m/z (or sometimes m/e).

For example, if an ion had a mass of 28 and a charge of 1+, its mass/charge ratio would be 28. An ion with a mass of 56 and a charge of 2+ would also have a mass/charge ratio of 28.

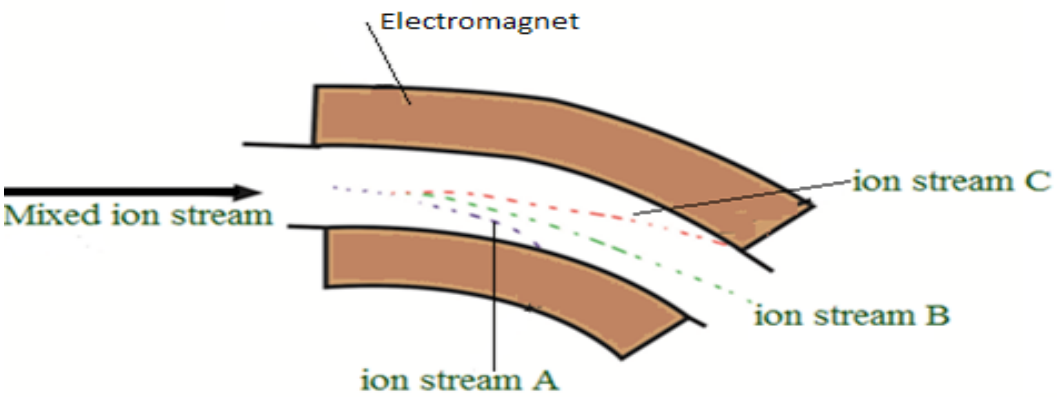


Figure 15.10: Magnetic field to separate positive ions of different m/z ratios

In the figure 15.10 above, ion stream A is most deflected: it will contain ions with the smallest mass/charge ratio. Ion stream C is the least deflected: it contains ions with the greatest mass/charge ratio. Assuming 1+ ions, stream A has the lightest ions, stream B the next lightest and stream C the heaviest. Lighter ions are going to be more deflected than heavy ones.

Stage 5: Detection

The beam of ions passing through the machine is detected electrically. As they pass out of the magnetic field, ions are *detected by an ion detector,* which records the position of the ions on the screen and the number of ions that hit the screen at each position. These two pieces of information are used to produce a mass spectrum for the sample.

A flow of electrons in the wire is detected as an electric current, which can be amplified and recorded. The more ions arriving, the greater the current.

Detecting the other ions

How might the other ions be detected (those in streams A and C which have been lost in the machine)?

Remember that stream A was most deflected. To bring them on to the detector, you would need to deflect them less by using a smaller magnetic field.

To bring those with a larger m/z value (the heavier ions if the charge is +1) to the detector you would have to deflect them more by using a larger magnetic field.

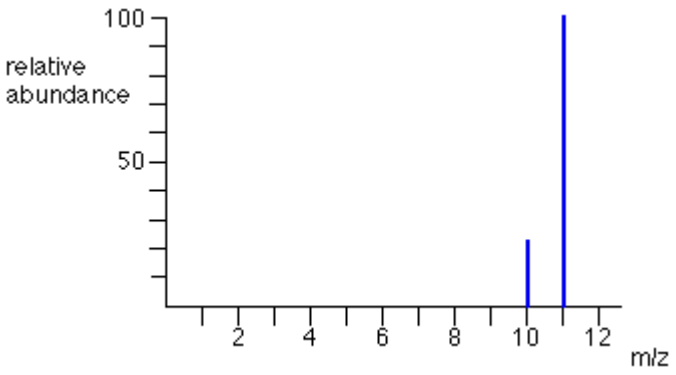
If you vary the magnetic field, you can bring each ion stream in turn on the detector to produce a current which is proportional to the number of ions arriving. The mass of each ion being detected is related to the size of the magnetic field used to bring it on to the detector. The machine can be calibrated to record current (which is a measure of the number of ions) against m/z directly. The mass is measured on the ^{12}C scale.

Note: The ^{12}C scale is a scale on which the ^{12}C isotope weighs exactly 12 units.

Recorder

The detector of a typical instrument consists of a counter, which produces a current that is proportional to the number of ions, which strike it. Using of electron multiplier circuits, this current can be measured so accurately that the current caused by just one ion striking the detector can be measured. The signal from the detector is fed to a recorder, which produces the mass spectrum. In modern instruments, the output of the detector is fed through an interface to a computer. The computer can store the data, provide the output in both tabular and graphic forms, and compare the data to standard spectra, which are contained in spectra libraries also stored in the computer.

The following figure is an example of a mass spectrum of unknown element that has 2 isotopes.



Application activity 15.4.1

1. Use the list of the words given below to fill in the blank spaces. Each word will be used once.

   Vaporization chamber, mass spectrum, velocity, ionization, deflection, detector, acceleration

   A sample of the element is placed in the _____ chamber where it is converted into gaseous atoms. The gaseous atoms are ionized by bombardment of high energy electrons emitted by a hot cathode to become positive ions (in practice, the voltage in the _____ chamber is set in such a way that only one electron is removed from each atom). The positive ions (with different masses) are then going faster to a high and constant _____by two negatively charged plates: the process is called_____. The positive ions are then deviated by the magnet field. This process is called _____ (ions with smaller mass will be deflected more than the heavier ones). These ions are then detected by the ion _____. The information is fed into a computer which prints out the_____ of the element.

2. The correct order for the basic features of a mass spectrometer is...

   a) acceleration, deflection, detection, ionization

   b) ionisation, acceleration, deflection, detection

   c) acceleration, ionisation, deflection, detection

   d) acceleration, deflection, ionisation, detection

3. Which one of the following statements about ionisation in a mass spectrometer is incorrect?

   a) gaseous atoms are ionised by bombarding them with high energy electrons

b) atoms are ionised so they can be accelerated

c) atoms are ionised so they can be deflected

d) it doesn't matter how much energy you use to ionise the atoms

4. The path of ions after deflection depends on...

   a) only the mass of the ion

   b) only the charge on the ion

   c) both the charge and the mass of the ion

   d) neither the charge nor the mass of the ion

5. Which of the following species will be deflected to the greatest extent?

   a) $^{37}Na^{+}$

   b) $^{35}Na^{+}$

   c) ^{37}Na

   d) $^{35}Na^{2+}$

6. Which of the following separates the ions according to their mass-to-charge?

   a) Ion source

   b) Detector

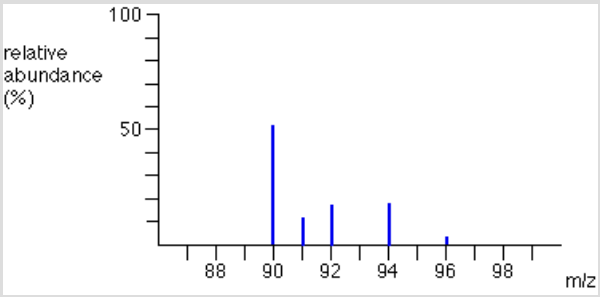
   c) Magnetic sector

   d) Electric sector

15.4.2. Interpretation of mass spectra and uses of the mass spectrometer

Activity 15.4.2

The mass spectrum of zirconium looks like this:



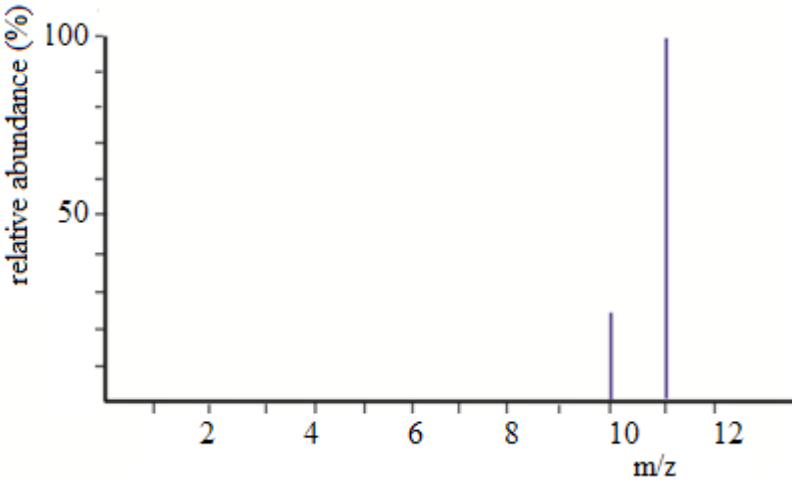
e) What does m/z mean?

f) Explain as fully as possible what the mass spectrum shows about zirconium. (You are not required to read actual values from the relative abundance axis.)

g) The spectrum shows lines for 1+ ions. If there were also peaks for 2+ ions, where would you expect to find them, and what would you predict about their heights relative to the 1+ peaks?

The mass spectrum of an element shows how you can find out the masses and relative abundances of the various isotopes of the element and use that information to calculate the relative atomic mass of the element.

Example 1: The mass spectrum of boron



The mass spectrum of boron may be used to know the number of boron isotopes and their relative abundances.

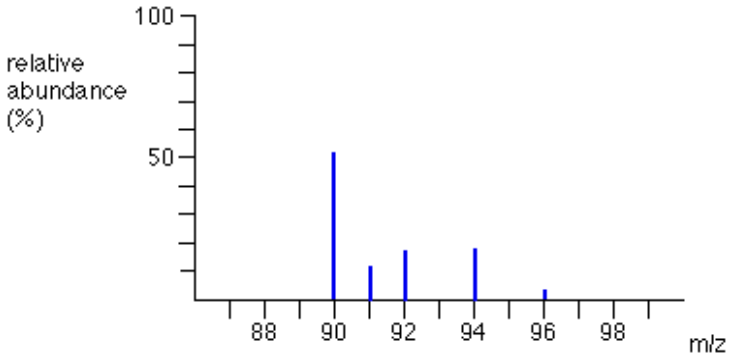
The two peaks in the mass spectrum shows that there are 2 isotopes of boron with relative isotopic masses of 10 and 11 on the ^{12}C scale.

The relative size of the peaks gives you a direct measure of the relative abundances of the isotopes. The tallest peak is often given an arbitrary height of 100 but you may find all sorts of other scales used; it doesn't matter. You can find the relative abundances by measuring the lines on the stick diagram.

In this case, the two isotopes (with their relative abundances) are:

boron-10	23
boron-11	100

Example 2: The mass spectrum for zirconium



The mass spectrum of zirconium may be used to know the number of zirconium isotopes and their relative abundances

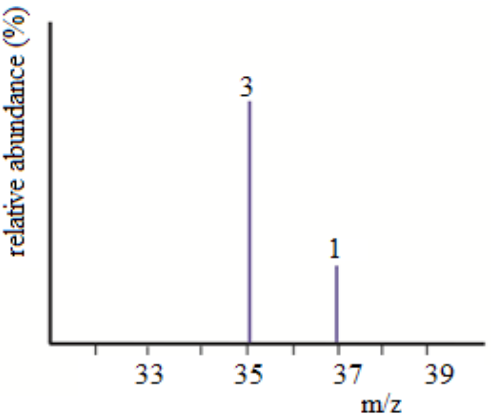
The five peaks in the mass spectrum show that there are five isotopes of zirconium with relative isotopic masses of 90, 91, 92, 94 and 96 on the ^{12}C scale.

This time, the relative abundances are given as percentages. Again you can find these relative abundances by measuring the lines on the stick diagram. In this case, the five isotopes (with their relative percentage abundances) are:

zirconium-90	51.5%	zirconium-92	17.1%	zirconium-96	2.8%
zirconium-91	11.2%	zirconium-94	17.4%		

Example 3: The mass spectrum of chlorine

Chlorine is taken as typical of elements with more than one atom per molecule. Chlorine has two isotopes, ^{35}Cl and ^{37}Cl, in the approximate ratio of 3 atoms of ^{35}Cl to 1 atom of ^{37}Cl. You might suppose that the mass spectrum would look like this:



But it is not true. The problem is that chlorine consists of molecules, not individual atoms. When chlorine is passed into the ionization chamber, an electron is knocked off the molecule to give a *molecular ion,* Cl_2^+. These ions won't be particularly stable, and some will fall apart to give a chlorine atom and a Cl^+ ion. The term for this is *fragmentation.*

$$Cl_2^+ \longrightarrow Cl + Cl^+$$

If the Cl *atom* formed isn't then ionized in the ionization chamber, it simply gets lost in the machine (neither accelerated nor deflected).

The Cl^+ ions will pass through the machine and will give lines at 35 and 37, depending on the isotope and you would get exactly the pattern in the last diagram. The problem is that you will also record lines for the *unfragmented* Cl_2^+ ions.

At the end the spectrum will show peaks due to ionized atoms, Cl^+ at 35, and 37, and ionized molecule Cl_2^+ at 70, 72, 74 as in the figure 15.12 below.

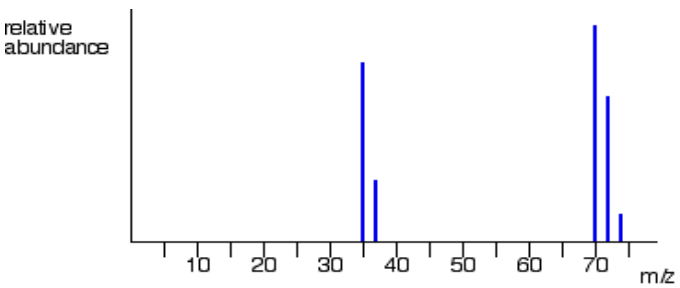
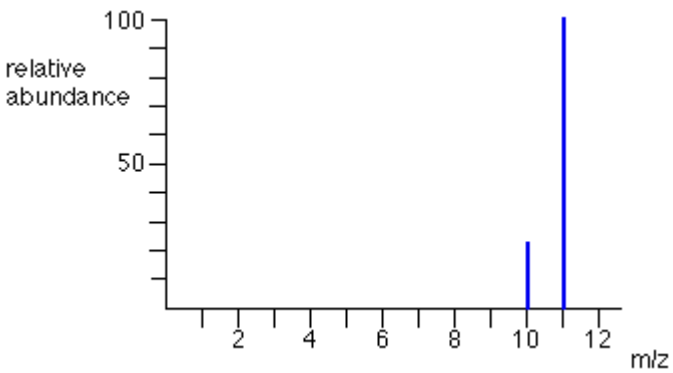


Figure 15.12 Mass spectrum of chlorine



Relative atomic masses can be calculated using the information from the mass spectrum.

Example 1: **The mass spectrum of boron is given below:**

Determine the relative atomic mass of boron

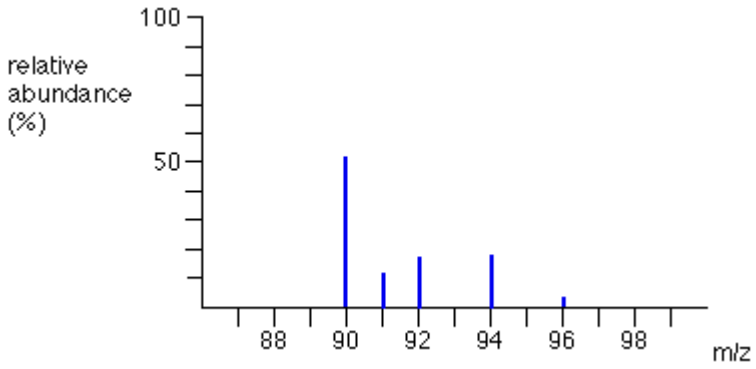
From the mass spectrum given, we have123 typical atoms of boron (sum of relative abundances). 23 of these would be ^{10}B and 100 would be ^{11}B.

The total mass of these would be (23 x 10) + (100 x 11) = 1330

The average mass of these 123 atoms would be 1330 / 123 = 10.8 (to 3 significant figures).

10.8 is the relative atomic mass of boron.

Example 2: **The figure below represents the mass spectrum of zirconium**



Suppose you had 100 typical atoms of zirconium (sum of relative abundances). 51.5 of these would be ^{90}Zr, 11.2 would be ^{91}Zr and so on. The total mass of these 100 typical atoms would be

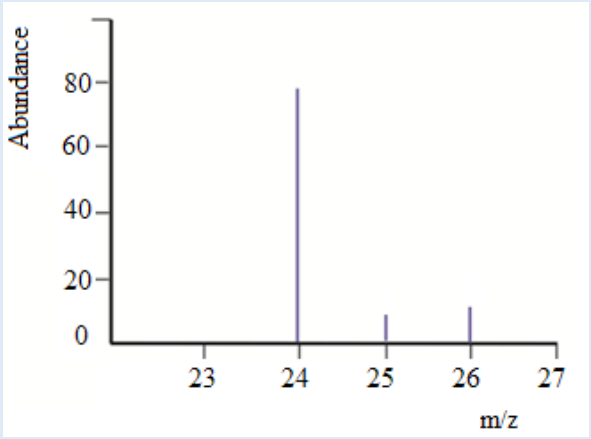
(51.5 x 90) + (11.2 x 91) + (17.1 x 92) + (17.4 x 94) + (2.8 x 96) = 9131.8

The weighted average mass of these 100 atoms would be 9131.8 / 100 = 91.3 (to 3 significant figures).

91.3 is the relative atomic mass of zirconium.

Application activity 15.4.2

1. The mass spectrum of magnesium is given below.



   b) How many isotopes does magnesium possess

   c) Estimate the isotopic mass of each of the magnesium isotopes

   d) Estimate the relative abundance for each of the isotopes of magnesium

2. Which of the following is not done through mass spectrometry?

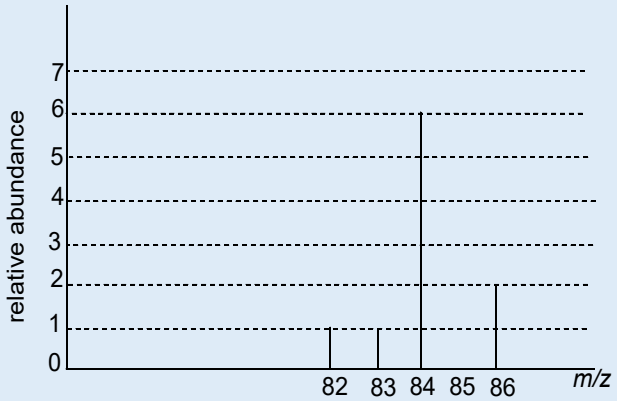
   e) Calculating the isotopic abundance of elements

   f) Investigating the elemental composition of planets

   g) Confirming the presence of O-H and C=O in organic compounds

   h) Calculating the molecular mass of organic compounds

3. When a pure, gaseous sample of element X is introduced in a mass spectrometer, four mononuclear, singly charged ions are detected as shown in the spectrum below.



i) Describe the process by which the gaseous sample of X is converted into ions in a mass spectrometer.

j) What adjustment is made to the operating system in order to direct the different ions, in turn, onto, the detector of a mass spectrometer?

k) What is the number of isotopes of element X? Justify your answer.

l) Use the data from the spectrum above to calculate the relative atomic mass of X.

m) Identify the element X.

Skills lab 15

When atoms of any element are subjected to high voltages, cathode rays are emitted and are the same for all elements.

During your free time, write an essay about all other radiations that may be emitted by atoms specifying the conditions under which they can be emitted and their properties. In your essay, you have to point out the benefits and the disadvantages of those radiations to humans and their environment.

End unit assessment 15

I. Multiple choice questions

1. Which of the following is true regarding a typical atom?
   a) Neutrons and electrons have the same mass.
   b) The mass of neutrons is much less than that of electrons.
   c) Neutrons and protons together make the nucleus electrically neutral.
   d) Protons are more massive than electrons
2. Which of the following statements is(are) *true?* For the false statements, correct them.
   a) All particles in the nucleus of an atom are charged.
   b) The atom is best described as a uniform sphere of matter in which electrons are embedded.
   c) The mass of the nucleus is only a very small fraction of the mass of the entire atom.
   d) The volume of the nucleus is only a very small fraction of the total volume of the atom.
   e) The number of neutrons in a neutral atom must equal the number of electrons.
3. Each of the following statements is true, but Dalton might have had trouble explaining some of them with his atomic theory. Give explanations for the following statements.
   a) Atoms can be broken down into smaller particles.
   b) One sample of lithium hydride is 87.4% lithium by mass, while another sample of lithium hydride is 74.9% lithium by mass. However, the two samples have the same chemical properties
4. In mass spectrometer, the sample that has to be analyzed is bombarded with which of the following?
   a) Protons
   b) Electrons
   c) Neutrons
   d) Alpha particles

5. Mass spectrometer separates ions on the basis of which of the following?

   a) Mass

   b) Charge

   c) Molecular weight

   d) Mass to charge ratio

6. In a mass spectrometer, the ions are sorted out in which of the following ways?

   a) By accelerating them through electric field.

   b) By accelerating them through magnetic field.

   c) By accelerating them through electric and magnetic field.

   d) By applying a high voltage.

7. The procedure for mass spectroscopy starts with which of the following processes?

   a) The sample is bombarded by electron beam.

   b) The ions are separated by passing them into electric and magnetic field.

   c) The sample is converted into gaseous state.

   d) The ions are detected.

8. Which of the following ions pass through the slit and reach the collecting plate?

   a) Negative ions of all masses.

   b) Positive ions of all masses.

   c) Negative ions of specific mass.

   d) Positive ions of specific mass.

9. Which of the following statements is not true about mass spectrometry?

   a) Impurities of masses different from the one being analyzed interferes with the result.

   b) It has great sensitivity.

   c) It is suitable for data storage.

   d) It is suitable for library retrieval.

10. In a mass spectrometer, the sample gas is introduced into the highly evacuated spectrometer tube and it is ionised by the electron beam.

    a) True

    b) False

II. Short and long answer questions

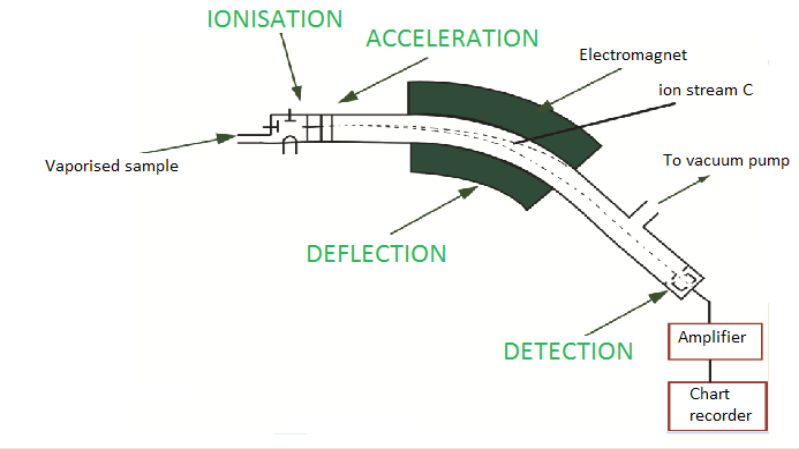
11. What are the three fundamental particles from which atoms are built? What are their electric charges? Which of these particles constitute the nucleus of an atom? Which is the least massive particle of the fundamental particles?

12. Verify that the atomic weight of lithium is 6.94, given the following information:

    ^{6}Li, mass = 6.015121 u; percent abundance = 7.50%

    ^{7}Li, mass = 7.016003 u; percent abundance = 92.50%

13. The diagram below shows the main parts of a mass spectrometer.



    a) Describe the different steps involved in taking a mass spectrum of a sample

        i. Which two properties of the ions determine how much they are deflected by the magnetic field? What effect does each of these properties have on the extent of deflection?

        ii. Of the three different ion streams in the diagram above, why is the ion stream C least deflected?

        iii. What would you have to do to focus the ion stream C on the detector?

    b) Why is it important that there is a vacuum in the instrument?

    c) Describe briefly how the detector works.

14.

(a) A mass spectrum of a sample of indium shows two peaks at m/z = 113 and m/z = 115. The relative atomic mass of indium is 114.5. Calculate the relative abundances of these two isotopes.

(b) The mass spectrum of the sample of magnesium contains three peaks with the mass-to-charge rations and relative intensities shown below

i. Explain why magnesium gives three peaks in mass spectrum?

m/z	25	25	26
relative intensities	1	0.127	0.139

ii. Use the information in the table above to calculate the accurate value for the relative atomic mass of magnesium

15. There exists 3 isotopes of oxygen that occur naturally with atomic mass 16, 17 and 18 with abundance 99.1% ; 0.89% and 0.01% respectively. Given that oxygen occurs naturally as diatomic molecule,

a) Predict the number of peaks that will be observed on the screen of mass spectrometer.

b) Show the molecular ions that are responsible of these peaks.

UNIT 16:

ELECTRON CONFIGURATIONS OF ATOMS AND IONS

Key Unit competence: Relate Bohr's model of the atom with hydrogen spectrum and energy levels, and interpret graphical information in relation to ionization energy of elements.

Introductory Activity 16

Observe the images below and answer the questions asked.

(a)Variation of potential energy up a ramp

(b)Variation of potential energy up steps

1. What can you see on the image above?
2. What type of motion is performed by the people on the image?
3. How does their potential energy change?

16.1. Bohr's atomic model: Concept of energy levels and spectra

Activity 16.1

1. Describe the Rutherford atomic structure?
2. What improvements have been brought by Bohr to the Rutherford atomic structure
3. Establish a relationship between the Bohr's atomic model and the illustration in the introductory activity.

16.1.1 Bohr's atomic model

From the observations in introductory activity 16, the potential energy of a person walking up ramp increases in uniform and continuous manner whereas potential energy of person walking up steps increases in stepwise and quantized manner. This can be explained by the values of energy which are continuous for the person walking up ramp while they are discrete (discontinued) for the person walkingup steps (Figures (a) and b of the introductory activity).

A **model** is a simplified representation used to explain the workings of a real world system or event. A model is useful because ***it helps you understand what's observed in nature***. It's not unusual to have more than one model represent and help people understand a particular topic.

There are two models of atomic structure in use today: the **Bohr model** and the **quantum mechanical model**. Of these two models, the Bohr model is simpler and relatively easy to understand.

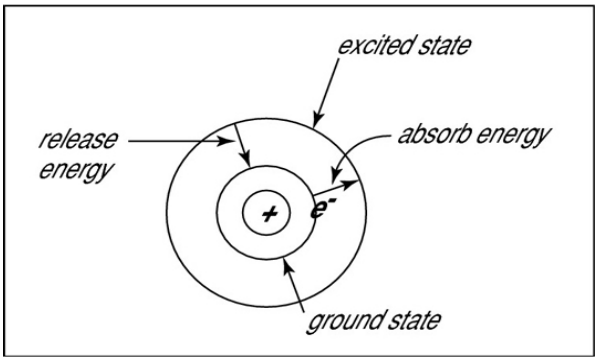
Have you ever bought colour crystals for your fireplace — to make flames of different colours? Or have you ever watched fireworks and wondered where the colours came from?

Colour comes from different elements. If you sprinkle table salt on a fire, you get a *yellow colour*. Salts that contain copper give a *greenish-blue flame*. And if you look at the flames through **a *spectroscope***, an instrument that uses a prism to break up light into its various components, you see a number of lines of various colours. Those distinct lines of colour make up a ***line spectrum***.

Niels Bohr, a Danish scientist, explained this line spectrum while developing a model for the atom:

- The Bohr model shows that the ***electrons in atoms are in orbits of differing energy around the nucleus*** (think of planets orbiting around the sun).
- Bohr used **the term *energy levels*** (or ***shells*****) to describe these orbits of differing energy**. He said that the energy of an electron is ***quantized***, meaning electrons can have one energy level or another but nothing in between.
- The energy level an electron normally occupies is called its ***ground state***. But it can move to a higher-energy, less-stable level, or shell, by absorbing energy. This higher-energy, less-stable state is called the electron's ***excited state***.

- After it is done being excited, the ***electron can return to its original ground state by releasing the energy it has absorbed***, as shown in the diagram below.



- Sometimes the energy released by electrons occupies the portion of **the *electromagnetic spectrum*** (the range of wavelengths of energy) that humans detect as **visible light**. Slight variations in the amount of the energy are seen as light of different colours.

The energy change is accompanied by absorption of radiation energy of $\mathbf{E=E_2\ E_1 = h}\ \upsilon$ where, h is a constant called 'Planck's constant' and υ is the frequency of radiation absorbed or emitted.

The value of $\mathbf{h}$ is $\mathbf{6.626\ x\ 10^{-34}}$ J.s. The absorption and emission of light due to electron jumps are measured by use of spectrometers.

Bohr found that the closer an electron is to the nucleus, the less energy it needs, but the farther away it is, the more energy it needs. So Bohr numbered the electron's energy levels.

The higher the energy-level number, the farther away the electron is from the nucleus and the higher the energy.

Bohr also found that the various energy levels can hold differing numbers of electrons: energy level 1 may hold up to 2 electrons, energy level 2 may hold up to 8 electrons, and so on.

Weakness of Bohr's Model

The Bohr model works well for very simple atoms such as hydrogen (which has 1 electron) but not for more complex atoms. Although the Bohr model is still used today, especially in elementary textbooks, a more sophisticated (and complex) model (the quantum mechanical model) is used much more frequently.

16.2.2. Absorption and emission spectra and energy associated

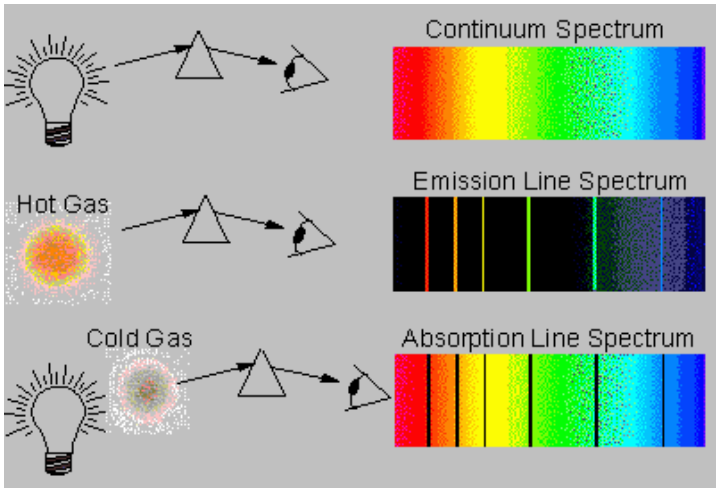
1. Atomic spectrum

The **atomic spectrum** is the range of characteristic frequencies or electromagnetic radiations that are readily absorbed and emitted by an atom.

A spectrum obtained from a glowing source is called an *emission spectrum*. When white light is passed through a prism we see a ***myriad of colours*** – specifically what we term to be a *rainbow*. This dispersion of white light demonstrates that white light contains all the wavelengths of colour and ***is thus considered to be continuous***. Each colour blends into the next with no discontinuity.

When elements are vaporized and then thermally excited, they emit light; however, this light was not in the form of a continuous spectrum as was observed with white light. Instead, a discrete **line spectrum** was seen when the light was passed through a narrow slit. A series of fine lines of different colours separated by large black spaces was observed. The wavelengths of those lines are characteristic of the element producing them – thus, elements can be identified based on the spectral line data that they produce.

Typically, we can examine the visible line spectra produced by an element in lab – using electricity, tubes filled with elements in the gaseous state and a spectroscope or diffraction grating which separates the light emitted by the gas into its components.



2. Emission and absorption spectra

Emission is the ability of a substance to give off light, when it interacts with heat. **Absorption** is the opposite of emission, where energy, light or

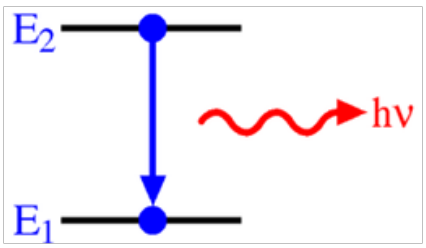
radiation is absorbed by the electrons of a particular matter.

Emission and absorption spectra are techniques that are used in chemistry and physics. Spectroscopy is the interaction of radiation and matter. Using spectroscopy, a scientist can figure out the composition of a certain matter. This is really beneficial, of dealing with unknown substances. Emission spectra and absorption spectra are different from each other but still related.

a. Emission spectrum

Every substances reacts differently when it interacts with light. The material starts off with being in the ground state, where all molecules are stable and settled. However when heat, energy or light is applied to a substance, some of the molecules transition into a higher energy state or an excited state. During this state ***the molecules are unstable and try to emit the energy in order to reach the state of equilibrium***. The **molecules emit energy in the form of photons or light**. The difference between the substance in ground state and excited state is then used to determine the emission level of the substance.

Each element or substances has a unique emission level or the amount of energy it radiates; this helps the scientists identify elements in unknown substances. The emission of an element is recorded on an **emission spectrum** or **atomic spectrum**. The emittance of an object measures how much light is emitted by it. The amount of emission of an object varies depending on the spectroscopic composition of the object and temperature. The ***frequencies on an emission spectrum are recorded in*** **light frequencies**, where the **colour of the light determines the frequency.**

The frequencies can be determined using the formula $\mathbf{E_{photon} = hv}$, where '$\mathbf{E_{photon}}$' is *the energy of the photon,* '**v**' is its *frequency,* and '**h**' is *Planck's constant.* Emission can happen in the form of light and rays, such as gamma and radio. The spectrum is a dark wavelength with bands of colour on it, which is used to determine the emission of the object	

b. Absorption spectrum

Absorption is ***the ability of a matter or electron to absorb light or radiation which makes them transition into a higher energy state***. Absorption is used to determine the absorption level of certain objects and their ability to retain heat.

Absorption spectrum is the plotting of the energy that is absorbed by an element or substance. Absorption **can be plotted in a *wavelength, frequency* or *wave number***. There are two types of absorption: *atomic absorption spectra* and *molecular absorption spectra*.

Absorption is used to determine the presence of a particular substance in a sample, or the quantity of the present substance in the sample. They are also used in molecular and atomic physics, astronomical spectroscopy and remote sensing. Absorption is primarily determined by the atomic and molecular composition of the material. They can also depend on temperature, electromagnetic field, interaction between the molecules of the sample, crystal structure in solids and temperature.

In order to determine the absorption level of a substance, ***a beam of radiation is directed at the sample and the absence of light that is reflected through the object can be used to calculate the absorption.*** The absorption spectrum **is usually light coloured, with dark bands that run through it**. These dark bands are used to determine the absorption of the object.

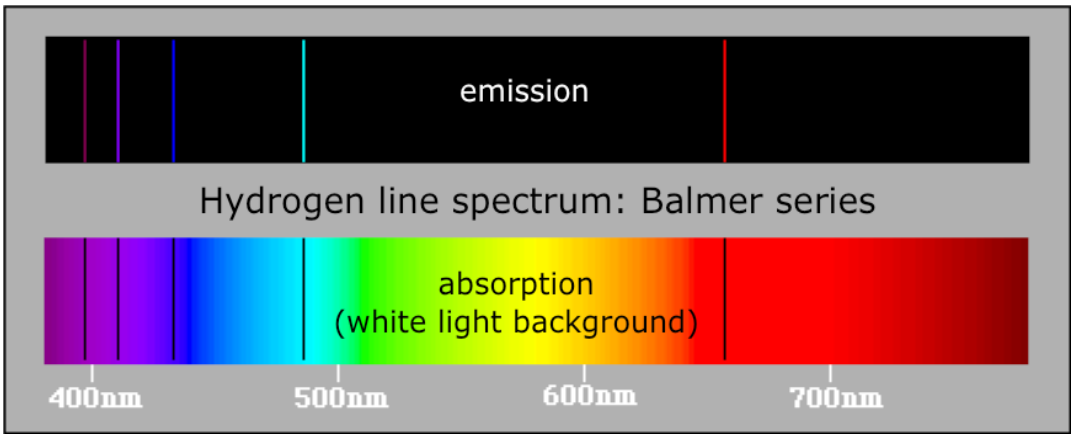


Table 16.1: Comparison between emission and absorption spectra

	Emission Spectra	Absorption Spectra
Description	Emission is the ability of a substance to give off light, when it interacts with heat.	Absorption is the opposite of emission, where energy, light or radiation is absorbed by the electrons of a particular matter.
Subjects	Chemistry and Physics	
Purpose	Can be used as part of spectroscopy to figure out the composition and the structure of a certain matter.	Can be used as part of spectroscopy to figure out the absorption level of certain objects and their ability to retain heat. Can also used in molecular and atomic physics, astronomical spectroscopy and remote sensing.
Types	_	Atomic absorption spectra and molecular absorption spectra.
Effect on molecules	When a substance interacts with light, then some of its molecules absorb the heat from the light and get excited. This causes them to become unstable and they try to emit the excess energy to get back to normal and stable state. The excited molecules release the excess energy in the form of photons, also known at light particles.	When a substance interacts with light, then some of its molecules absorb the light or the radiation. The types of light wavelength that get absorbed can be mapped.
Result	The type of photons emitted helps figure out the type of elements that the substance is formed of, as each element or substances has a unique emission level or the amount of energy it radiates	The type of light wavelengths that are absorbed helps figure out how much quantity of a substance is present in the sample.
In simple terms	Emission spectra records wavelengths emitted by materials, which had been stimulated by energy before.	Absorption spectra records the wavelengths absorbed by the material.

Looks like	*Dark coloured*, with light bands that run through it. These light bands are used to determine the types of photons emitted by the object.	*Light coloured*, with dark bands that run through it. These dark bands are used to determine the absorption of the object.
Units	The frequencies of emission can be determined using the formula $\mathbf{E_{photon} = hv}$	Can be plotted in a wavelength, frequency or wave number.

Application activity 16.1

1. Find out two more examples that you can use to illustrate the concept of quantization.
2. Discuss the main weakness of:

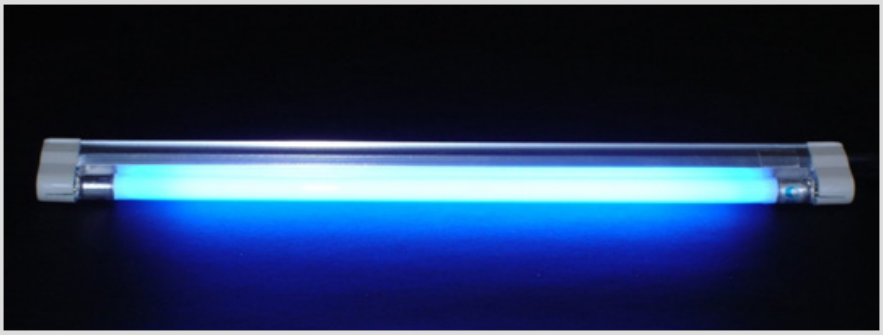
c) Rutherford's nuclear atom.

d) Bohr's atomic model.

16.2. Hydrogen spectrum and spectral line series

Activity 16.2

Look at the picture of neon tube light below and do research about how this neon tube light works to produce light and present your findings.



(Source: https://www.amazon.com/Blue-Replacement-Light-Purchased-Length/dp/B00AWPI2F8)

Bohr's atomic model allows explaining the emission spectra of atoms. This happens when excited electrons lose energy in form of electromagnetic radiation and fall to lower energy levels.

The wave-particle nature of the light

Light as a wave

The light is a wave-like phenomenon as shown in Figure 16.1.

It is characterized by its wave length, generally symbolized by the Greek letter lambda, λ, and its frequency, represented by the Greek letter nu, ν.

As shown in the Figure 16.1 below, the wavelength represents the distance between two successive summits/peaks (or two successive troughs).

The frequency represents the number of complete wavelengths made by the light per second, also called cycles per second.

Visible light is composed by different visible lights with different λ and ν.

But all those lights have the same speed, the speed of light, which, in a vacuum, is equal to: 3.00×10^8m/s; although different types of light have different λ and ν, they move at the same speed c. This results in the relation between the speed of light and its wavelengthand frequency: $c = \nu\lambda$

From this relation, and since c is constant, we can conclude that:

- Light with long wavelength has low frequency, whereas
- Light with short wavelength has high frequency.

Let's take an example to illustrate: light$_1$ has λ_1 equal to 10^5m whereas light$_2$ has λ_2 equal to 10^{-5}m. After 1 second, both would have travelled 3.00×10^8m, the speed of light, but their frequencies will be different:

$\nu_1 = c/\lambda_1 = 3.00 \times 10^8 ms^-/10^5 m = 3.00 \times 10^3 s^- = 3{,}000$ cycles/s

$v_2 = c/\lambda_2 = 3.00 \times 10^8 ms^-/10^{-5} m = 3.00 \times 10^{13} s^- = 30{,}000{,}000{,}000{,}000$ cycles/s

Light is energy, represented by: $E = h\nu$ where h is Plank's constant ($h = 6.626 \times 10^{-34}$Js)

Hence energy in the light is proportional to its frequency; the higher the frequency of the light, the higher is its energy and vice-versa.

The different colours of the visible light differ by their wavelength as shown in the Figure 16.2

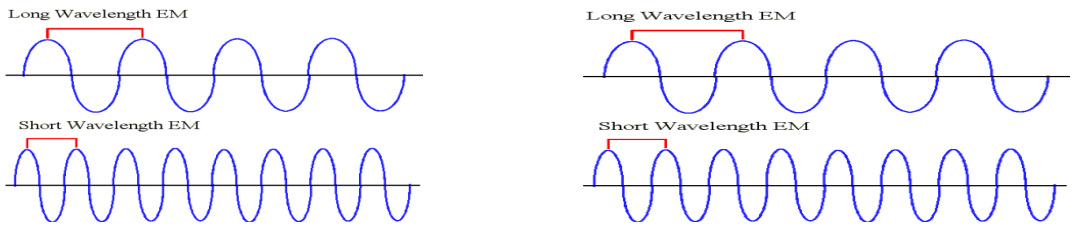


Figure 16.1: Wave-like phenomenon

As illustrated in Figure 16.2 below, the right side of the spectrum consists of high-energy, high-frequency and short wavelength radiations. Conversely, the left side consists of low-energy,low-frequency and long wavelength radiations.

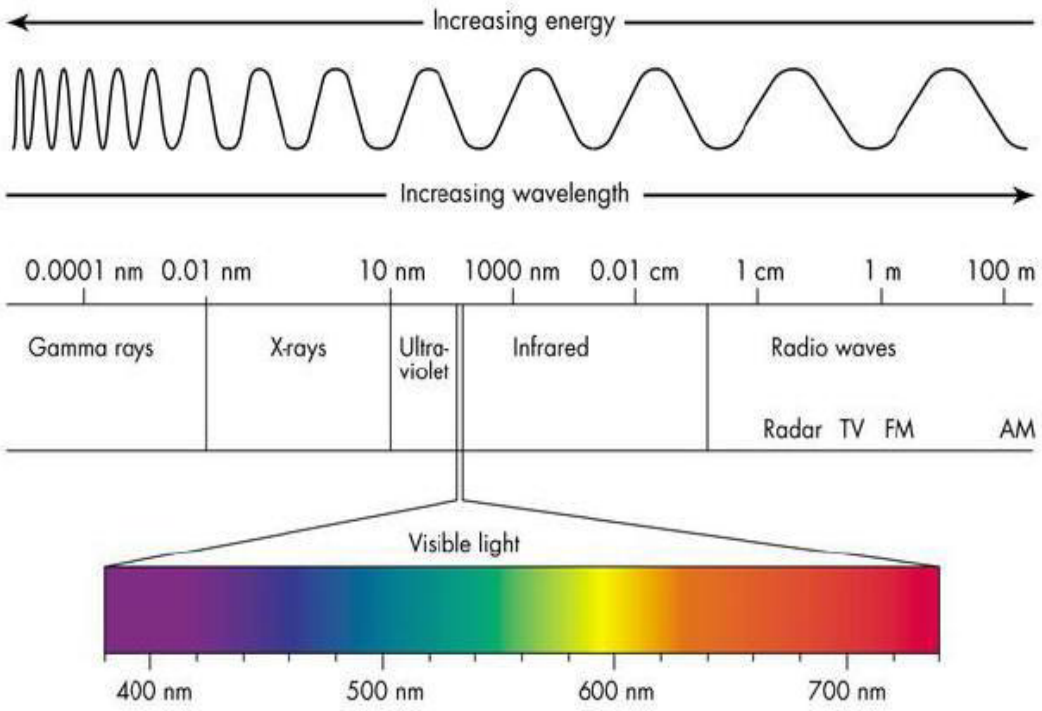


Figure 16.2: Electromagnetic radiation

(Source: http://psychelic-information-theory.com/em_spectrum)

When an electron is excited or de-excited, the energy absorbed or emitted corresponds to the difference of energy, ΔE, between the final energy level of the electron, E_2, and the starting energy level of the electron,

E_1: $E_2 - E_1 = \Delta E = h\nu$. ΔE is positive when $E_2 > E_1$, this is the case of absorption and excitation of electron; on the other hand ΔE may be negative when $E_2 < E_1$, in case of emission and de-excitation of electron.

Figure 16.3 below shows the different series of emission spectra of hydrogen. As you can see, the difference between those series is the final energy level where the electron fall after de-excitation.

The series have been named according to the scientists who discovered them. Ionization of an atom or loss of an electron corresponds to excitation of an electron to the level n=∞.

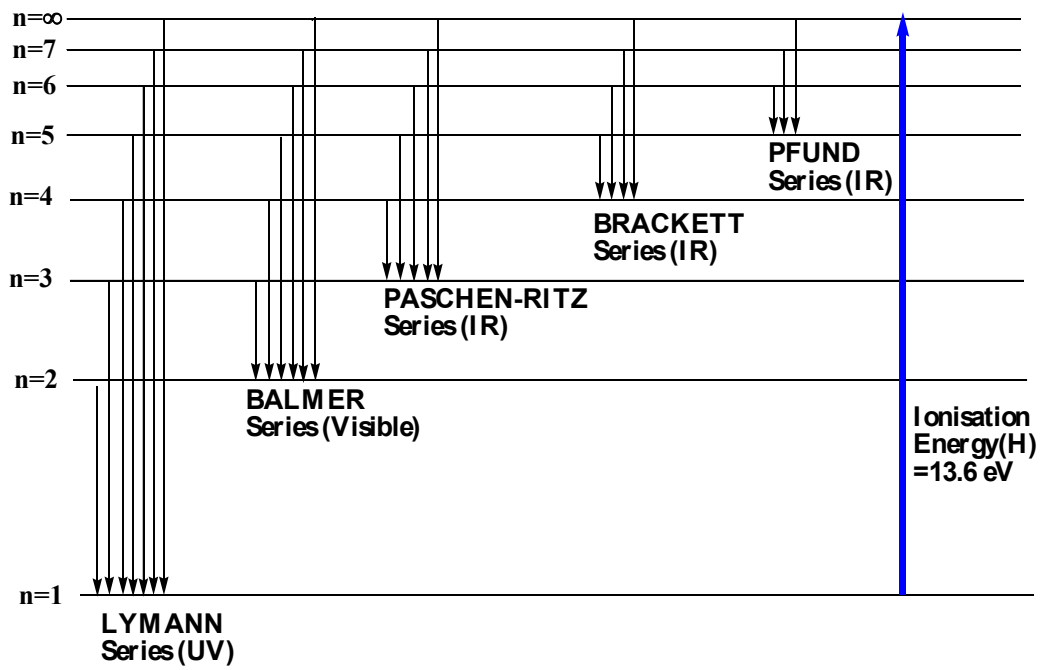


Figure 16.3: Energy-Level diagram and emission line spectra for the hydrogen atom

The wavelength of lines in the series is given by

$$\frac{1}{\lambda} = R_H\left(\frac{1}{n_1^{\,2}} - \frac{1}{n_2^{\,2}}\right)$$

where, λ is the wave length

$\mathbf{R_H}$, is the Rydberg's constant ($109667 cm^{-1}$)

$\mathbf{n_1}$ = final energy level and n_2 = initial energy level

Examples

1. Find the wavelength and frequency in Balmer series associated with a drop of an electron from the fourth orbit.

Answer

$$\frac{1}{\lambda} = R(\frac{1}{n_1^{\,2}} - \frac{1}{n_2^{\,2}})$$

$$= 109667(\frac{1}{2^2} - \frac{1}{4^{\,2}})$$

$$= 109667(\frac{1}{4} - \frac{1}{16})$$

$$= 109667x\frac{3}{16}$$

$$\frac{1}{\lambda} = \frac{20562.5625}{1}$$

$\lambda = 4.86 \text{ X } 10^{-5}\text{cm}$

$$\upsilon = \frac{C}{\lambda}$$

$$= \frac{3\,x\,10^{8}\,m/s}{4.86\,x\,10^{-7}\,m}$$

$= \mathbf{6.17\ x10^{14}\ s^{-1}}$

2. Find the wave length, frequency and energy of the third line in the Lyman series.

Answer

$$\frac{1}{\lambda} = R(\frac{1}{n_1^{\,2}} - \frac{1}{n_2^{\,2}}) = 109667(\frac{1}{1^2} - \frac{1}{4^{\,2}}) = 109667(\frac{1}{1} - \frac{1}{16}) = 109667x\frac{15}{16}$$

$$\frac{1}{\lambda} = \frac{102812.8125}{1} \quad \lambda = \frac{1}{102812.8125}$$

$\lambda = \mathbf{9.73\ x\ 10^{-6}cm}$

$$\upsilon = \frac{C}{\lambda} = \frac{3\ x\ 10^{8}\ m/s}{9.73\ x\ 10^{-8}\ m} = \mathbf{3.08\ x\ 10^{15}\ s^{-1}}$$

$E = h\upsilon = 6.63 \times 10^{-34} \times 3.08 \times 10^{15} = 2.042 \times 10^{20}$J/atom

Application activity 16.2

1. What is the meaning of infinity level in the hydrogen spectral lines?
2. Given a transition of an electron from n=5 to n=2. Calculate
   c) Energy
   d) Frequency
   e) Wavelength
3. What is the wavelength (in nanometres) of a photon emitted during a transition from the ni = 5 state to the nf = 2 state in the hydrogen atom?

16.3. Quantum theory of the atom

Activity 16.3

1. Recall the Bohr's model of an atom.
2. What were the limitations of that model?
3. What improvements other scientists have brought to that model?

In 1913, physicist **Niels Bohr** described as an atom as a small, positively charged nucleus surrounded by electrons that travel in circular orbits around the nucleus—similar in structure to the solar system, but with attraction provided by electrostatic forces rather than gravity. An electron must absorb or emit specific amounts of energy to transition between these fixed orbits.

1. Despite its success in accounting for spectral lines of the H atom, the Bohr model failed to predict the spectrum of any other element. The model worked well for one-electron species, but not for atoms or ions with more than one electron.
2. According to Bohr, the circular orbits in which electrons revolve are planar.
3. Bohr's theory fails to account for Zeeman Effect and Stark Effect.
4. Bohr assumes that the electron revolves around the nucleus in circular orbits at fixed distance from the nucleus and with a fixed velocity.

In 1926, **Erwin Schrödinger** used this idea to develop a mathematical model of the atom how electrons move in wave form, and developed the Schrodinger equation which describes how the quantum state of a system changes with time.

This was the beginning of **Quantum Mechanics** (or Quantum Theory). A consequence of using waveforms to describe particles is that it is impossible to obtain precise values for both the position and velocity of a particle at the same time. This became known as the "**uncertainty principle**" formulated by Werner Heisenberg in 1926. This model was able to explain observations of atomic behaviour that previous models could not. Afterwards the planetary model of the atom was discarded in favour of one that described atomic orbital zones around the nucleus where a given electron is most likely to be observed.

Thus, the quantum mechanical model is based on mathematics. Although it is more difficult to understand than the Bohr model, it can be used to explain observations made on complex atoms.

Max Planck proposed that energy emitted is not done so in a continuous manner but is given off in small packets which he called **quanta**. He determined that an atom can emit only certain amounts of energy and therefore they must contain certain quantities of energy and that those are fixed. Thus, the energy of an atom is quantized. The change in the atom's energy results from the gain or loss of one or more packets of energy. Planck derived an equation to explain this quantized form of energy (as opposed to the idea that energy emitted was continuous)

$$E = h\nu$$

where h = Planck's constant = 6.626 x 10^{-34} J•s and ν = frequency (as above)

According to quantum theory, energy is always emitted in integral multiples of $h\nu$; ($h\nu$, 2 $h\nu$, 3 $h\nu$ . . .), but never, for example, 1.67 $h\nu$ or 4.98 $h\nu$. At the time Planck presented his theory, he could not explain why energies should be fixed or quantized in this manner.

Therefore, despite the fact that Planck thought that energy was quantized, physicists continued to think of energy as travelling in waves. Energy as waves, however, could not explain the photoelectric effect. The quantum mechanical model is based on quantum theory, which says matter also has properties associated with waves. According to quantum theory, it is impossible to know the exact position and momentum of an electron at the same time. This is known as the Uncertainty Principle.

The quantum mechanical model of the atom uses complex shapes of orbitals(sometimes called electron clouds), volumes of space in which there is *likely* to be an electron. So, this model is based on probability rather than certainty.

The existence of discrete atomic energy levels is retained from Bohr's model in the current atomic model.

The allowed wave-like motion of the electron leads to an atom with certain fixed energy states much like Bohr assumed. The electron's exact location cannot be determined.

Solutions of Schrödinger's wave equation are functions that describe atomic orbitals.

Each function describes a fixed-energy state the electron can occupy and gives the probability of finding it in a given 3-dimensional space.

Applying wave mathematics to the electron wave, Erwin Schrödinger derived an equation that is the basis for the quantum-mechanical model of hydrogen atom.

This probability can be shown pictorially by means of an electron probability density diagram, or simply, an electron density diagram.

Electron probability density in the H atom ground state.

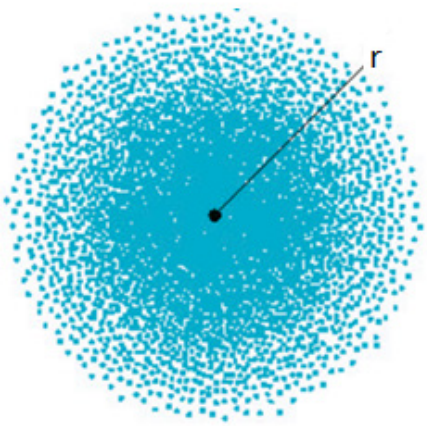


Figure 16.4: Electron "cloud" model

The distinction between atomic orbital and probability of the electron being at a distance r from the nucleus.

The probability of the electron being far from the nucleus is very small, but not zero.

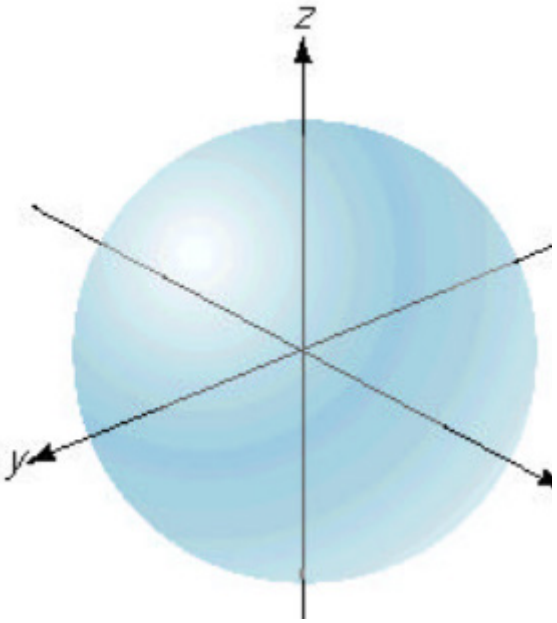


***Figure 16**:5:A 90% contour depiction of the ground state; the volume in which the electron spends 90% of its time*

***Source:** https://www.uh.ed*

The maximum radial probability for the ground-state of H atom appears at (0.529 Å, or 5.29 x 10^{-10} m), the same as 1st Bohr orbit.

The electron spends most of its time at the same distance that the Bohr model predicted it spent all of its time. Each atomic orbital, has a distinctive radial probability distribution and probability contour diagram.

Four numbers, called ***quantum numbers***, were introduced to describe the characteristics of electrons and their orbitals:

Principal quantum number: **n**

Angular momentum quantum number: **ℓ**

Magnetic quantum number: m_ℓ

Spin quantum number: m_s

Application activity 16.3

Discuss the improvements brought by the quantum theory in the understanding of the atomic structure

16.3.1. Quantum numbers for energy levels, sub-energy levels and orbitals

Activity 16.3.1

1. Explain the difference between orbit and orbitals.
2. The energy of both orbit and orbital is determined by a number n known as principal quantum number. Explain.
3. An electron in an orbital is even described by other quantum numbers. What are they?
4. Do research to precise the relationship between these quantum numbers, their role in the description of electrons in orbitals and their number in each energy level.

We have seen the weakness and critics against the atomic Bohr's model. In order to answer the questions not answered by that model, other atomic models were proposed. One of those models is the Quantum model that has been developed by the Australian physicist Erwin Schrödinger (1887-1961). The model is based on a mathematical equation called Schrödinger equation. This model is based on the following assumptions or hypotheses:

An electron is in continuous movement around the nucleus but cannot be localized with precision; only the high probability of finding it in a certain region around the nucleus can be known.

The region where the probability of finding electron is high, at more than 95%, is called "orbital"; in other words, the orbital is the volume or the space (three-dimensional) around the nucleus where there is a high probability of finding the electron.

Without going into the mathematical development of the Schrödinger equation, we can say that the energy of the electron depends on the orbital where it is located. And an atomic orbital is described by a certain number of "quantum numbers" according to the solution of Schrodinger equation, i.e. 3 whole numbers:

1. **The principal quantum number** (n) is a positive integer which varies from 1 to ∞. The principal quantum number indicates the energy level in an atom where electrons can be located: the higher the n value, the higher the energy level. An electron in energy level n=1 has lowest energy in an atom. The principal quantum number, n, has been traditionally given names by the letters: K (n=1), L (n=2), M (n=3), N (n=4), O (n=5), P (n=6).

In the Bohr's atomic model, K, L, M, … were used to represent different orbits or shells of electrons. Later on, the term shell sometimes is used to describe a group of orbitals with the same principal quantum number. The term subshell describes a group of orbitals with the same principal and second quantum number. The maximum number of orbitals and electrons that can be found in an energy level n are n^2 and $2n^2$, respectively (Table 16.2). The maximum number of sub shells in an energy level n equals n.

In summary:

n = principal quantum number; all orbitals with same n are in the same shell.

l = secondary (azimuthal) quantum number; divides shells into subshells.

ml = magnetic quantum number; divides subshells into individual orbitals.

Table 16.2: Relation between the principal quantum number, the number of orbitals and the maximum number of electrons.

N	Maximum Number of subshells	Maximum Number of orbitals	Maximum Number of electrons
1	1	1	2
2	2	4	8
3	3	9	18
4	4	16	32

2. The angular momentum quantum number *(l)*

The second quantum number is the angular quantum number represented by the letter, l: it is an integer which can take any value from zero or higher but less than n-1, i.e. equal to: 0, 1, 2, 3,….up to n-1. For example if n= 1, l is equal to 0, if n = 2, l can be 0, 1.

It is also called secondary or azimuthal quantum number. It indicates the shape of the orbital and is sometimes called the orbital shape quantum number. By tradition, those different shapes of orbitals have been given names or letter symbols: l = 0 = s, l =1 = p, l = 2 = d, l=3 = f

3. Magnetic quantum number (m_l)

The **magnetic quantum number** describes the **spatial orientation of the orbital**. It is an integer that **varies from -*l* to +*l*.** For example if: $l = 0$, m_l can only be 0; if $l = 1$, $m_l = -1, 0, +1$; if $l=2$, $m_l = -2, -1, 0, 1, 2$. As you can see for each value of l there are $(2l+ 1)$ values of m_l corresponding to $(2l + 1)$ orientations under the influence of magnetic field.

It is called the *magnetic* quantum number because the effect of different orientations of orbitals was first observed in the presence of a magnetic field.

Table 16.3: Relationship between the principle quantum (n), the azimuthal quantum number (l) and magnetic quantum number (m_l)

n	Possible values of *l*	Sub-shell designation	Possible values of m_l	Orbital designation	Number of orbitals in sub-shell	Total of orbitals in shell
1	0	1s	0	1s	1	1
2	0	2s	0	2s	1	
	1	2p	-1, 0, +1	2p	3	4
3	0	3s	0	3s	1	
	1	3p	-1, 0, +1	3p	3	
	2	3d	-2, -1, 0, +1, +2	3d	5	9
4	0	4s	0	4s	1	
	1	4p	-1, 0, +1	4p	3	
	2	4d	-2, -1, 0, +1, +2	4d	5	
	3	4f	-3, -2, -1, 0, +1, +2, +3	4f	7	16

The table 16.3 shows that, apart s sub-level that has only one orbital, other sub-levels have a certain number of different orbitals; those orbitals have the same energy but differ in their specific orientations. Example p orbitals are 3 with different orientations: $p_x p_y p_z$.

4. The spin quantum number (m_s)

The fourth quantum number is the spin quantum number, represented by the symbol m_s. It describes the spin of an electron that occupies a particular orbital. The electron behaves as a spinning magnet. The spin quantum number is the property of the electron, not the orbital.

This number describes *the spinning direction of the electron in a magnetic field. The direction could be* either ***clockwise*** or ***counter clockwise.*** The electron behaves as if it were spinning about its axis, thereby generating a magnetic field whose direction depends on the direction of the spin. The two directions for the magnetic field correspond to the two possible values for the spin quantum number, m_s (S).

Only two values are possible: $m_s = +1/2$ and $-1/2$ as shown in the Figure 16.4 below.

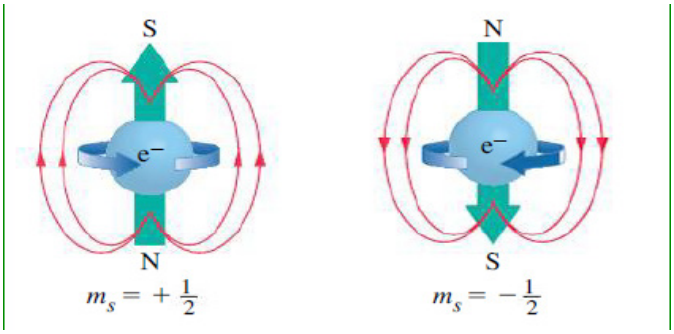


Figure 16.4: Electron spin (Source: Petrucci et al., 2011)

m_s = +1/2 and m_s = -1/2 are commonly represented by ↑ and ↓ respectively.

In conclusion an electron in any given atom is described by 4 quantum numbers: (i) three quantum numbers which describe the orbital where the electron is located: n, *l* and m_l and (ii) one quantum number describing the spin of the electron, m_s.

- Values: +1/2 or – ½
- Electrons will spin opposite each other in the same orbital

Application activity 16.3.1

1. Which of the following sets of quantum numbers are not allowed? For each incorrect set, state why it is incorrect.
   a) (1, 0, ½, ½)
   b) (3, 0, 0, ½)
   c) (2, 2, 1, ½)
   d) (4, 3, -2, ½)
   e) (3, 2, 1, 1).
2. Which of the following are possible sets of quantum numbers?
   a) n = 1, l =1, m_l = 0, m_s = +1/2
   b) n = 2, l =0, m_l = 0, m_s = +1/2
   c) n = 3, l =2, m_l = −2, m_s = −1/2
   d) n = 3, l = 3, m_l = 0, m_s = -1/2
   e) n = 4, l = 3, m_l = 2, m_s = -1/2
   f) n = 4, l = 1, m_l = 1, m_s = +1/2
   g) n = 2, l = 1, m_l = -1, m_s = -1
   h) n = 5, l = -4, m_l = 2, m_s = +1/2
   i) n = 3, l = 1, m_l = 2, m_s = -1/2

16.3.2. Number and shape of "s" and "p" orbitals

Activity 16.3.2

1. Explain the terms orbital and quantum numbers.
2. How the terms explained in 1. above are related?
3. Do research and point out the possible shapes assigned to s and p orbitals.

An orbital is a region of space around the nucleus in which there is a great chance (high probability) to find an electron. It is described by three quantum numbers.

These quantum numbers describe the size, shape, and orientation in space of the orbitals on an atom.

The principal quantum number (n) describes the size of the orbital.

The angular quantum number (l) describes the shape of the orbital.

The magnetic quantum number (ml), describes the spatial orientation of a particular orbital.

All electrons that have the same value for n (the principal quantum number) are in the same shell. Within a shell (same n), all electrons that share the same value for l(the angular momentum quantum number, or orbital shape) are in the same subshell. The electrons in subshell have same principal quantum number, same azimuthal quantum number and differ in magnetic and spinquantum number. When electrons share the same n, l, and m, they are said to be in the same orbital. That is, they have the same energy level, shape, and orientation. The electrons in the same orbital differ only in spin quantum number.

Chemists describe the shell and subshell in which an orbital belongs with a two-character code such as $2p$ or $4f$. The first character indicates the shell ($n = 2$ or $n = 4$). The second character identifies the subshell. By convention, the following lowercase letters are used to indicate different subshells.

$$s: l = 0$$

$$p: l = 1$$

$$d: l = 2$$

$$f: l = 3$$

The number of subshells in a shell is equal to the principal quantum number for the shell. For example the n = 3 shell, contains three subshells: the 3s, 3p, and 3d orbitals.

There is only one orbital in the n = 1 shell because there is only one way in which a sphere can be oriented in space. The only allowed combination of quantum numbers for which n = 1 is the following.

n	l	m	
1	0	0	1s

There are four orbitals in the n = 2 shell.

n	l	m	
2	0	0	2s
2	1	-1	
2	1	0	2p
2	1	+1	

There is only one orbital in the 2s subshell. But, there are three orbitals in the 2p subshell because there are three directions in which a p orbital can point. One of these orbitals is oriented along the X axis, another along the Y axis, and the third along the Z axis of a coordinate system, as shown in the figure below. These orbitals are therefore known as the $2p_x$, $2p_y$, and $2p_z$ orbitals.

The number of orbitals in a shell is the square of the principal quantum number: $1^2 = 1$, $2^2 = 4$, $3^2 = 9$. There is one orbital in an s subshell (l = 0), three orbitals in a p subshell (l = 1), and five orbitals in a d subshell (l = 2). The number of orbitals in a subshell is therefore 2(l) + 1.

Orbitals have shapes that are best described as spherical (l = 0), polar (l = 1), or cloverleaf (l = 2).

Shapes of “s” and “p” orbitals

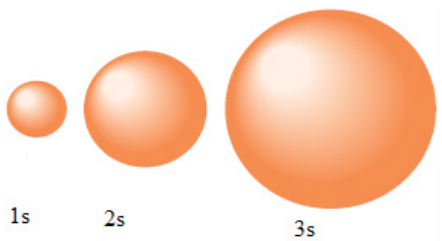


Figure 16.4: Representation of s orbitals (Source: Zumdahl S.& Zumdahl A., 2007)

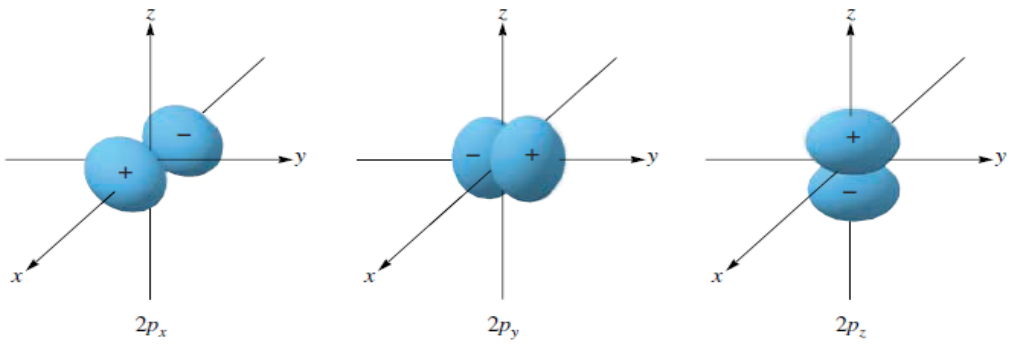


Figure 16.5: Representation of p orbitals (Source: Zumdahl S.&Zumdahl A., 2007)

Application activity 16.3.2

1. Determine the number of orbitals that can exist at energy level corresponding to n=1, 2, 3.
2. Outline diagrams representing the **s** and **p** orbitals.

16.4. Electronic configuration of atoms and ions

16.4. 1. Rules governing the electronic configurations

Activity 16.4.1

1. Using your knowledge acquired so far in chemistry, define the term electron configuration.
2. How this concept can apply to the quantum theory of the atomic structure?
3. Using available resources, figure out the rules that dictate the electronic configuration.

The electron configuration is the distribution of electrons of an atom in its atomic orbitals. The electronic configuration of an atom is governed by three main rules including Aufbau principle, Pauli Exclusion Principle and Hund's rule.

1. Pauli Exclusion Principle

No two electrons in the same atom can have the same set of the four quantum numbers. If two electrons have the same values of n, l, ml, they must have

different values of ms. Then, since only two values of msare allowed, an orbital can hold only two electrons, and they must have opposite spins.

2. Hund's rule

Electrons occupy all the orbitals of a given sublevel singly before pairing begins.

Spins of electrons in different incomplete orbitals are parallel in the ground state. The most stable arrangement of electrons in the subshells is the one with the greatest number of parallel spins.

That is, the most stable arrangement for electrons in orbitals of equal energy (degenerate) is where the number of electrons with the same spin is maximized • Example: Carbon - 6 electrons

For example, for an element with atomic number equal to 6, the electronic configuration is:$1s^2 2s^2 2p_x^1 p_y^1 p_z^0$

When building the electronic configuration of elements, the above principles are applied. writing the principal quantum number in Arabic number, followed by the orbitals immediately followed by the number of electrons in the orbital as superscript.

An atom X: $1s^2$: has only two electrons in s orbital at the 1^{st} energy level

An atom Y: $1s^2 2s^2 2p^3$: has electrons in 2 levels of energy: level n=1, and level n= 2. In level 1, it has 2 electrons in s orbital. In level 2, it has 2 electrons in s orbital and 3 electrons in p orbitals.

Figure 2.8 is a useful and simple aid for keeping track of the order in which electrons are first filled for each atomic orbital. The different orbitals are filled in the order *1s, 2s, 2p, 3s, 3p, 4s, 3d, 4p, 5s, 4d, 5p, 6s, 4f, 5d, 6p, 7s, 5f, 6d, 7p*.

Notice that as energy levels increase starting from n=3, 4s orbital is filled before 3d, 5s before 4d, etc… as shown in the diagram below. But when ionized, 4s electrons are ionized before 3d, and 5s before 4d.

3. Aufbau Principle

The Aufbau principle or build up principle or construction principle state that "Electrons fill lower energy orbitals (closer to the nucleus) before they fill higher energy ones".

Electron filling order

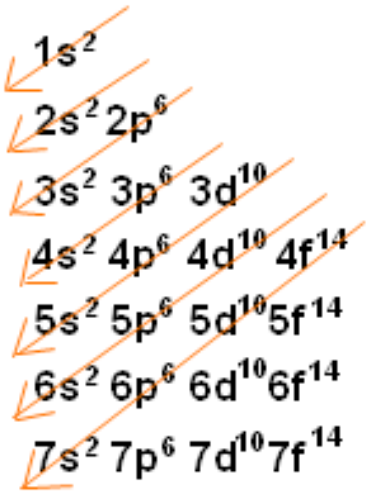


Figure 16.6: The order of filling orbitals with electrons

Orbitals in atomic ground state electron configuration are filled in order of increasing $n+l$. For equal $n+l$ values, the orbital with the lower n is most often filled first.

Examples:

Using, s, p, d, f notation, write the electronic configuration for elements of atomic numbers: 16, 23, 37

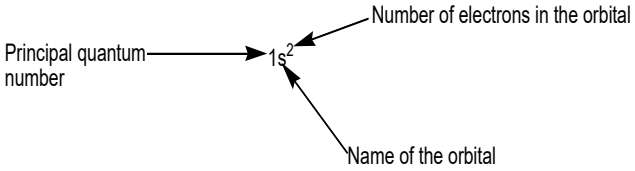
(Z = 16): $1s^2 2s^2 2p^6 3s^2 3p^4$

(Z = 23): $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^3$

(Z = 37): $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^{10} 4p^7$

The s, p, d and f notation uses numbers to designate a principal shell and the letters to identify a subshell; a superscript number indicates the number of electrons in a designated subshell

Example:



Note that $1s^2$ is read "one s two," not "one s squared."

Application activity 16.4.1.a

1. Using the spdf notation, build the electronic configuration of the following atoms: $_{1}H$, $_{3}Li$, $_{5}B$, $_{11}Na$, $_{18}Ar$, $_{19}K$, $_{21}Sc$, $_{24}Cr$, $_{26}Fe$, $_{29}Cu$

Expanded notation

Expanded notation is another method of writing the s, p, d and f notation. The method uses the same concept as s, p, d and f notation except that each individual orbital of a sub-level having many orbitals is represented with a subscript letter indicating the orientation of the orbital. This applies for p, d, and f orbitals.

Considering that p-orbital has three components p_x, p_y, and p_z, the expanded electronic configuration of some elements is given hereafter.

1. H (Z=2): $1s^1$
2. He (Z=3): $1s^2$
3. Li (Z=3): $1s^2 2s^1$
4. Be (Z=4): $1s^2 2s^2$
5. B (Z=5): $1s^2 2s^2 2p_x^1$
6. C(Z=6): $1s^2 2s^2 2p_x^1 2p_y^1$
7. N (Z=7): $1s^2 2s^2 2p_x^1 2p_y^1 2p_z^1$
8. O (Z=8): $1s^2 2s^2 2p_x^2 2p_y^1 2p_z^1$
9. F(Z=9): $1s^2 2s^2 2p_x^2 2p_y^2 2p_z^1$
10. Ne(Z=10): $1s^2 2s^2 2p_x^2 2p_y^2 2p_z^2$
11. $Al^{3+}(Z=13)$: $1s^2 2p_x^2 p_y^2 p_z^2$
12. $S^{2-}(Z=16)$: $1s^2 2s^2 2p_x^2 p_y^2 p_z^2 3s^2 3p^6$

Application activity 16.4.1.b

Write the expanded electronic configuration for each of the following atom/ions.

S(z=16), P^{3-}(z=15), Mg^{2+}(z=12)

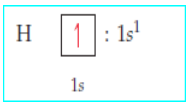
Orbital box representation

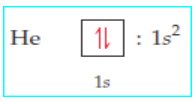
An orbital box representation consists of a box for each orbital in a given energy level, grouped by sublevel, with an arrow indicating an electron and its spin.

Note that two electrons in the same orbital have necessarily opposite spins as indicated in the examples below.

The table 16.4 shows the electronic configuration of some elements using orbital box representation and applying Hund's rule.

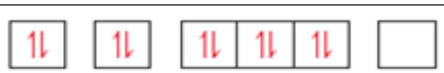
Examples





N.B:An orbital box representation doesn't show the real form of the orbital; the forms of the different orbitals are shown in Figures 16.4, 16.5 and 16.6 above.

Table 16.4: Electronic configuration using orbital box representation

Element	Total electrons	Orbital diagram
		1s 2s 2p 3s
Li	3	
Be	4	
B	5	
C	6	
N	7	
Ne	10	
Na	11	

Application activity 16.4.1.c

Using boxes to represent orbitals, draw the electronic configuration of

N^{3-} (z=7), Ti^{4+}(z=22), Mg^{2+}(z=12), Ar(z=18)

Identify the isoelectronic species that are present.

Noble Gas Notation or condensed electron configuration

All noble gases have completely filled subshells and can be used as a shorthand way of writing electron configurations for subsequent atoms.

When using this method, the following steps are respected.

a). Identify the noble gas whose electronic configuration is included in that of the concerned element.

b). Write the chemical symbol of the identified noble gas within square brackets. We call this the noble gas core.

c). Add electrons beyond the noble gas core. Note that electrons that are added to the electronic level of the highest principal quantum number (the outermost level or valence shell) are called valence electrons.

Example: Given the electronic configurations of the noble gases Ne and Ar, one can write the electronic configuration of some elements as:

$$Ne(Z=10): 1s^2 2s^2 2p^6 \text{ or } [\text{Ne}]$$

$$Ar(Z=18): 1s^2 2s^2 2p^6 3s^2 3p^6 \textbf{ or [Ar]}$$

1) $Na: 1s^2 2s^2 2p^6 3s^1$ **or** $[Ne]3s^1$

2) $Mg(Z=12): 1s^2 2s^2 2p^6 3s^2$ **or** $[Ne]3s^2$

3) $Ca(Z=20): 1s^2 2s^2 2p^6 3s^2 3p^6 4s^2$ **or** $[Ar]4s^2$

4) $Cr(Z=24): 1s^2 2s^2 2p^6 3s^2 3p^6 3d^5 4s^1$ **or** $\mathbf{[Ar]4s^1 3d^5}$

5) $Cr(Z=24): 1s^2 2s^2 2p^6 3s^2 3p^6 3d^{10} 4s^1$ **or** $\mathbf{[Ar]4s^1 3d^{10}}$

Application activity 16.4.1.d

Using the noble gas notation, write the electronic configuration of the following atoms/ions.

a). Ge (Z=32)
b). S (Z=16)
c). Co^{2+} (Z=27)
d). Br^{-} (Z=35)
e). Sr (Z=38)

16.4.2. Electronic configuration and stability

Activity 16.4.2

State the Hund's rule and explain how the rule is important in the understanding of the behaviour of an element.

Hund's rule states that electrons first occupy the similar energy orbitals that are empty before occupying those which are half-full. This is especially helpful when determining unpaired electrons. The Aufbau process denotes the method of "building up" each sub-shell before moving on to the next;

Almost all the elements follow the same trend for writing electronic configuration.

Sometimes electron configurations of certain elements appear to violate the rules that govern the electron configuration.

For example, the electron configuration of chromium is $[Ar]3d^5 4s^1$ rather than the $[Ar]3d^4 4s^2$ configuration we might have expected.

Similarly, the configuration of copper is $[Ar]3d^{10} 4s^1$ instead of $[Ar]3d^9 4s^2$.

This anomalous behaviour is largely a consequence of the closeness of the $3d$ and 4s orbital energies.

The orbitals in which the sub-shell is exactly half-filled or completely filled are more stable because of the symmetrical distribution of electrons.

Application activity 16.4.2

1. Write the electronic structure of the following chemical species.

   K (Z=19), Ne (Z=10), Al^{3+} (Z=13), Cl (Z=17), O^{2-} (Z=16)

2. a) Write the electronic configuration for each of the following pairs of ions. State the more stable ion in gaseous state and explain your choice.

   i. Cu^{+} and Cu^{2+} (ii) $Fe^{2+}$and Fe^{3+}

   b) Using information in question (a) specify the species from each pair has a more stable electronic configuration. Explain.

16.5. Relationship between ionization energy, energy levels and factors influencing ionization energy

16.5.1. The graphs of ionization energy versus the number of electrons removed

Activity 16.5.1

1. Write the electronic configuration of the following elements/ions, use s, p, d, …) Sodium, magnesium, magnesium ion (Mg^{2+}), aluminium, aluminium ion (Al^{3+}), oxide ion (O^{2-}).
2. Identify the common feature of ions in (1) and why do they have such feature.
3. Suggest what happened to aluminium atom when it changed to aluminium ion (Al^{3+}).
4. Identify the group and the period of aluminium, sodium and oxygen atom.

Concept of Ionization energy

The ionization energy is a measure of the energy needed for an atom, in gaseous state, to lose an electron and become positive ion.

The first ionisation energy is the energy required to remove one electron from an atom in its gaseous state. The example below shows how to represent the successive ionization energies of an atom M.

First ionisation energy: $M(g) \rightarrow M^{+}(g) + e^{-}$

Second ionisation energy and n^{th} ionisation energy: Two or more electrons can be removed and we have successive ionization energies.

$$M^{+}(g) \rightarrow M^{2+}(g) + e^{-}$$

$$M^{(n-1)+}(g) \rightarrow M^{n+}(g) + e^{-}$$

The ionization energy is usually expressed in kilojoules per mole ($kJ.mol^{-1}$). This energy is required to overcome the attractive force between the nucleus and the electron and then remove the electron. Theoretically there are as many successive ionisation energies as there are electrons in the original atom. In figure 2.9, someone can make an interpretation of successive ionization energies of an atom.

- Successive ionisation energies for an element increase, since the remaining electrons are pulled closer to the nucleus and are more tightly held.
- Evidence for the arrangements of electrons in shells of different energies is provided by values of successive ionisation energies for elements.
- The figure below shows a graph of the logarithm of the ionisation energy required for the removal of one electron after another from the potassium atom.
- A logarithmic scale is used in order to give a condensed graph, since there is a large range in values of successive ionisation energies.

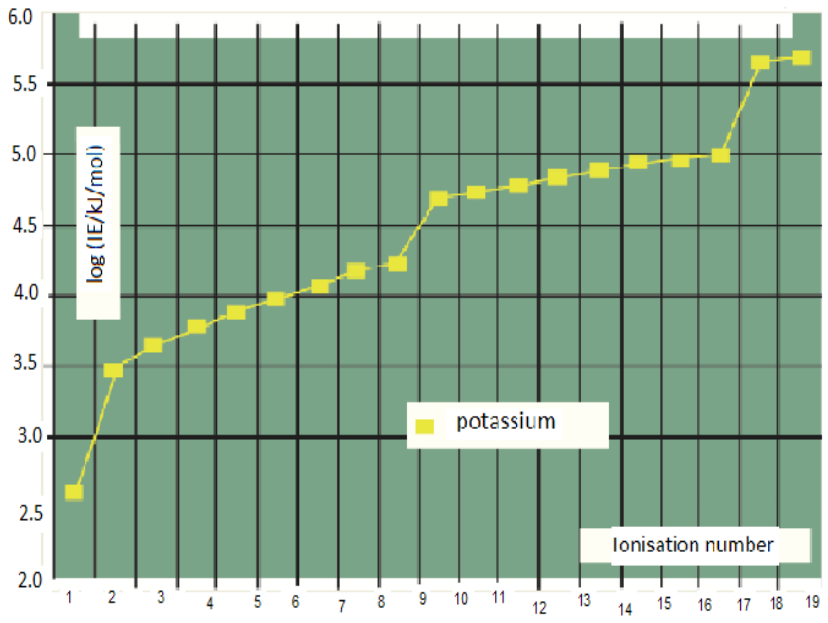


Figure 16.7: Successive ionisation energies of potassium

- The graph that has four parts shows three breaks. These three breaks are evidence of principal energy level.
- The first break occurs after the first electron has been removed from the forth energy level (n = 4). The second breaks after the ninth electron has been removed from the third energy level (n = 3).
- The third break occurs after the 17^{th} electron is removed from a second energy level.
- The two electrons with the highest ionisation energy are close to the nucleus and form the n = 1 shell.

Application activity 16.5.1

1. What is meant by the term first ionisation energy?
2. Explain why helium has the highest first ionisation of all the elements.
3. Explain why the first ionisation energy of sodium is less than that of neon, but the second ionisation energy of sodium is greater than the second ionisation energy of neon.
4. Sketch a graph to show the successive ionisation energies for aluminium. What does the shape of your graph tells you about the electron configuration of aluminium?

16.5.2. Interpretation of a graph of first ionization energy versus the atomic numbers of elements.

Activity 16.5.2

Using the s, p, d, f notation, predict the electronic configurations of the first twenty elements.

Use the established electronic configurations to predict the trends in the first ionisation energy values for the elements above.

The graph below shows the first ionisation energies for the first 20 element

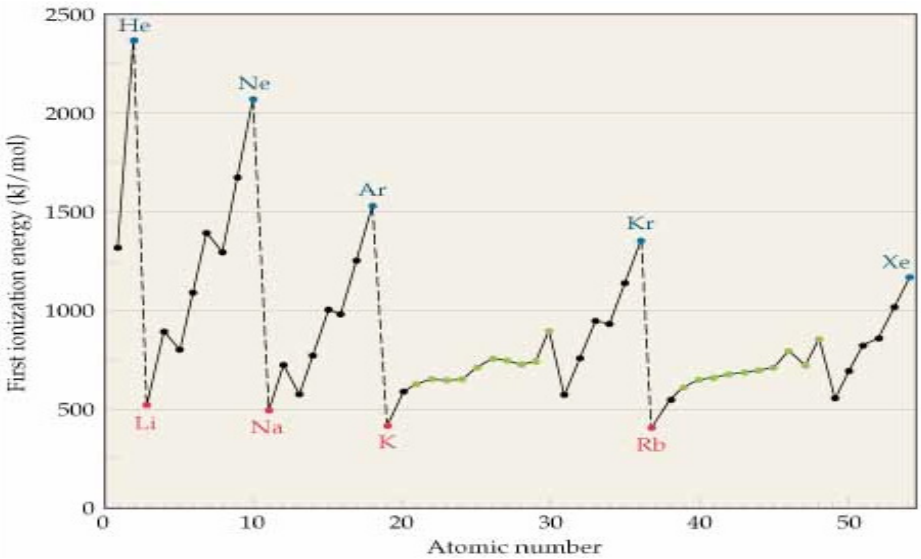


Figure 16.8.Variation of the first ionization energy with atomic number
Source: Chemistry The central science

Note that the noble gases have high ionization energies, whereas the alkali metals and alkaline earth metals have low ionization energies.

Note also that, apart from small irregularities, the first ionization energies of elements in a period increase with increasing atomic number.

Application activity 16.5.2

1. Write an equation to illustrate the process which occurs during the first ionization of neon.
2. Explain why the value of the first ionization energy of magnesium is higher than that of sodium.
3. Write an equation to illustrate the process occurring when the second ionization energy of magnesium is measured.
4. The Ne atom and the Mg^{2+} ion have the same number of electrons. Give two reasons why the first ionization energy of neon is lower than the third ionization energy of magnesium.
5. Explain why the first ionization of aluminum is smaller than that of magnesium greater than the first ionization energy of magnesium.

16.5.3. Factors influencing the magnitude of ionization energy

Activity 16.5.3

1. Why is the first ionisation of neon higher than that of fluorine?
2. Explain why the first ionisation energy of potassium is lower than that of sodium?

The ionization energy is a physical property of elements that can be influenced by some factors:

1. Size of atom

The atomic size is the distance between the nucleus and valence shell. As the number of energy levels (shells) increases, the force of attraction between nucleus and valence electron decreases. Therefore, the valence electrons are loosely held to the nucleus and lower energy is required to remove them, i.e. ionization energy decreases with increase in atomic size and vice versa. This is what happens when you go down a Group.

2. Nuclear charge

The nuclear charge is the total charge of all the protons in the nucleus. As the nuclear charge increases, the force of attraction between nucleus and valence electrons on the same valence energy level increases and hence makes it difficult to remove an electron from the valence shell. The higher the nuclear charge, the higher the ionization energy. This is what happens when you cross a period from left to right.

3. Screening effect or Shielding effect

The Screening effect or Shielding effect is due to the presence of inner electrons which have a screening or shielding effect against the attraction of the nucleus towards the outermost electrons. The electrons present in inner shells between the nucleus and the valence shell reduce the attraction between nucleus and the outermost electrons. This shielding effect increases with the increasing number of inner electrons. A strong Shielding effect makes it easier to remove an external electron and hence lowers the ionization energy.

4. Electronic Configuration

Electronic configuration plays an important role in determining the value of ionization energy. Atoms having stable configuration (i.e. fully filled or half filled) have least tendency to lose electrons and hence have high of ionization energy values.

Application activity 16.5.3

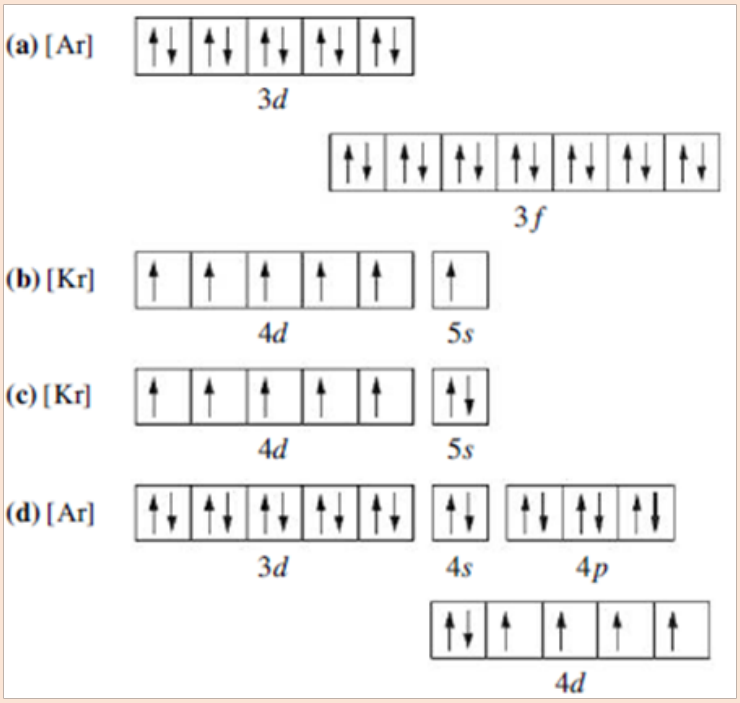
1. Why is the first ionization energy of krypton lower than the first ionization energy of argon?
2. Explain why the value of the first ionization energy of sulphur is lower than that of phosphorus.
3. Explain why the third ionisation energy of magnesium is very much larger than the second ionisation energy of magnesium.

Skills lab 16

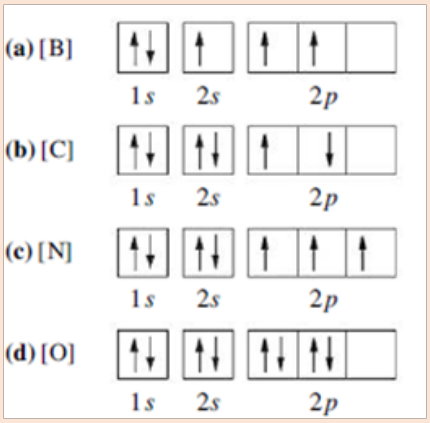
Absorption and emission spectra are like finger prints. Assume that you have to market some goods using advertising lamps that can emit light with at least 10 alternating colours. Write paper specifying the chemical elements that may be involved in the production of these advertising lamps.

End unit assessment 16

1. Which of the following is the correct representation of the ground-state electron configuration of molybdenum? Explain what is wrong with each of the others.



2. Which of the following electron configurations are correct and which ones are wrong? Explain.



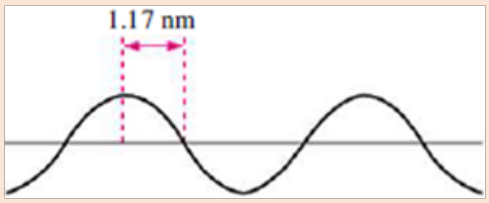
3. Photosynthesis uses 660 nm light to convert CO_2 and H_2O into glucose and O_2. Calculate the frequency of this light.
4. Which of the following orbital designations are incorrect: 1s, 1p, 7d, 9s, 3f, 4f, 2d?
5. The data encoded on CDs, DVDs, and Blu-ray discs is read by lasers. What is the wavelength in nanometres and the energy in joules of the following lasers?

   CD laser, $\nu = 3.85 \times 10^{14}\ Hz$

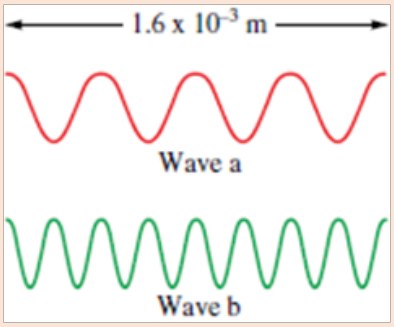
   DVD laser, $\nu = 4.62 \times 10^{14}\ Hz$

   Blu-ray laser, $\nu = 7.41 \times 10^{14}\ Hz$

6. Concerning the concept of energy levels and orbitals,
   a) How many subshells are found in $n=3$?
   b) What are the names of the orbitals in $n=3$?
   c) How many orbitals have the values $n=4$ and $l=3$?
   d) How many orbitals have the values $n=3$, $l=2$ and $m_l = -2$?
   e) What is the total number of orbitals in the level $n=4$?
7. A hypothetical electromagnetic wave is pictured here. What is the wavelength of this radiation?



8. Consider the following waves representing electromagnetic radiation:



a) Which wave has the longer wavelength?

b) Calculate the wavelengths of the two radiations

c) Which wave has the higher frequency and larger photon energy?

9. Order the orbitals for a multielectron atom in each of the following lists according to increasing energy:

a). 5p, 5d

b). 4s, 3d

c). 6s, 4d

10. According to the Aufbau principle, which orbital is filled immediately after each of the following in a multielectron atom?

a) 4s

b) 3d

c) 5f

d) 5p

11. According to the Aufbau principle, which orbital is filled immediately before each of the following?

a) 3p

b) 4p

c) 4f

d) 5d

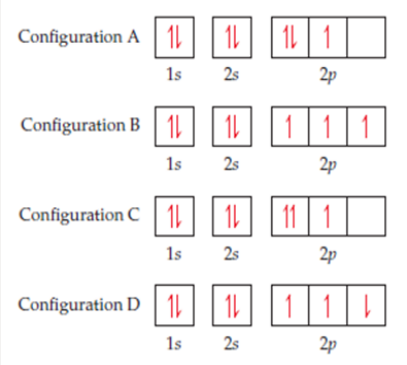
12. The ground-state electron configurations listed here are incorrect. Explain what mistakes have been made in each and write the correct electron configurations.

Al: $1s^2 2s^2 2p^4 3s^2 3p^3$

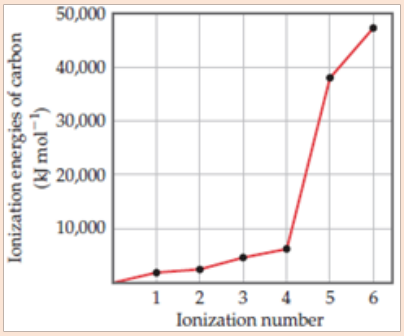
B: $1s^2 2s^2 2p^5$

F: $1s^2 2s^2 2p^6$

13. Four possible electron configurations for a nitrogen atom are shown below, but only one represents the correct configuration for a nitrogen atom in its ground state. Which one is the correct electron configuration? Which configurations violate the Pauli Exclusion Principle? Which configurations violate Hund's rule?



14. Explain the variation in the ionization energies of carbon, as displayed in this graph.



15. The first seven ionization energies of an element W are shown below

1^{st}I.E	2^{nd}I.E	3^{rd}I.E	4^{th}I.E	5^{th}I.E	6^{th}I.E	7^{th}I.E
785	1581	3231	4361	16002	20001	23602

What factors determine the magnitude of the first ionization energy

16. The following table shows the ionisation energies (in kJ mol-1) of the five elements lettered A, B, C, D and E.

Element	1^{st} IE	2^{nd} IE	3^{rd} IE	4^{th} IE
A	500	4600	6900	9500
B	740	1500	7700	10500
C	630	1600	3000	4800
D	900	1800	14800	21000
E	580	1800	2700	11600

a) Which of these elements is most likely to form an ion with a charge of 1+? Give reasons for your answer.

b) Which two of the elements are in the same group of the periodic table? Which group do they belong to?

c) In which group of the periodic table is element E likely to occur? Give reasons for your answer.

d) Which element would require the least energy to convert one mole of gaseous atoms into ions carrying two positive charges?

UNIT 17:

IONIC AND METALLIC BONDS

Key Unit competence: Describe how properties of ionic compounds and metals are related to the nature of their bonding.

Introductory Activity

Consider the following figures and answer the related questions

Figure 17.1

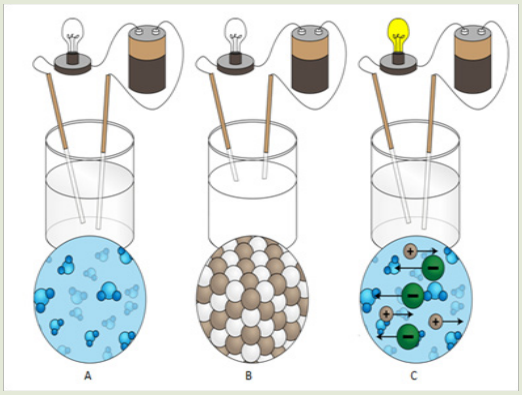


Figure17.2

1. Figure 1 shows materials commonly used at home. If you reflect back around your house/home you will see hundreds of objects made from different kinds of materials.
   a) Observe the objects (in picture) and classify them according to the materials they are made of.
   b) Have you ever wondered why the manufacturers choose the material they did for each item?
   c) Why are frying saucepans made of metals and dishes, cups and plates often made of glass and ceramic?
   d) Could dishes be made of metal? And saucepans made of ceramic and glass?

2. Figure 17.1 (b) shows the electric conductivity of distilled water, solid table salt and a solution of a table salt respectively.

   a) Use the diagrams A, B and C to explain the observations from the set up.
      i. No light is given out by bulb in A
      ii. No light is given out by bulb in B
      iii. Light is given out in C

   b) Suppose that you have a 30 cm bar made of table salt. Suggest the change, if there is any, that can occur and deduce the property related, when this salt bar is:
      i. Dropped from a table of 1 m high to the floor
      ii. Immersed in water found in a water bath.
      iii. Dry heated to 100^{o}C

17.1. Explanations of why atoms of elements form bonds

Activity 17.1

Consider Chlorine (Cl, Z = 17) and Argon (Ar, 18) atoms of the elements of Period 3 in the Periodic Table.

d) Which of these atoms is more reactive?

e) Suggest the reasons for your answer in (**a)** in terms of the electronegativity and electronic structure.

f) Choose, between Chlorine and Argon, which one has lower energy potential.

The atoms of most elements form chemical bonds because the atoms become more stable when bonded together. Electric forces attract neighboring atoms to each other, making them stick together.

In atoms, electrons are arranged into complex layers called shells. For most atoms, the outermost shell is incomplete, and the atom shares electrons with other atoms to fill the shell.

The type of chemical bond maximizes the stability of the atoms that form it.

An ionic bond, where one atom essentially donates an electron to another, forms when one atom becomes stable by losing its outer electrons and the other atoms become stable (usually by filling its valence shell) by gaining the electrons.

*Covalent bonds form when **sharing atoms** results in the highest stability.*

Other types of bonds besides ionic and covalent chemical bonds exist, too.

*Atoms with **incomplete shells** are said to have **high potential energy**; atoms whose **outer shells arefull** have **low potential energy**. In nature, objects with high potential energy "seek" a lower energy, becoming more stable as a result. Atoms form chemical bonds to achieve lower potential energy.*

Application activity 17.1

Explain why atoms of elements form bonds

17.2. Gain of stability by losing and gaining electrons

Activity 17.2

Observe the electronic configuration of the following atoms in groups and discuss the following questions.

Na: $1s^2 2s^2 2p^6 3s^1$

Cl: $1s^2 2s^2 2p^6 3s^2 3p^5$

Ne:$1s^2 2s^2 2p^6$

a) How many electrons does sodium have on its outer most shell?
   i. How many electrons does Sodium need to be stable?
   ii. What is the easiest way for Sodium to be stable?

b) How many electrons does Chlorine have on its outer most shell?
   i. How many electrons does Chlorine need missing to get stability?
   ii. What is the easiest for Chlorine to be stable?

c) Does Neon need more electrons to be stable? If Yes why? If no why?

Like people always relate and connect to others depending on their values, interests and goals so does unstable atoms. They combine together to achieve stability. We know that noble gases are the most stable elements in the periodic table. They have a filled outer electron energy level.

When an atom loses, gains, or shares electrons through bonding to achieve a filled outer electron energy level, the resulting compound is often more stable than individual separate atoms.

- Neutral sodium has one valence electron. When it loses this electron to chlorine, the resulting Na+ cation has an outermost electron energy level that contains eight electrons. It is isoelectronic (same electronic configuration) with the noble gas neon.
- On the other hand, *chlorine* has an outer electron energy level that contains *seven electrons*. When chlorine *gains* sodium's electron, it becomes an anion that is isoelectronic with the *noble gas argon*. The fact that atoms need to form bonds with other atoms by loosing or gaining electrons to attain stability is called the octet rule; this means to have 8 electrons at the outermost shell.

It is easiest to apply the "Octet Rule" to predict whether two atoms will form bonds and how many bonds they will form. Most atoms need 8 electrons to complete their outer shell. So, an atom that has 2 outer electrons will often form a chemical bond with an atom that lacks two electrons to be "complete". The octet rule states that elements gain or lose electrons to attain an electron configuration of the nearest noble gas. Octet comes from Latin language meaning "eight".

Note that the "Duet Rule" is also applied. The noble gas HELIUM has two electrons (a doublet) in its outer shell, which is very stable. Hydrogen only needs one additional electron to attain this stable configuration, while lithium needs to lose one.

Low atomic weight elements (the first twenty elements) are most likely to adhere to the Octet Rule. For example,

- A sodium atom has one lone electron in its outer shell.
- A chlorine atom, in contrast, is short one electron to fill its outer shell.
- Sodium readily donates (looses) its outer electron (forming the Na+ ion, since it then has one more proton than it has electrons), while chlorine readily accepts (gains) a donated electrons(making the Cl- ion, since chlorine is stable when it has one more electron than it has protons).
- Sodium and chlorine form an ionic bond with each other, to form table salt or sodium chloride.

Application activity 17.2

d) State the following Rule?

i. Octet Rule

ii. Duet Rule

b) Answer to the following questions

i. Does sodium need to gain electron than chlorine? (Yes or No)

ii. Explain the target of sodium when it is seeking to lose electron and chlorine to gain electron.

c) Which of the following is stable? Explain why?

i. Na^+ ii. Na iii.Cl iv.Cl^-

17.3. Ionic bonding

Activity 17.3

In the Ordinary Level, you learnt that there exist three main types of chemical bonding namely, covalent, ionic and metallic.

b) Recall the definition of the ionic bond.

c) State the properties of a table salt and use it to generalize the properties of ionic compounds. (Appearance, Solubility in water and in petrol, Temperature required to melt, Electrical conductivity of solid and aqueous solution)

17.3.1. Concept of ionic bonding

The ionic bond is formed by ***complete transfer of electrons from one to the other, being a metal atom and the other a non-metal atom.***

Due to the atoms are neutral:

- When an atom gives up an electron, it is positively charged, forming what is called ***a positive ion*** or **cation**. The positive charge of a monovalent cation is equal in magnitude but opposite to that of the electron ($1,602 \times 10^{-19}$ C) sign.
- If an atom captures one electron, it will be negatively charged, thereby forming ***a negative ion*** or ***anion***. The negative charge of a monovalent anion is therefore the same as the electron.

When two counter-ions have been formed, i.e. a cation and an anion, they attract each other through **electrostatic forces**, and so, they can form a stable molecule.

These ***electrostatic attractive forces*** (sometimes called **Coulomb forces)** are therefore responsible for the formation of ionic compounds.

Suppose the simple case of *sodium chloride* or *common salt*. The sodium atom (Z = 11) has a ***single (unpaired) electron*** in orbital **3s**, somewhat isolated from other pairs; while in the chlorine atom, there is also ***a single (unpaired) electron***, but in this case it is in the $\mathbf{3p_z}$ orbital, other **3s**, $\mathbf{3p_x}$ and $\mathbf{3p_y}$ orbitals are being inhabited by respective pairs of electrons.

The alone electron passes from sodium to chlorine; which, besides forming a pair with the electron in $\mathbf{3p_z}$ orbital, it will be found surrounded by other couples.

All the atoms of the alkali (group 1) metals have an external electronic configuration type $\mathbf{ns^1}$, i.e., with a single electron in the outermost orbit. This electron, which is often called **valence electron** is quite far from the core (the nucleus), which is separated also by the other electrons, called **internal (electrons)**, which largely core-shielding attraction on said valence electron.

So, it is quite easy to remove this electron, for which a little energy is spent. This is why the ionization energy, which, for the alkali metals, is very small.

When an alkali metal atom has been easily removed its valence electron, however, it is difficult to remove a second electron, as their ionization energy is very high. Therefore, the alkali metal cations are relatively easily *monovalent* ($\mathbf{M^+}$).

For the group 13 elements (B, Al, Ga, In, T*l*), it costs much energy to remove valence electrons, which, in this case, are the type $\mathbf{ns^2np^1}$, so it is difficult. Hence, they form ***trivalent cations***, except heavier atom of T*l* that can form in certain cases, ***monovalent cation***.

The elements of groups 14, 15, 16 and 17 of the Periodic Table, as well as the noble gases (group 18), the ionization energy is increasing, so it is very difficult for these elements to form positive ions.

- Only the heaviest elements of group 14, tin and lead, form, in some cases, ***divalent cations*** (loss of two of the valence electrons $\mathbf{np^2}$). The rest of the elements of these groups form covalent bonds or negative ions.
- The atoms of group 17 elements of the Periodic Table (halogens) have an outer electron configuration $\mathbf{ns^2np^5}$, that is, they lack one electron to complete the **p** orbitals and thus to form the electron configuration

of noble gas that follows in the same period. Therefore, it is easy to understand that if one of these atoms is joined by a new electron; a more stable configuration is obtained, shedding energy in this process. This energy is called electron affinity, which for the halogens is high. These elements will form in a relatively easily way, ***monovalent anions***.

- These anions have no tendency to take a second electron, so it would have to stand, alone in an outermost orbit, without the nucleus exercised about it any attractive force. Halogens form only ***monovalent anions***.
- Atoms of oxygen family elements (group 16 of the Periodic Table) are missing two electrons to complete the external orbitals np and acquire the noble gas configuration. Therefore, these elements tend to form ***divalent ions***, although in this process the energy balance is slightly negative.
- Nitrogen family elements (group 15 of the periodic table are hardly ***trivalent anion***; while carbon group (group 14) is almost impossible the form ***tetravalent anions***. Therefore, the compounds of the nitrogen family are largely covalent part and those of carbon family are typically covalent.

17.3.2. Ionic bond formation

Once the oppositely charged ions form when electrons are transferred from one atom to another, they ***are attracted by their positive and negative charges*** (by electrostatic forces) and form an ionic compound. Ionic bonds are also formed when there is a large electronegativity difference between two atoms. This difference causes an unequal sharing of electrons such that ***one atom completely loses one or more electrons and the other atom gains one or more electrons.*** For example, in the creation of an ionic bond between a metal atom, sodium (electronegativity = 0.93) and a non-metal, fluorine (electronegativity = 3.98).

Let us take a look of how ***sodium and fluorine bond to form sodium fluoride***.

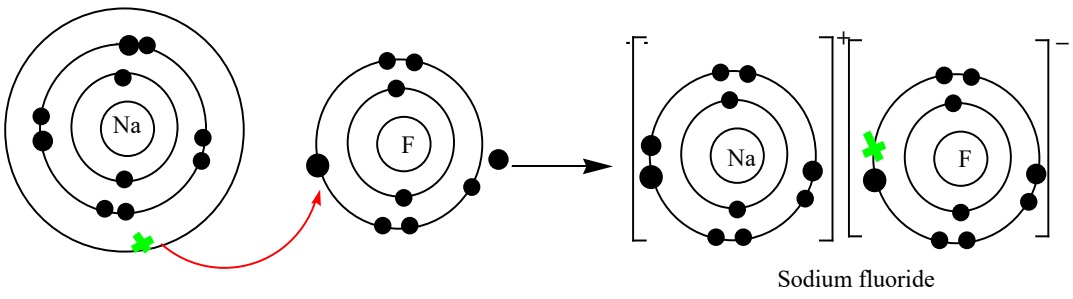


Figure 17.2: Formation of NaF

The *curved arrow* between sodium and fluorine atoms represents the transfer of an electron from a sodium atom to a fluorine atom to form oppositely charged ions. These two ions ***are strongly attracted to each other because of their opposite charges***. A bond is now formed and the resulting compound is called **Sodium fluoride**

Another example of ionic bonding formation is the ***formation of magnesium oxide***.

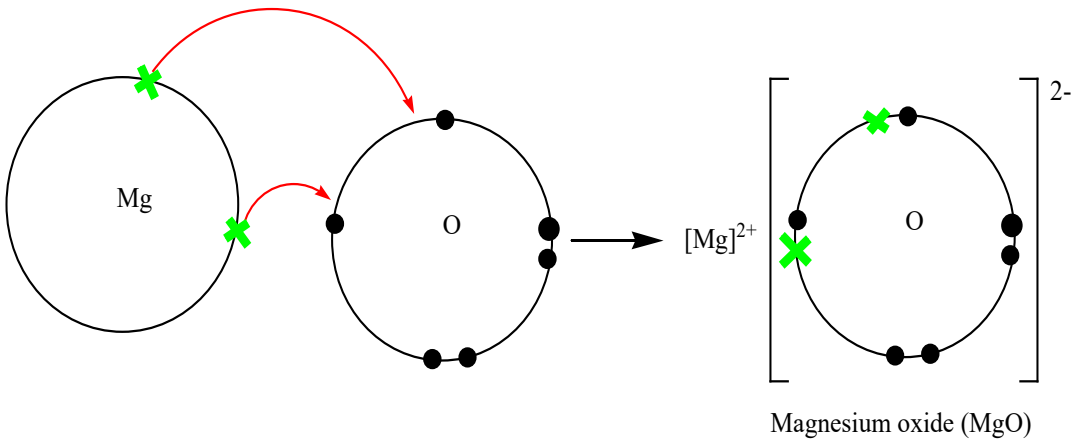


Figure 17.3: Formation of MgO

17.3.3. Properties of ionic compounds

Here are the properties shared by the **ionic compounds**. Notice that the properties of ionic compounds relate to ***how strongly the positive and negative ions attract each other*** in an ionic bond.

1. They have high melting points and high boiling points

In an ionic lattice, there are ***many strong electrostatic attractions between oppositely charged ions.*** We therefore expect that ionic solids will have high melting points. On melting although the regular lattice is broken down, there will still be significant attractions between the ions in the liquid. This should result in high boiling points also.

The factors which affect the melting point of an ionic compound are:

- **The charge on the ions of ionic compound:** "The greater the charge, the greater the electrostatic attraction, the stronger the ionic bond, the higher the melting point."

  For example, Melting Point of **NaCl** is **801 °C** and that of **MgO** is **2,800 °C**.

- **The size of the ions of ionic compound**: "Smaller ions can pack closer together than larger ions so the electrostatic attraction is greater, the ionic bond is stronger, the melting point is higher." For example, Melting Point of **NaF** is **992 °C** and that of **CsF** is **2,800 °C**.

2. Most ionic compounds are soluble in water

This is because the electrostatic forces of the polar water molecules are stronger than the electrostatic forces keeping the ions together. When an ionic compound like NaCl is added to water, water molecules attract the positive and negative salt ions. Water molecules surround each ion and move the ions apart from each other. The separated ions dissolve in water. There are several exceptions, however, where the electrostatic forces between the ions in an ionic compound are strong enough that the water molecules cannot separate them. Despite these few limitations, water's ability to dissolve ionic compounds is one of the major reasons it is so vital to life on Earth. Ionic compounds are generally insoluble in non-polar solvents like kerosene.

3. They are hard and brittle

Ionic crystals are hard because the positive and negative ions are strongly attracted to each other and difficult to separate, however, ionic solids are brittle.

When a stress is applied to the ionic lattice, the layers shift slightly. The layers are arranged so that each cation is surrounded by anions in the lattice. If the layers shift then ions of the same charge will be brought closer together. Ions of the same charge will repel each other, so the lattice structure breaks down into smaller pieces.

4. They conduct electricity when molten or dissolved in water

In order for a substance to conduct electricity, it must contain mobile particles capable of carrying charge.

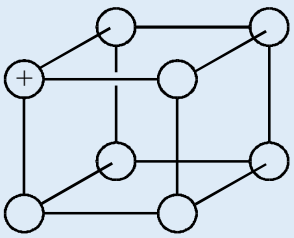
- Solid ionic compounds do not conduct electricity because the ions (charged particles) are locked into a rigid lattice or array. The ions cannot move out of the lattice, so the solid cannot conduct electricity.
- When is molten, the ions are free to move out of the lattice structure.
  - Cations (positive ions) move towards the negative electrode (cathode): $M^+ + e^- \rightarrow M$
  - Anions (negative ions) move towards the positive electrode (anode): $X^- \rightarrow X + e^-$
- When the ionic compound is **dissolved in water** to form an aqueous solution, the ions are released from the lattice structure and **are free to move** so the solution conducts electricity just like the molten (liquid) ionic compound.

5. They form crystals

Ionic compounds form crystal lattices rather than amorphous solids. Although molecular compounds form crystals, they frequently take other forms but molecular crystals typically are softer than ionic crystals. At an atomic level, an ionic crystal is a regular structure, with the cation and anion alternating with each other and forming a three-dimensional structure based largely on the smaller ion evenly filling in the gaps between the larger ion.

Application activity 17.3

1. The diagram below represents a part of the structure of sodium chloride. The ionic charge is shown on the centre of only one of the ions.



   a) On the diagram, mark the charges on the four negative ions.

   b) What change occurs to the motion of the ions in sodium chloride when it is heated from room temperature to a temperature below its melting point?

   c) Sodium chloride can be formed by reacting sodium with chlorine. A chloride ion has one more electron than a chlorine atom. In the formation of sodium chloride, from where does this electron come?

2. Draw diagrams to illustrate the formation of ionic compounds in the following substances:
   a) Calcium chloride
   b) Sodium peroxide
   c) Iron (III) chloride
   d) Sodium sulphide

3. Solid sodium chloride and solid magnesium oxide are both held together by ionic (electrovalent) bonds.

   a) Using s, p and d notation write down the symbol for and the electronic configuration of (i) a sodium ion; (ii) a chloride ion; (iii) a magnesium ion; (iv) an oxide ion.

b) Explain what holds sodium and chloride ions together in the solid crystal

c) Sodium chloride melts at 1074 K; magnesium oxide melts at 3125 K. Both have identical structures. Why is there such a difference in their melting points?

17.4. Metallic bonding

Activity 17.4

1. Give three examples of substances which are malleable, ductile, good conductor of heat and electricity, and having a characteristic luster. Here you can use a dictionary or other searching tools to find the meaning for any unfamiliar word.
2. Suggest another property, apart from those given, of the substances you have given in **(1)**.
3. Choose from the examples given in **(1)**, one which is most common and well known.

   a) This substance is seen to be composed by atoms of one element. Which one?

   b) Use a labeled drawing to show the internal structure of that kind of substance.

A metallic bond is a type of chemical bond formed between positively charged atoms in which the free electrons are shared among a lattice of cations. In contrast, covalent and ionic bonds form between two discrete (separate) atoms.

Metallic bonding is the main type of chemical bonds that forms between metal atoms (pure metals and alloys and some metalloids). A metal is a lattice of positive metal 'ions' in a 'sea' of delocalised electrons.

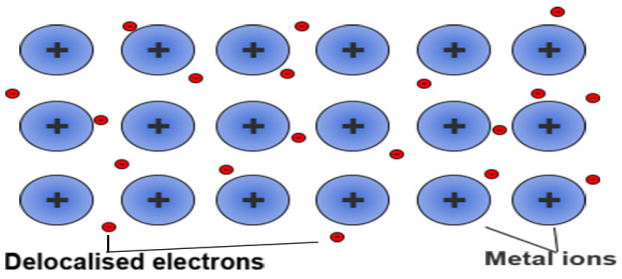


Figure 17.4: Electron Sea Model of Metallic bonding

17.4.1. How metallic bonds work

The outer energy levels of metal atoms (the s and p orbitals) overlap. At least one of the valence electrons participating in a metallic bond is not shared with a neighbor atom, nor is it lost to form an ion. Instead, the electrons form what may be termed an “electron sea” in which valence electrons are free to move from one atom to another. Metallic bonding refers to the interaction between the delocalised electrons and the metal nuclei.

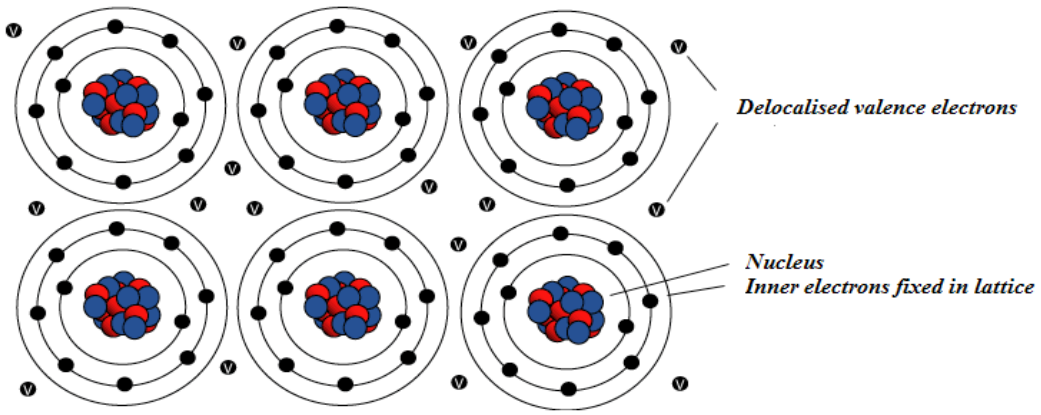


Figure 17.5:Electron Sea Model of Metallic bonding

The electron sea model is an oversimplification of metallic bonding. Calculations based on electronic band structure or density functions are more accurate. Metallic bonding may be seen as a consequence of a material having many more delocalized energy states than it has delocalized electrons (electron deficiency), so localized unpaired electrons may become delocalized and mobile. The electrons can change energy states and move throughout a lattice in any direction.

Bonding can also take the form of metallic cluster formation, in which delocalized electrons flow around localized cores. Bond formation depends heavily on conditions. For example, hydrogen is a metal under high pressure! As pressure is reduced, bonding changes from metallic to non-polar covalent.

17.4.2. Physical properties of metals

Because electrons are delocalized around positively-charged nuclei, metallic bonding explains many properties of metals.

1. Electrical Conductivity

Most metals are *excellent electrical conductors* because the *electrons in the electron sea are free to move and carry charge*. For example, electric wires in our homes are made of aluminium and copper. They are good conductor

of electricity. Electricity flows most easily through gold, silver, copper and aluminium. Gold and silver are used for fine electrical contacts in computers.

2. Thermal Conductivity

Metals conduct heatbecause the free electrons are able to transfer energy away from the heat source and because vibrations of atoms (phonons) move through a solid metal as a wave. Cooking utensils and water boilers are also made of iron, copper and aluminium, because they are good conductors of heat.

3. Ductility

Metalstend to be ductile or able to be drawn into thin wires because local bonds between atoms can be easily broken and also reformed. Single atoms or entire sheets of them can slide past each other and reform bonds. Wires are mainly made from copper, aluminium, iron and magnesium.

4. Malleability

Metals are often malleable or capable of being molded or pounded into a shape, again because bonds between atoms readily break and reform. This ability to bend or be shaped without breaking occurs because the electrons simply slide over each other instead of separating. The binding force between metals is non-directional, so drawing or shaping a metal is less likely to fracture it. Electrons in a crystal may be replaced by others. Gold and Silver metals are the most malleable metals. They can be hammered into very fine sheets. Thin aluminium foils are widely used for safe wrapping of medicines, chocolates and food material.

5. Metallic Luster

Metals tend to be shiny or display metallic luster. They are opaque once a certain minimum thickness is achieved.

The electron sea reflects photons off the smooth surface therefore there is an upper frequency limit to the light that can be reflected. Silver is a very good reflector. It reflects about 90% of the light falling on it. All modern mirrors contain a thin coating of metals. Due to their shiny appearance they can be used in jewellery and decorations.

Application activity 17.4

1. Magnesium has a higher melting and boiling point than sodium. This can be explained in terms of the electronic structures, the packing, and the atomic radii of the two elements.

a) Explain why each of these three things causes the magnesium melting and boiling points to be higher.

b) Explain why metals are good conductors of electricity.

c) Explain why metals are also good conductors of heat.

2. Pure metals are usually malleable and ductile.

a) Explain what those two words mean.

b) If a metal is subjected to a small stress, it will return to its original shape when the stress is removed. However, when it is subjected to a larger stress, it may change shape permanently. Explain, with the help of simple diagrams why there is a different result depending on the size of the stress.

c) When a piece of metal is worked by a blacksmith, it is heated to a high temperature in a furnace to make it easier to shape. After working it with a hammer, it needs to be re-heated because it becomes too difficult to work. Explain what is going on in terms of the structure of the metal.

d) Why is brass harder than either of its component metals, copper and zinc?

Skills lab 17

Experiment to demonstrate the malleability and ductility of metals

Materials: Wires, nails, hammer, piece of cloth.

Procedure:

1. Wrap the material to be test in a heavy plastic or cloth to avoid pieces flying from the material.
2. Place the material on a flat hard surface
3. Use a harmer to pound the material flat
4. Record your observations as malleable or non-malleable.

End unit assessment 17

1. State whether the following statement is True or False. Justify your answer. “Sodium Chloride has a higher melting point than Magnesium Oxide”.
2. Why are ionic compounds brittle?
3. Why do ionic compounds have high melting points?
4. What happens when an electric current is passed through a solution of an ionic compound?
5. This question is about metallic bonding.
   a) Describe the bonding that is present in metals.
   b) Explain how the bonding and structure lead to the typical metallic properties of electrical conductivity and malleability.
   c) Suggest a reason why aluminium is a better conductor of electricity than magnesium.
6. Silver and sodium chloride melt at similar temperatures. Give two physical properties of silver which are different from those of sodium chloride and, in each case, give one reason why the property of silver is different from that of sodium chloride.
7. This question is about calcium oxide (CaO).
   a) Describe the nature and strength of the bonding in solid calcium oxide.
   b) Use the kinetic theory to describe the changes that take place as calcium oxide is heated from 25°C to a temperature above its melting point.
   c) State two properties of calcium oxide that depend on its bonding.

UNIT 18:

VARIATION IN TRENDS OF THE PHYSICAL PROPERTIES

Key Unit competence: Use atomic structure and electronic configuration to explain the trends in the physical properties of the elements.

Introductory Activity 18

An extended family is a family that extends beyond the nuclear family, consisting of parents like father, mother, their children, aunts, uncles, grandparents and cousins. They all live in the same household. It is easily to see characteristics that qualify as physical traits include hair color and type, attached earlobes, adorable freckles, eye color and hairline among family members.

Think about this activity in order to understand well and answer the following questions.

1. Have you ever noticed physical similarities between relatives at a family reunion?
2. What do physical characteristics indicate among the family members?
3. What are the physical characteristics that may be possessed by the same family members?

18.1. Historical background of the Periodic Table

Activity 18.1

Use textbooks and internet connection and point out the main steps in the development of the periodic table.

Scientists have always been searching for patterns and similarities in the properties and reactions of substances.

For a long time chemists look at ways of dividing up the two big group, metallic and non-metallic elements into smaller sub-groups. They grouped together very reactive metals like lithium, sodium and potassium, slightly

less reactive metals such as calcium, strontium and barium and very reactive non-metals like chlorine, bromine and iodine.

Some elements were discovered which had properties in between metallic and non-metallic. These elements were described as metalloids (e.g. silicon).

- In 1669, Hennig Brand a German merchant and amateur alchemist invented the Philosopher's Stone; an object that supposedly could turn metals into pure gold. He heated residues from boiled urine, and a liquid dropped out and burst into flames. He also discovered phosphorus.
- In 1680 Robert Boyle also discovered phosphorus without knowing about Henning Brand' discovery.
- In 1809, curiously 47 elements were discovered and named, and scientists began to design their atomic structures based on their characteristics.
- In 1829, Johann Dobereiner classified some elements into groups of three, which he called triads.
  The elements in a triad had similar chemical properties and orderly physical properties.
- Examples of triads : Cl, Br, I and Ca, Sr, Ba
- Dobereiner showed that when the three elements in each triad were written in order of atomic mass, the middle element had properties in between those of the other two.
- In 1863, John Newlands (1838 - 1898) suggested that elements be arranged in "octaves" because he noticed (after arranging the elements in order of increasing atomic mass) that certain properties repeated every 8th element. (Law of Octaves).His law of octaves failed beyond the element calcium.
- In 1869, Dimitri Mendeleev based on John Newlands' ideas started the development of elements organized into the periodic table. The arrangement of chemical elements were done by using atomic mass as the key characteristic to decide where each element belonged in his table. The elements were arranged in rows and columns. He predicted the discovery of other elements, and left spaces open in his periodic table for them.
- At the same time, Lothar Meyer published his own periodic table with elements organized by increasing atomic mass.
- In 1886, French physicist Antoine Becquerel first discovered radioactivity. During the same period of 1886, Ernest Rutherford named three types of radiation; alpha, beta and gamma rays.

- In 1886, Marie and Pierre Curie started working on the radioactivity and they discovered radium and polonium. They discovered that beta particles were negatively charged.
- In 1895, Lord Rayleigh discovered a new gaseous element named argon which proved to be chemically inert. This element did not fit any of the known periodic groups.
- In 1898, William Ramsay suggested that argon be placed into the periodic table between chlorine and potassium in a family with helium, despite the fact that argon's atomic weight was greater than that of potassium. This group was termed the "zero" group due to the zero valency of the elements. Ramsey accurately predicted the future discovery and properties neon.
- In 1913, Henry Moseley worked on X-rays and determined the actual nuclear charge (atomic number) of the elements. He has rearranged the elements in order of increasing atomic number.
- In 1897 English physicist J. J. Thomson discovered small negatively charged particles in an atom and named them as electrons; John Sealy Townsend and Robert A. Millikan investigated the electrons and determined their exact charge and mass.
- In 1900, Antoine Becquerel discovered that electrons and beta particles as identified by the Curies are the same thing.
- In 1903, Ernest Rutherford proclaimed that radioactivity is initiated by the atoms which are broken down.
- In 1911, Ernest Rutherford and Hans Geiger discovered that electrons are moving around the nucleus of an atom.
- In 1913, Niels Bohr suggested that electrons move around a nucleus in discreete energy levels called orbits. He observed also that light is emitted or absorbed when electrons transit from one orbit to another.
- In 1914, Rutherford identified protons in the atomic nucleus. He also transformed a nitrogen atom into an oxygen atom for the first time. English physicist Henry Moseley provided atomic numbers, based on the number of electrons in an atom, rather than based on atomic mass.
- In 1932 James Chadwick discovered neutrons, and isotopes were identified. This was the complete basis for the periodic table. In that same year Englishman Cockroft and the Irishman Walton first split an atom by bombarding lithium in a particle accelerator, changing it to two helium nuclei.The last major changes to the periodic table give rise from Glenn Seaborg's work in the middle of the 20th Century. In 1940,

he discovered plutonium and all the transuranic elements from 94 to 102.

- In 1944, Glenn T. Seaborg discovered 10 new elements and moved out 14 elements of the main body of the periodic table to their current location below the lanthanide series. These elements were known as Actinides series.
- In 1951, Seaborg was awarded the Nobel Prize in chemistry for his work. Element 106 has been named seaborgium (Sg) in his honor.
- Presently, 118 elements are in the modern Periodic Table.

Although Dimitri Mendeleev is often considered as the "father" of the periodic table, however the work of many scientists contributed to its present form. The representation of a modern Periodic Table of Elements is shown below.

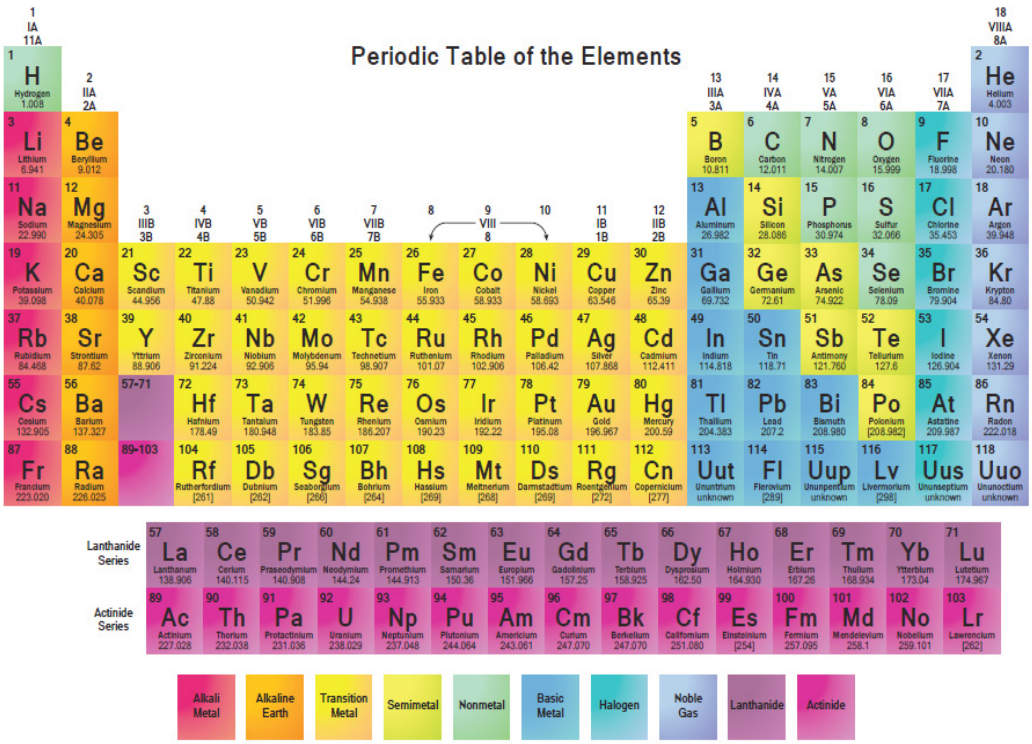


Figure 18.1: The periodic table of elements

Source:**https://0.tqn.com/z/g/chemistry/od/homeschoolchemistry/l/PeriodicTable.pdf**

The elements also belong to the families (chemical families). These elements have similar physical and chemical properties exist among the families and the periods of the periodic table.

Application activity 18.1

The periodic table is an important tool used in chemistry:

1. Why scientists have developed a classification of chemical elements?
2. Explain why the elements are classified in groups and periods of the periodic table
3. Choose one element of Group 1 and one of group 17 and make their electronic configurations using orbitals.

18.2. Comparison of Mendeleev's table with the modern Periodic Table

Activity 18.2

1. Look at the modern periodic table and write down four things it tells you.
2. Explain the gaps found in the Mendeleev periodic table compared to the modern one?
3. How many elements does the modern periodic table contain?

The periodic table is the arrangement of chemical elements according to their chemical and physical properties. The modern periodic table was created after a series of different versions of the periodic table. The Russian Chemist/ Professor Dmitri Mendeleev was the first to come up with a structure for the periodic table with columns and rows. This feature is the main building block for the modern periodic table as well. The columns in the periodic table are called groups or families, and they group together elements with similar properties. The rows in the periodic table are called periods, and they represent sets of elements that get repeated due the possession of similar properties. The main difference between Mendeleev and Modern Periodic Table are shown in the Table 18.1.

Table 18.1. Differences between Mendeleev' table and the modern Periodic Table

CRITERIA	MENDELEEV PERIODIC TABLE	MODERN PERIODIC TABLE
Definition	Was created on the basis of periodic functions of the elements, leaving room for future findings of the missing elements at that time.	Is the one used at the moment, as a collective improvement of the works of so many chemists and scientists in an effort to order the chemical elements to resemble the similarities in their properties.
Ordering	Orders the elements based on their atomic weight.	Orders the elements based on their atomic number.
Gaps	Gaps for the missing elements at that time.	Has no concept as such.
Number of Columns and Rows	Has 8 vertical columns called ***groups*** and 12 horizontal rows called ***periods***.	Has 18 columns called groups and 7 rows called periods.
Existence of Isotopes	Does not support the fact of the existence of isotopes.	Supports this fact as the classification is based upon the atomic number, rather than the atomic weight of the element.
Defining Atomic Structure	Does not support the concept of the atomic structure.	Supports this fact by grouping the elements in such a manner that their electronic configuration can be deduced easily.
Characteristics of Grouped Elements	Has elements with dissimilar properties in the same group sometimes.	Elements have similar properties repeated at regular intervals.

Application activity 18.2

1. Discuss the similarities and differences of Mendeleev's table and modern periodic Table.
2. How were the positions of cobalt and nickel resolved in the modern periodic table?

18.3. Location of the elements in the Periodic Table based on the electronic configuration

Activity 18.3

1. Based on knowledge gained in the previous years:
   a) Represent the electronic configuration of the elements ^{25}X and ^{11}Y.
   b) Discuss the information given by the number of electrons in the last orbitals of the above element about their position in the periodic table?
   c) Explain the period and the group of the periodic table in which the above elements are located.
2. Is it possible to have an element with atomic number 1.5 between hydrogen and helium?

18.3.1. Major Divisions of the Periodic Table

The periodic table is a tabular of the chemical elements organized on the basis of their atomic numbers, electron configurations, and chemical properties.

In the periodic table, the elements are organized by periods and groups. The period relates to the principal energy level which is being filled by electrons. Elements with the same number of valence electrons are put in the same group, such as the halogens and the noble gases. The chemical properties of an atom relate directly to the number of valence electrons, and the periodic table is a road map among those properties such that chemical properties can be deduced by the position of an element on the table. The electrons in the outermost or valence shell are especially important because they participate in forming chemical bonds.

18.3.2. Location of elements in modern Periodic Table using examples

In the periodic table, the elements are located based on groups and periods.

a. Finding Period of Elements

Period of the element is equal to highest energy level of electrons or principal quantum number. Look at following examples for better understanding;

Examples

$_{16}$**S:** $1s^2 2s^2 2p^6 3s^2 3p^4$, 3 is the highest energy level of electrons or principal quantum number. Thusperiod of S is 3.

$_{23}$**Cr:** $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^4$, 4 is the highest energy level of electrons or principal quantum number. Thusperiod of Cr is 4.

b. Finding Group of Elements

Group of element is equal to number of valence electrons of element or number of electrons in the highest energy level of elements.

Examples

$_{19}$**K:** $1s^2 2s^2 2p^6 3s^2 3p^6 4s^1$, 1 is the number of valence electrons of Potassium. Thus group of K is 1.

$_{16}$**S:** $1s^2 2s^2 2p^6 3s^2 3p^4$, 2+4=6 is the number of valence electrons of Sulphur. Thus Group of S is 6.

Another way of finding group of element is looking at sub shells.

- If last sub shell of electron configuration is “s” or “p”, then group becomes A.

Examples

$_{19}$**K:** $1s^2 2s^2 2p^6 3s^2 3p^6 4s^1$. Since last sub shell is “s” group of K is A.

$_{35}$**Br:** $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^{10} 4p^5$. Since last sub shell is “p” group of Br is A.

- Elements in group B have electron configuration **ns** and **(n-1)d**, total number of electrons in these orbitals gives us group of element. Look at following

Examples

$_{26}$Fe: $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^6$, 6+2=8 B group

c. Finding Block of Elements:

The block of element corresponds to the last sub-energy levels which is, as we know, the highest one.

Examples

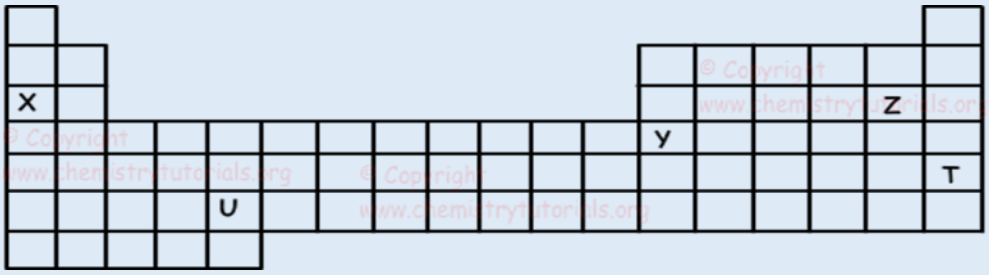
$_{16}$**S:** $1s^2 2s^2 2p^6 3s^2 3p^4$, p is the highest sub-energy level of electrons. Thus block of S is p. It belongs into p-block.

$_{23}$Cr: $1s^2 2s^2 2p^6 3s^2 3p^6 4s^2 3d^4$, d is the highest sub-energy level of electrons.

Thus block of Cr is d. It belongs to d-block.

Application activity 18.3

1. The elements X, Y, Z, T and U are given in the picture below.



Which one(s) of the following statements are correct, which one(s) are false for these elements.

a) X is alkaline metal

b) Z is halogens

c) U is lanthanide

d) T is noble gas

2. Explain the major parts of the periodic table?

18.4. Classification of the elements into blocks (s, p, d, f-block)

Activity 18.4

1. Using the s p d f notation write condensed electronic configuration for the following elements. Al (Z = 13), K (Z = 19), Sr (Z = 38) and Fe (Z = 26). Br(Z = 35), Zr(Z = 40), Nd(Z = 60) ,Yb (Z= 70)
2. For each element, identify the last subshell that is occupied by electrons and draw adequate conclusions.

The long form of periodic table can be divided into four main blocks. These are: s- block, p- block, d- block, and f-block.

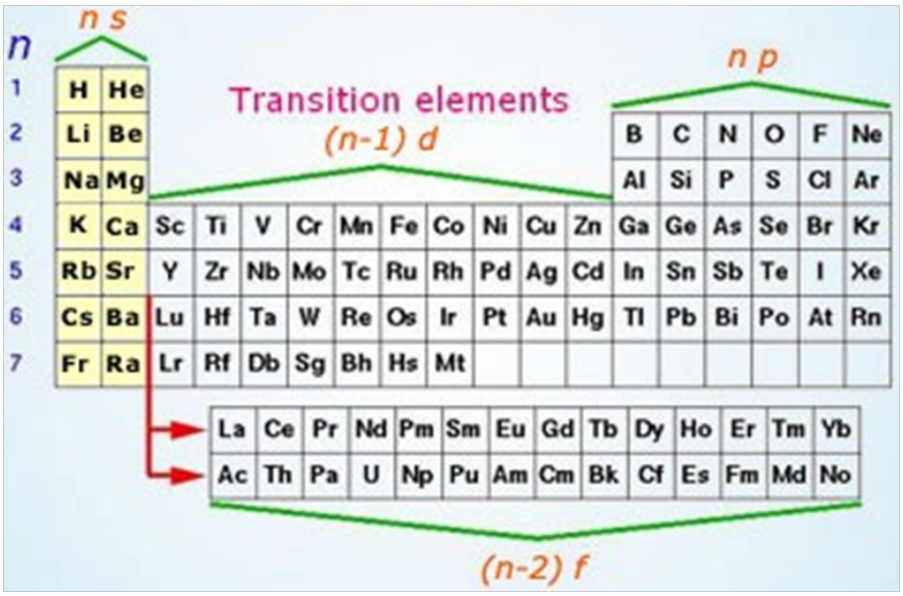


Figure 18.2: Blocks of the periodic table ofchemical elements

1. s-Block Elements

The elements in which the last electron enters the s-subshell of their outermost energy level are called s-block elements.

This block is situated at extreme left of the periodic table. It contains elements of groups 1 and 2.

Their general configuration is ns^{1-2}, where n represents the outermost shell.

The elements of group 1 are called alkali metals whereas the elements of group 2 are called alkaline earth metals.

2. p-Block Elemets

The elements in which the last electron enters the p-sub-shell of their outermost energy level are called p-block elements.

The general configuration of their outermost shell is $ns^2\ np^{1-4}$. The only exception is helium (ls^2). Strictly, helium belongs to the s-block but its positioning in the p-block along with other group 18 elements is justified because it has completely filled valence shell ($1s^2$) and as a result, exhibits properties characteristic of other noble gases. This block is situated at the extreme right of the periodic table and contains elements of groups 13, 14, 15, 16, 17 and 18 of the periodic table.

Most of these elements are non-metals, some are metalloids and a few others are heavy elements which exhibit metallic character.

The non-metallic character increases as we move from left to right across a period and metallic character increases as we go down the group.

3. d-Block Elements

The elements in which the last electron enters the d-subshell of the penultimate energy level are called d-block elements.

Their general valence shell configuration is (n-1) d^{1-10} ns^{1-2}, where n represents the outermost energy level. d-block contains three complete rows of ten elements in each. The fourth row is incomplete.

The three rows are called first, second and third transition series. They involve the filling of 3d, 4d and 5d orbitals respectively. The d-block contains elements of groups 3 to 12 of the periodic table.

4. f-Block Elements

The elements in which the last electron enters the f-sublevel of the anti-penultimate (third to the outermost shell) shell are called f-block elements.

Their general configuration is (n–2)f^{1-14} (n–1) d^{0-1} ns^2, where n represents the outermost shell.

They consist of two series of elements placed at the bottom of the periodic table.

The elements of first series follow lanthanum (^{57}La) and are called lanthanides. The elements of second series follow actinium (^{89}Ac) and are called actinides. Actinide elements are radioactive.

Many of them have been made only in nanogram quantities or even less by nuclear reactions. Chemistry of the actinides is complicated and is not fully studied.

Briefly, in the periodic table, elements are divided into:

- The ***s-block*** (contains reactive metals of Group 1A (1) and 2A (2)),
- The ***p-block*** (contains metals and non metals of Group 3A (13) through 8A (18)),
- The ***d-block*** (contains transition metals Group 3B (3) through Group 2B (12)), and
- The ***f-block*** (contains lanthanide and actinide series or inner transition metals).

Application activity 18.4

1. a) How many blocks into which chemical elements are classified and how are they named?

   b) What is the criterion used for this classification?

2. Which block of elements will tend to form: positive ions? Negative ions?
3. Give any two examples of elements of s block, p block, d block and f block.
4. Why d-block elements are called transition elements?
5. Why f-block elements are called inner transition elements?

18.5. Factors that influence the change of each physical property of the elements across a period and down a group

Activity 18.5

The elements in the periodic table display many trends which can be used to predict their physical properties.

Use available resources and explain three of the factors that you think can influence the physical properties of elements in the periodic table.

In the Periodic Table, there are a number of physical properties that are not really "similar" as it was previously defined, but are more trend-like. This means is that as you move down a group or across a period, you will see a trend-like variation in the properties.

The actual trends that are observed with the physical properties have to do with the following factors:

1. The number of protons in the nucleus (called the **nuclear charge**).
2. The number of electrons held between the nucleus and its outermost electrons (called the **shielding effect**).

Explanation:

Electrons in inner levels or shells tend to shield outer electrons from the full nuclear charge, which is reduced to ***effective nuclear charge*** (Z_{eff}).

Electrons are held in an atom or ion by the electrostatic attraction between the positively charged nucleus and the negatively charged electrons. In multi-electron species, the electrons do not experience the full positive charge of the nucleus due to shielding by electrons which lie between the electron of interest and the nucleus. The amount of positive charge that actually acts on an electron is called the ***effective nuclear charge***.

Electrons that have a greater *penetration* shield others more effectively. For example, electrons in level n = 1 shield those in level n = 2 very effectively, and those in n = 1 and n = 2 shield electrons in level n = 3. Electrons at the same level, but in different sublevels, also shield other electrons to some extent. The extent of penetration and shielding effect is in the order:

$s>p>d>f$.

The effective nuclear charge (Z_{eff}) greatly influence atomic properties. In general,

- Z_{eff} increases significantly across a period (left-to-right)
- Z_{eff} increases slightly down a group.

Application activity 18.5

Periodic Table Trends are influenced by the following factors:

a) Size of the atom

b) Nuclear charge

c) Shielding effect

Which of them affect the variation of physical properties in a group, in a period?

18.6. Variation of the physical properties down the group and across the period

Activity 18.6

Analyze the table about the molar ionization energy values for some elements and answers the questions that follow.

1. How is the variation of these values if you consider elements in the same

   a) period?

b) group?

Using available resources including internet, attempt to give a plausible explanation to the trend observed.

2. Analyze the trends of others physical properties and try to find each time appropriate reasons for those trends.

Periodic trends are specific patterns that are present in the periodic table that illustrate different aspects of a certain element, including its size and its electronic properties.

Major periodic trends include: **atomic radius, electronegativity, ionisation energy, electron affinity, melting point, density, metallic character**. Periodic trends, arising from the arrangement of the periodic table, provide chemists with an invaluable tool to quickly predict an element's properties. These trends exist because of the similar atomic structure of the elements within their respective group families or periods, and because of the periodic nature of the elements.

The elements in the periodic table are arranged in order of increasing atomic number. All of these elements display several other trends and we can use the periodic table to predict their physical properties. There are many noticeable patterns in the physical and chemical properties of elements as we descend in a group or move across a period in the Periodic Table.

18.6.1. Atomic radius

The atomic radius is the distance from the centre of the nucleus to the outermost shell of an atom when it is in the ground state.

The covalent radius is one-half the distance between the two nuclei of identical atoms that are joined together by a single covalent bond.

The metallic radius is one-half the distance between the nuclei of two atoms in contact in the crystalline solid metal.

A **cation** is an atom that has lost one of its outer electrons. Cations have a smaller radius than the atom that they were formed from.

An **anion** is an atom that has gained an outer electron. Anions have a greater radius than the atom that they were formed from.

Going **across a period**, the atomic radius **decreases**. This is caused by the **increase** in the number of protons and electrons across a period. One proton has a greater effect than one electron; thus, electrons are pulled towards the nucleus, resulting in a smaller radius.

Atomic radius **increases down the group.** This is caused by electron shielding. The valence electrons occupy higher levels due to the increasing quantum number (n). As a result, the valence electrons are further away from the nucleus as '**n**' increases. Electron shielding prevents these outer electrons from being attracted to the nucleus; thus, they are loosely held, and the resulting atomic radius is large.

18.6.2. Electronegativity

Electronegativity is a measure of the relative ability of an atom to attract the pair of electrons in a covalent bond. Two factors help to determine the electronegativity value of an element:

- The size of the nuclear charge
- The size of the atom

Going across a period, the electronegativity value increases: as the nuclear charge increases, the size of the atom decreases and hence there is a greater attraction between the nucleus and the pair of electrons in a covalent bond.

Going down a group, the electronegativity value decreases. The effect of the increase in the nuclear charge is less than the increase in atomic radius and the shielding of the inner electrons.

Important exceptions of the above rules include the noble gases, lanthanides and actinides. The noble gases possess a complete valence shell and do not usually attract electrons. The lanthanides and actinides possess more complicated chemistry that does not generally follow any trends. Therefore, noble gases, lanthanides, and actinides do not have electronegativity values.

As for the transition metals, although they have electronegativity values, there is little variance among them across the period and up and down a group. This is because their metallic properties affect their ability to attract electrons as easily as the other elements.

According to these two general trends, the most electronegative element is fluorine, with 3.98 Pauling units.

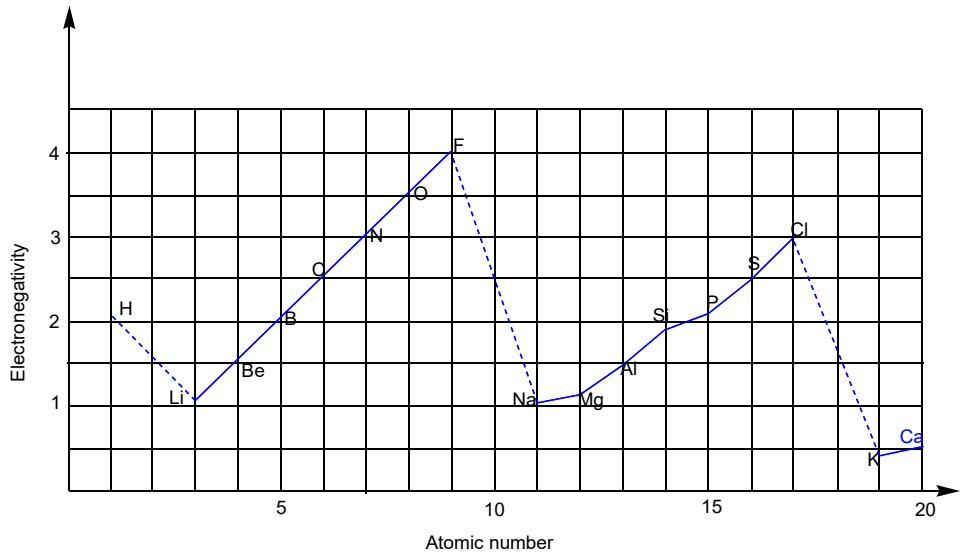


Figure 18.3: Variation in electronegativity with atomic number

18.6.3.Ionization energy (I.E)

Ionisation energy is the minimum energy (in kJ/mol) required to remove one mole of electrons from one mole of gaseous atom in the ground state.

In general, going across a period, the ionisation energy increases because the nuclear charge increases and hence electrons are more strongly attracted to the nucleus.

In general, the second ionisation energy of an element is always greater than the first ionisation. This is explained as follows: every time you remove an electron from an atom, the remaining electrons are more strongly attracted by the nucleus and it requires more energy to remove other electrons from the atom.

Hence: 1^{st} IE < 2^{nd} IE < 3^{rd} IE

Ionisation energy of rare gases or any species with an octet electronic structure show very high IE because the electron is being removed from a very stable electronic structure.

Down a group, the ionisation energy decreases because the electron which is removed is further from the nucleus. The nuclear charge also increases, but the extra inner electrons reduce the effect of the nuclear charge by shielding the outer electrons from the nucleus.

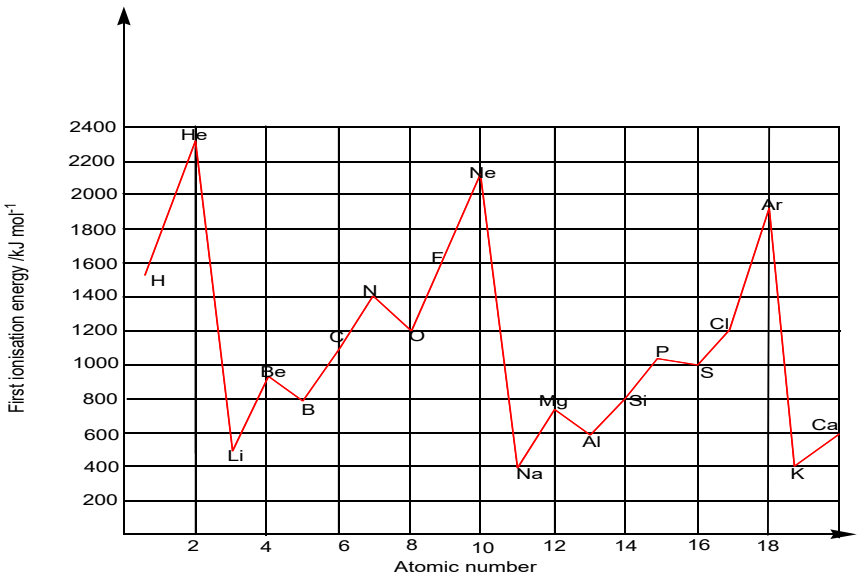


Figure18.4: First ionisation energies for the first twenty elements

18.6.4. Electron affinity

As the name suggests, **electron affinity*is the ability of an atom to accept an electron***. Unlike electronegativity, electron affinity is a quantitative measurement of the energy change that occurs when an electron is added to a neutral gas atom. The more negative the electron affinity value, the higher an atom's affinity for electrons.

That is the electron affinity (E.A.) is the energy change that occurs when an electron is added to a gaseous atom.

It can be either positive or negative value. The greater the negative value, the more stable the anion is.

The electron affinity ***is positive***: **$X(g)+e^- \rightarrow X^- +$ Energy** (Exothermic)

The electron affinity ***is negative:*** **$X(g)+e^- +$ Energy $\rightarrow X^-$** (Endothermic)

Electron affinity increases from left to right within a period. This is caused by the decrease in atomic radius. Moving from left to right across a period, atoms become smaller as the forces of attraction become stronger. This causes the electron to move closer to the nucleus, thus increasing the electron affinity from left to right across a period.

Electron affinity decreases from top to bottom within a group. This is caused by the increase in atomic radius. With a larger distance between the negatively-charged electron and the positively-charged nucleus, the force of attraction is relatively weaker. Therefore, electron affinity decreases.

18.6.5. The melting points and boiling points

Trends can be a bit complicated due to significant structural change from one element to another in the same group.

For groups 1 and 2, the melting and boiling points decrease down the group. As the atomic radius increases, the strength of the metallic bonding decreases.

For groups 7/17(halogens) and 18/0(noble gases), the melting and boiling points increase down the group, as the molecule becomes bigger with more electrons, the Van der Waals forces increase.

Across a period, melting and boiling points depend upon the structure and the bonding in the elements.

Table 18.2: Melting and boiling point of the elements in the third period

Element	Na	Mg	Al	Si	P_4	S_8	Cl_2	Ar
Type of structure	Metallic	Metallic	Metallic	Giant covalent	Simple molecule	Simple molecules	Simple molecules	Mono-atomic
Melting point/°C	98	650	660	1410	44	119	-101	-189
Boiling point/°C	890	1120	2450	2680	280	445	-34	-186

Sodium, magnesium and aluminium are metals. The strength of the metallic bonding depends upon the number of delocalised electrons in the metal structure. The melting point increases from sodium to aluminium.

Silicon exists as a giant covalent structure. Each silicon atom is covalently bonded to four other silicon atoms in a tetrahedral structure (similar to diamond). A great number of covalent bonds have to be broken to break up the giant structure.

Phosphorus, sulphur and chlorine are all simple molecular species. Phosphorus consists of P_4 molecules; sulphur consists of S_8 molecules; chlorine consists of Cl_2 molecules. The strength of the Van der Waals forces increases as the size of the molecule increases.

Argon exists as isolated atoms (it is monatomic) with weak van der Waals forces between atoms.

The illustrations below show the variation of melting and boiling point for some elements of the periodic table (Figures 18.13 and 18.14).

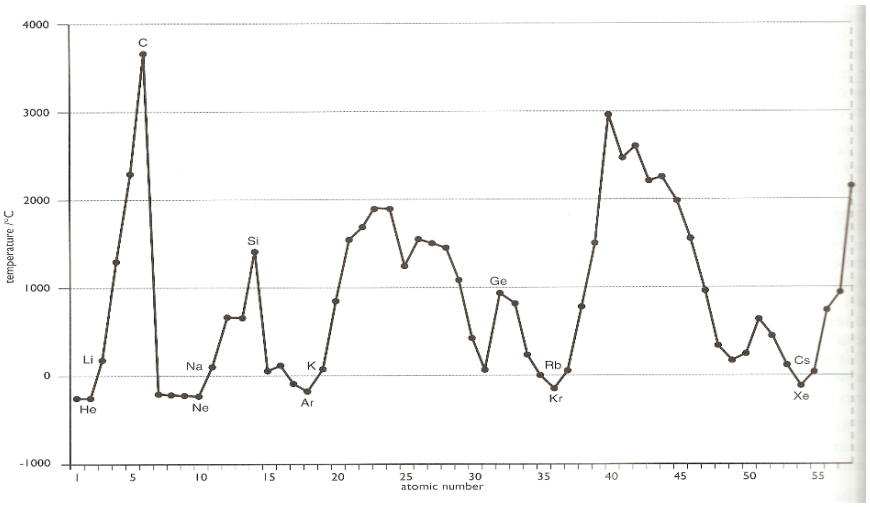


Figure 18.5: Variation of melting point with atomic number

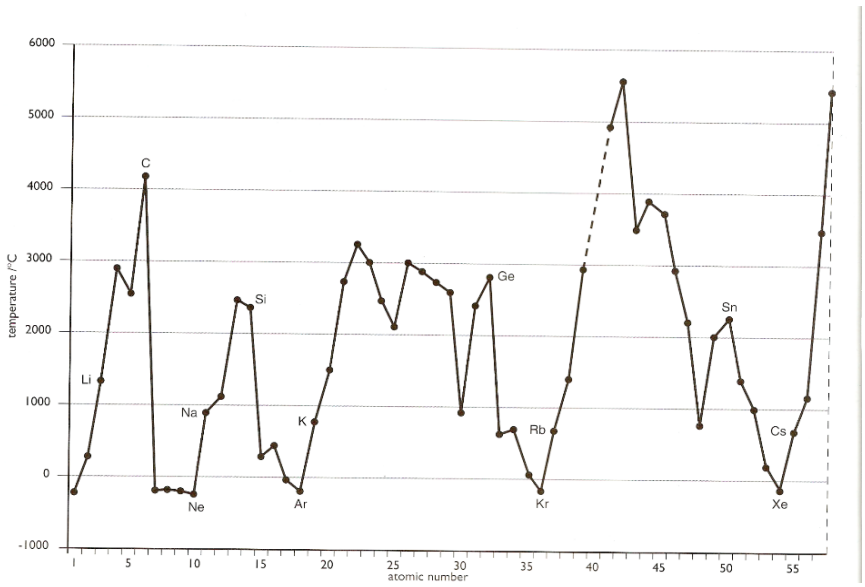


Figure 18.6: Variation of boiling point with atomic number

18.6.6. The density

The density of a substance is its mass per unit volume, usually in g/cm^3. The density is a basic physical property of a homogeneous substance; it is an intensive property, which means it depends only on the substance's composition and does not vary with size or amount.

The trends in density of elements can be observed in groups and periods of the periodic table. In general in any period of the table, the density first increases from group 1 to a maximum in the centre of the period because the

mass increases while the size decreases, and then the density decreases again towards group 18 because of the nature of bonds.

Going down a group gives an overall increase in density because even though the volume increases down the group, the mass increases more.

The variation of density with atomic number is shown in the Figure 18.7.

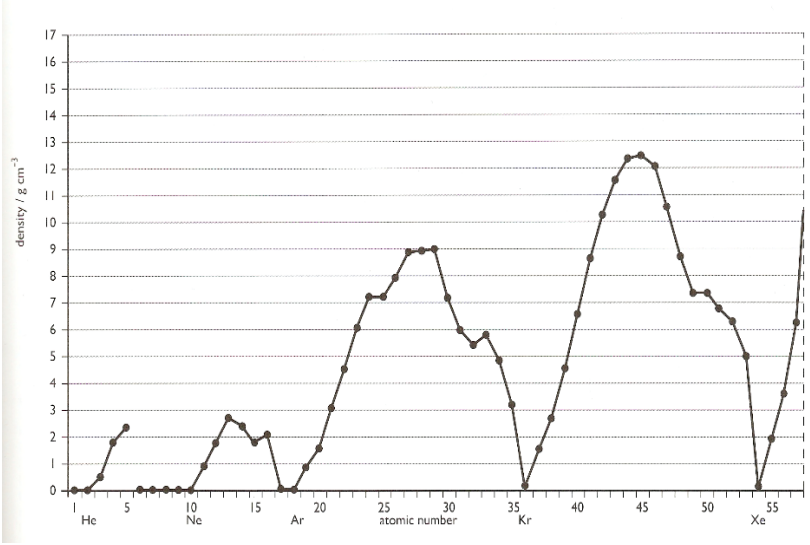


Figure 18.7: Variation of the density with atomic numbers

18.6.7. Electrical and thermal conductivity

The electrical conductivity is the ability of a substance to conduct an electric current.

Across a period, the conductivity increases as the number of delocalised electrons increases and then decreases as the metallic character decreases because electrons

within the covalent bonds are held much more tightly in these elements than in metals.

Down a group, the conductivity increases. The delocalised electrons become less attracted by the nuclei.

18.6.8. Electropositivity or Metallic character

Electropositivity or Metallic character refers to the level of reactivity of a metal. Metals tend to lose electrons in chemical reactions, as indicated by their low ionization energies.

Metals are located in the left and lower three-quarters of the periodic table, and tend to lose electrons to non-metals. Non-metals are located in the upper

right quarter of the table, and tend to gain electrons from metal. Metalloids are located in the region between the other two classes and have properties.

Metallic character is strongest for the elements in the leftmost part of the periodic table and tends to decrease as we move to the right of any period.

Within any group of the representative elements, the metallic character increases progressively going down.

Application activity 18.6

1. Explain why:
   a) the atomic radius decreases across a period.
   b) electronegativity decreases down a group.
   c) Electron affinity increases across a period
2. Describe and explain the trend in:
   a) Melting point
   b) metallic character
      i. Across a period (ii) down the group
3. Why is it always the case that the 2^{nd} ionisation energy is larger than the 1^{st} I.E, the 3^{rd} larger than the second I.E and so on?
4. In each of the following pairs, indicate which one of the two species is smaller: (a) Cl or Cl^-, (b) Na or Na^+, (c) O^{2-} or S^{2-}, (d) Mg^{2+} or Al^{3+}, (e) Au^+ or Au^{3+}.
5. The first and second ionization energies of K are 419 kJ/mol and 3052 kJ/mol, and those of Ca are590 kJ/mol and 1145 kJ/mol, respectively. Compare their values and comment on the differences

Skills lab 18

Consider an element with these properties and answer the questions that follow.

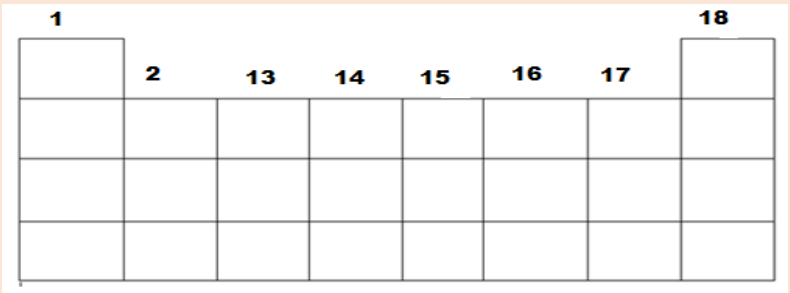
- Reacts mildly with 6 M HCl.
- Does not react with base or with water at room temperature.
- Has a good electrical conductivity.
- Is relatively cheap.

1. What element has these properties?

2. If this element undergoes combustion in air, write a balanced chemical equation for the reaction.
3. How much energy would it take (in kJ) to melt 1 kg of this element?
4. Plot a graph of the electrical conductivity for all elements in the group that includes this element. Is this element the best choice in its group for an electrical conductor? Explain.
5. Would this element make a good copper replacement for wiring in a house? Explain why or why not.

End unit assessment 18

1. What were the limitations of Newland's classification?
2. The following are coded groups/families of the representative elements of the periodic table (first 4 periods, s, p blocks only). The groups are in number of particular order. Use the hints below to identify the group and place of three elements of each group in their correct location in the periodic table: AOU, BVW, CKM, DLQ, ENT, FIJ, GPY, and HRS.



Hints

A has only one electron in p subshell

B is more electronegative than V

C has a larger atomic radius than both M and W

D has electronic configuration ending in p5

E is one of the most reactive metals

F has a smaller ionization energy than J

G has only 1 energy level with any electrons

H has one more proton than O and is in the same period as O

I is the largest alkaline earth metal

J has one more proton than E

K has electron configuration ending in p3

L has more filled energy levels than D

M is larger than K

N has the largest radius in its family

O is smaller than F but in the same energy level as F

P is smaller than Y

Q is the most reactive non-metal

R has the highest electronegativity in its family

T has the lowest density in its family

U more easily loses electrons (think about ionization energy) than either A or O

V has only 4 electrons in a p-subshell

W has 3 completely filled energy levels

Y has the lowest ionization energy in its family.

3. Justify the following statements:

   a) The first ionization energy of nitrogen is higher than that of oxygen even though nuclear charge of nitrogen is less compared to oxygen.

   b) Noble gases are having high ionization energies.

4. Explain why:

   a) Alkali metals (group 1 elements) are not found free in nature.

   b) Atomic radius of gallium is smaller than that of aluminium.(Z of Al = 13, Z of Ga = 31)

5. Arrange these elements in order of decreasing atomic size; sulphur, chlorine, aluminium and sodium. Does your arrangement demonstrate a group or a periodic trend?

6. List these ions in order of increasing ionic radius: N^{3-}, Na^{+}, F-, Mg^{2+}, O^{2-}

UNIT 19:

COVALENT BOND AND MOLECULAR STRUCTURES

Key Unit competence: Demonstrate how the nature of the bonding is related to the properties of covalent compounds and molecular structures.

Introductory Activity 19

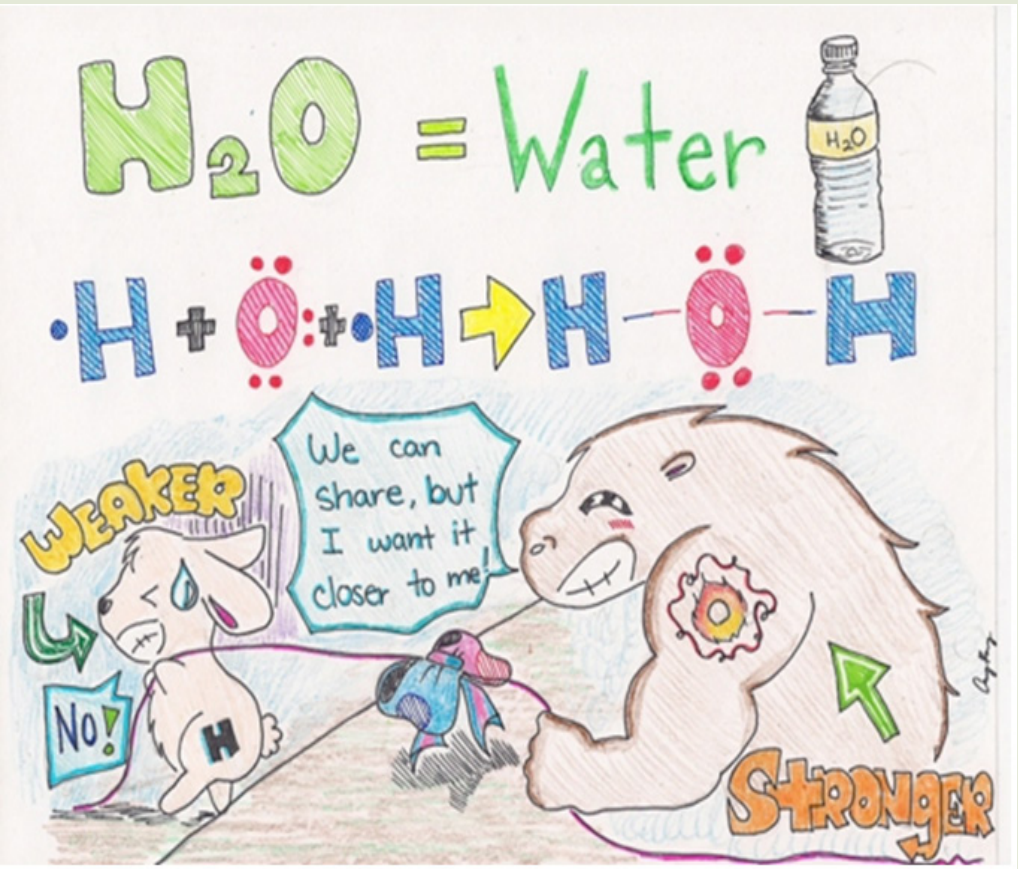


Figure 19.1

As you can see from the picture above, Oxygen is the big buff creature with the tattoo of "O" on its arm. The little bunny represents a Hydrogen atom. The blue and red bow tied in the middle of the rope, pulled by the two creatures represents the shared pair of electrons, a single bond.

Because the Hydrogen atom is weaker, the shared pair of electrons will be pulled closer to the Oxygen atom.

1. Suggest the measure used in Chemistry to describe the strength dedicated to oxygen, the stronger.
2. Suppose that the rope being pulled represents a single covalent bond. The electron contributed by hydrogen, the weaker, will be transferred to oxygen the stronger. If not, why?
3. Suppose again that we have two oxygens. They have the same strengths. What will happen to the pulled rope, or the shared pair of electrons?
4. Suggest a reason why, from the figure, one oxygen needed sharing with two hydrogens to form water.
5. Conclude about the possible types of covalent bonds.

19.1. Theories on the formation of covalent bond

Activity 19.1

1. Referring to the types of bonding already known, let us base on the electrovalent (ionic) bonding. Suppose that potassium metal and chlorine atom combine to form potassium chloride. Recall that each of those atoms engages in that bonding to get stability. How atoms in electrovalent bonding get the stability?
2. In covalent bonding, the same purpose remains but each atom contributes the same number of electrons to share in bonding. This is achieved by atoms of equal or close electronegativities.

   a) Draw the diagram to show the outer Bohr energy level for each of the following: Hydrogen (Z = 1), chlorine (Z = 17), potassium (Z = 19) and nitrogen (Z = 7).

   b) Basing on the information given, construct the H2, Cl2 and NH3 molecules by only showing the electrons on the outermost shell. Be informed that only the unpaired (single) electrons need to participate in the bonding. The paired ones are stable.

The sharing of pair of electrons between two atoms is referred to as a **covalent bond**. Normally, each atom that is participating in the covalent bond formation, contributes equal number of electrons to form pair(s) of electrons. The pair of electrons shared between the atoms is also known as **bond pair**.

The bond pair is ***strongly attracted by the nuclei of two atoms*** and thus by reducing the potential energy of them. This is the driving force of formation of covalent bond, which stabilizes the two atoms.

If two atoms share only one bond pair, that bond is referred to as a single bond. If two bond pairs are shared, that is known as a double bond. Likewise, a triple bond is formed when the atoms share three bond pairs.

A covalent bond is formed between two atoms when their electronegativity difference is less than 1.7 on Pauling's scale. Usually it is formed between two non-metals. For example;

- H_2 molecule: The electronegativity difference is **zero**.
- Cl_2 molecule: The electronegativity difference is **zero**.
- Hydrogen chloride (HCl): The electronegativity difference between them is 3.5 - 2.1 = **1.4**.
- Ammonia (NH_3): The electronegativity difference between them is 3.0 - 2.1 = **0.9**.
- H_2O molecule: The electronegativity difference between them is 3.5 - 2.1 = **1.4**.

To explain the formation of covalent bond, a simple qualitative model was developed by **Gilbert Newton Lewis** in **1916**.

According to this model:

- **Octet rule:**The inert gas atoms with 8 electrons in their outer shell (also known as valence shell) are ***highly stable***. The Helium atom with 2 electrons in its outer shell is also stable.
- Hence every atom tries to get nearest inert gas configuration by sharing electrons.

- The bond formed due to sharing of electrons is otherwise known as a covalent bond.
  - Only the electrons in the valence shell are contributed for sharing. The inner electrons, which are also known as core electrons do not participate in the bond formation.
  - In the formation of covalent bond between two atoms, each atom contributes its valence electrons to form pair(s) of electrons, which in turn is/are shared by both of them.

- Due to sharing of electrons, each atom gets nearest inert gas configuration.
- **Covalency:** The number of electrons contributed by the atom of an element in the formation of covalent compound is known as covalency of that element.
- In Lewis dot model, the **electrons in the valence shell of the atom are shown as dots around it.**

19.1.1. Lewis structures using octet rule

Lewis structures (also known as Dot and cross structures, Lewis dot diagrams, Lewis dot formulas, Lewis dot structures, and electron-dot structures) are diagrams that show the bonding between atoms of a molecule and the lone pairs of electrons that may exist in the molecule. A Lewis structure can be drawn for any covalently bonded molecule, as well as coordination compounds.

The Lewis Structure was named after Gilbert Newton Lewis, who introduced it in his 1916 article "The Atom and the Molecule".

Lewis structures extend the concept of the electron dot diagram by adding lines between atoms to represent shared pairs in a chemical bond.

Lewis structures show each atom and its position in the structure of the molecule using its chemical symbol. Lines are drawn between atoms that are bonded to one another (pairs of dots canbe used instead of lines). Excess electrons that form lone pairs are represented as pairs of dots, and are placed next to the atoms.

How to draw Lewis Structures

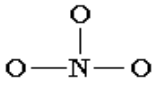
Let us use the **nitrate ion (NO_3^-)** as a typical example. An outline of how to determine the "best" Lewis structure for NO_3^- is given below:

1. Determine the **total number of valence electrons** in a molecule.

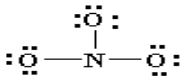
N	5
3 O	18
-ve	1
	24

2. Draw a **skeleton for the molecule which connects all atoms using only single bonds.** In simple molecules, the atom *with the most available sites for bonding* is usually placed central. The number of bonding sites is determined by considering the number of valence electrons and the ability of an atom to expand its octet.

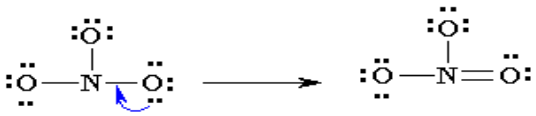
As you become better, you will be able to recognize that certain groups of atoms prefer to bond together in a certain way!



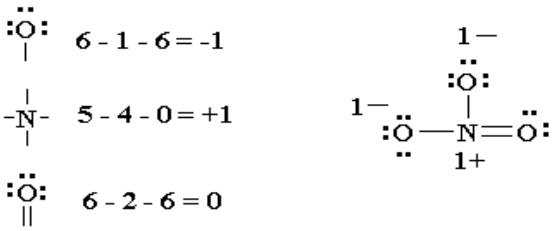
3. Of the 24 valence electrons in NO_3^-, 6 were required to make the skeleton. Consider the remaining 18 electrons and **place them so as to fill the octets of as many atoms as possible (start with the most electronegative atoms first then proceed to the more electropositive atoms).**



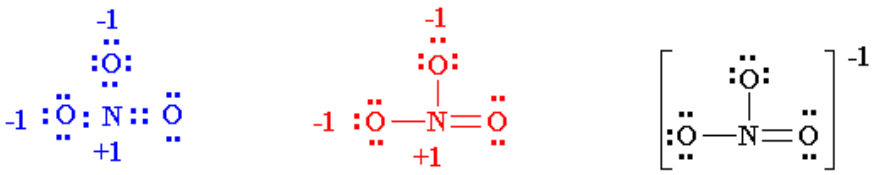
4. Are the **octets of all the atoms filled?** If not then fill the remaining octets by making multiple bonds (**make a lone pair of electrons, located on a more electronegative atom, into a bonding pair of electrons that is shared with the atom that is electron deficient).**



5. Check that you have the ***lowest formal charges (F.C.)*** possible for all the atoms, without violating the octet rule; F.C. = (valence e^-) - (1/2 bonding e^-) - (lone electrons).



6. Thus the Lewis structure of NO_3^- ion can be written in the following ways:



19.1.2. Lewis structures of unusual compounds that do not obey Octet Rule

There are ***three general ways*** in which the octet rule breaks down:

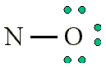
- Molecules ***with an odd number of electrons***
- Molecules ***in which an atom has less than an octet***
- Molecules ***in which an atom has more than an octet***

1. Odd number of electrons

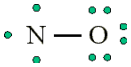
Consider the example of the Lewis structure for the molecule nitrous oxide **(NO)**:

- Total electrons: 6 + 5 = 11
- Bonding structure:

- Octet on “outer” element:



- Remainder of electrons (11-8 = 3) on “central” atom:

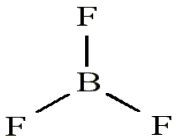


There are currently 5 valence electrons around the nitrogen. A double bond would place 7 around the nitrogen, and a triple bond would place 9 around the nitrogen. ***We appear unable to get an octet around each atom.***

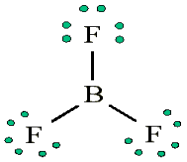
2. **Less than an octet** (most often encountered with elements of Boron and Beryllium)

Consider the example of the Lewis structure for boron trifluoride (BF_3):

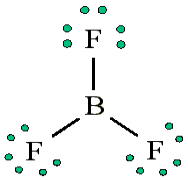
- Add electrons (3 x 7) + 3 = 24
- Draw connectivities



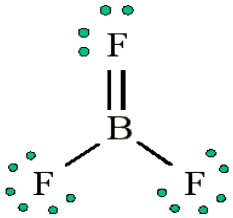
- Add octets to outer atoms:



- Add extra electrons (24 – 24 = 0) to central atom:



- Does central electron have octet? No, ***it has 6 electrons***. Add a multiple bond (double bond) to see if central atom can achieve an octet:

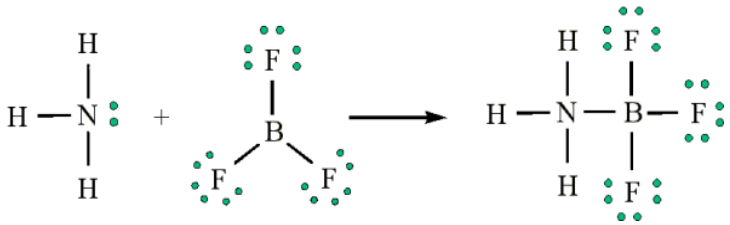


The central Boron now has an octet (there would be three resonance Lewis structures).

However, in this structure with a double bond the fluorine atom is sharing extra electrons with the boron.

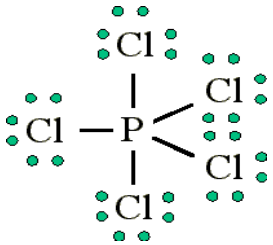
The fluorine would have a positive '**+**' partial charge, and the boron a negative '-' partial charge, this is inconsistent with the electronegativities of fluorine and boron. Thus, the structure of BF_3 with single bonds, and 6 valence electrons around the central boron is the most likely structure.

BF_3 reacts strongly with compounds which have an unshared (lone) pair of electrons which can be used to form a bond with the boron. Example: Reaction of BF_3 with ammonia.



3. More than an octet ***(most common example of exceptions to the Octet Rule***)

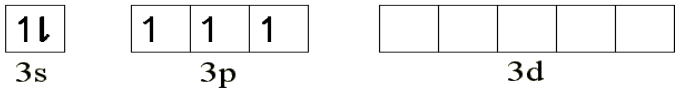
PCl_5 is a legitimate compound, whereas NCl_5 is not.



Expanded valence shells are observed only for elements in period 3 (i.e. n=3) and beyond.

The 'octet' rule is based upon available **ns** and **np-orbitals** for valence electrons (2 electrons in the **s** orbitals, and 6 in the **p** orbitals). Beginning with the **n=3** principle quantum number, the **d-orbitals become available** (**l=2**).

"The orbital diagram for the valence shell of phosphorous is:

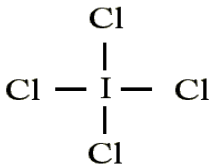


"Third period elements occasionally exceed the octet rule by using their empty d-orbitals to accommodate additional electrons."

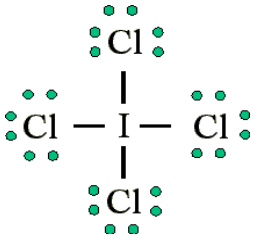
Size is also an important consideration: "The larger the central atom, the larger the number of electrons which can surround it". Expanded valence shells occur most often when the central atom is bonded to small electronegative atoms, such as **F**, **Cl** and **O**.

Example: Draw the Lewis structure for ICl_4^-

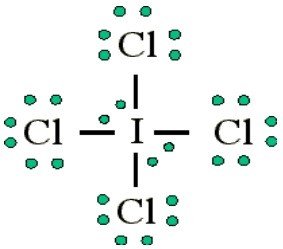
- Count up the valence electrons: 7 + (4 x 7) + 1 = 36 electrons
- Draw the connectivities:



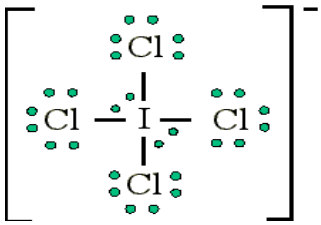
- Add octet of electrons to outer atoms:



- Add extra electrons (36-32=4) to central atom:



The ICl_4^- ion thus has 12 valence electrons around the central Iodine (in the 5*d* orbitals)



Other examples include: PCl_5 and SF_6

Application activity 19.1

1. Make a clear definition of the covalent bond.
2. For each of the following, write the electron configuration and choose one which violates the Octet Rule?

a) PCl_3
b) CBr_4
c) NF_3
d) OF_2
e) AsF_5
f) SO_4^{2-}
g) SF_6
h) HCl

19.2. Coordinate or dative covalent bond

Activity 19.2

The covalent bonding refers to the sharing of electrons to form a bond pair. There is a special type of covalence where the electrons shared are given by only one of the bonding atoms.

This special type of covalent bonding is described to be “dative”.

Formulate a succinct definition of the “dative covalent bond”.

A ***dative covalent bond***, or ***coordinate bond*** is a type of covalent bonding (i.e., electron sharing) where the shared electron pair(s) are completely provided by one of the participants in the union, and not by contributions from the two of them.

The contributors of these shared electrons are either ***neutral molecules*** which contain lone pair(s) of electrons on one of their atoms, or ***negatively charged groups*** (radicals) with free electrons to donate. Examples of these are: H_2O, NH_3 and CN^-.

Examples of coordinate bonding:

In the reaction between ammonia and hydrogen chloride a coordinate bond takes place forming solid ammonium chloride.

$$NH_3 + HCl \rightarrow NH_4Cl$$

In this reaction the hydrogen ion from the hydrogen chloride leaves its electrons and gets transferred to the lone pair of electrons on the ammonia molecule forming ammonium ions (NH_4^+). This is known as *a **coordinate bonding***.

Seeing that the hydrogen has left its electrons, the chloride will therefore have a negative charge while the ammonium will have a positive charge. The diagram below shows the reaction:

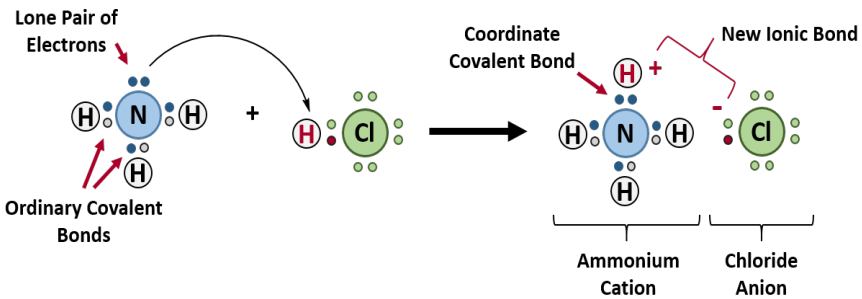


Figure 19.2: Reaction of ammonia with hydrogen chloride

Note: The complete compounds eventually formed comprises of the three types of bonding, i.e., ***covalent***, ***co-ordinate***and***electrovalent***. In NH_4Cl: Formation of NH_3 (covalency); formation of NH_4^+ (co-ordinate or dative bonding); and formation of NH_4Cl (electrovalency).

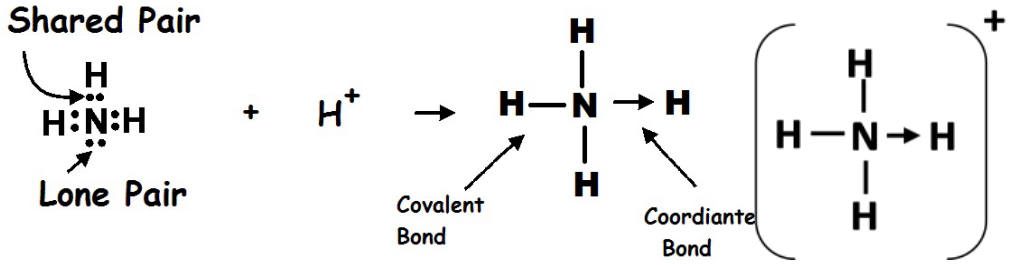


Figure 19.3: Reaction of ammonia with hydrogen ion

Dative covalent bonds are represented on drawings as an **"arrow"**, which usually points from the atom donating the lone pair to the atom accepting it.

Another example would be the ***reaction between ammonia and boron trifluoride***. Boron trifluoride is said to be ***electron deficient*** meaning it has 3 pairs of electrons at its bonding level but it is capable of having four pairs. In this reaction **the *ammonia is used to supply this extra lone pair*.**

A coordinate bond is formed where the lone pair from the nitrogen moves toward the boron. The end containing the nitrogen will therefore become more positive while the boron end will become more negative because it has received electrons.

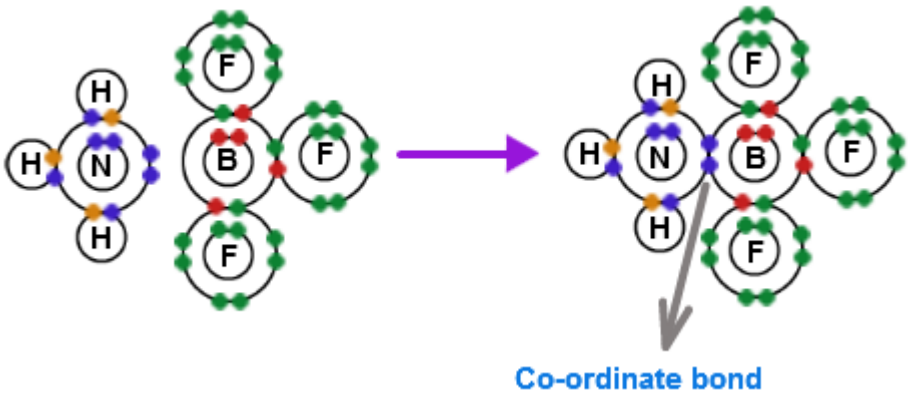


Figure 19.4: Reaction of ammonia with boron trichloride

Application activity 19.2

1. Give the difference and the similarity between a dative covalent bond and the normal covalent bond.
2. An aluminium chloride molecule reacts with a chloride ion to form the $AlCl_4^-$ ion.
   a) Name the type of bond formed in this reaction.
   b) Explain how this type of bond is formed in the $AlCl_4^-$ ion.
3. Co-ordinate bonding can be described as dative covalency.
   a) In this context, what is the meaning of each of the terms covalency and dative?
   b) Write an equation for a reaction in which a co-ordinate bond is formed.

19.3. Overlap of atomic orbitals to form covalent bonds

Activity 19.3

There are different theories describing the formation of a covalent bond. For your own, basing on the knowledge of the composition of the atom, formulate a paragraph which describe how two neutral atoms can be bound together to form a covalent bond (Include the way there will be attracted or repelled and what make them to remain together)

Covalent bonding ***occurs when atoms share electrons*** (Lewis Model), ***concentrating electron density between nuclei***. The build-up of electron density between two nuclei ***occurs when a valence atomic orbital of one atom merges with that of another atom*** (Valence Bond Theory).

The orbitals share a region of space, i.e. they overlap. *The overlap of orbitals allows two electrons of opposite spin to share the common space between the nuclei, forming a covalent bond.*

The simplest case to consider is the **hydrogen molecule, H_2**. When we say that the two electrons from each of the hydrogen atoms are shared to form a covalent bond between the two atoms, what we mean in Valence Bond Theory terms is that the ***two spherical 1s orbitals overlap,*** allowing the two electrons to form a pair within the two overlapping orbitals.

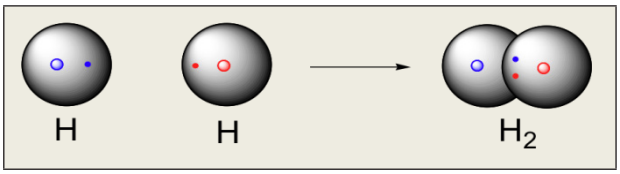


Figure 19.5: Overlaps of s-orbitals in hydrogen molecule

These two electrons are now attracted to the positive charge of both of the hydrogen nuclei, with the result that they serve as a sort of "chemical glue" holding the two nuclei together.

How far apart are the two nuclei? That is a very important issue to consider.

- If ***they are too far apart***, their respective *1s-orbitals cannot overlap*, and thus no covalent bond can form - they are still just two separate hydrogen atoms.
- As ***they move closer and closer together***, *orbital overlap begins to occur*, and a bond begins to form. This *lowers the potential energy* of the system, as new, *attractive* positive-negative electrostatic interactions become possible between the nucleus of one atom and the electron of the second. However, something else is happening at the same time: as the atoms get closer, the *repulsive* positive-positive interaction between the two nuclei also begins to increase.
- When ***the two nuclei are'too close'***, we have a ***very unstable, high-energy*** situation.

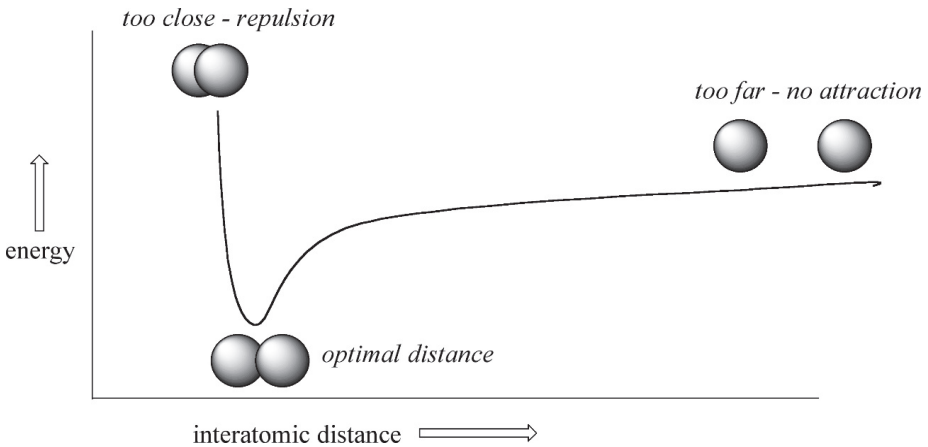


Figure 19.6: Bond length and bond energy

There is a defined optimal distance between the nuclei in which the potential energy is at a minimum, meaning that the combined attractive and repulsive forces add up to the greatest overall attractive force. This optimal inter-nuclear distance is the bond length. For the H2 molecule, this distance is $74x10^{-12}$ meters, or 0.74 Å (Å means angstrom, or 10^{-10} meters).

Likewise, the difference in potential energy between the lowest state (at the optimal inter-nuclear distance) and the state where the two atoms are completely separated is called the **bond energy**. For the hydrogen molecule, this energy is equal to about 104 kcal/mol.

Every covalent bond in a given molecule has a characteristic *length and strength*. Most covalent bonds in organic molecules range in strength *from just under 100 kcal/mol* (for a carbon-hydrogen bond in ethane, for example) *up to nearly 200 kcal/mol*.

Application activity 19.3

1. Explain the following terms:
   a) Bond length
   b) Bond energy
2. Account for the Valence Bond Theory.
3. Describe how the Lewis and Valence Bond Theories complement each other in explaining the formation of a covalent bond.

19.4. The concept of valence bond theory and formation of (σ) and (π) bonds

Activity 19.4

1. Recall and draw the shapes of s-orbital and p-orbital.
2. Using drawing, try to show different possible combinations of those orbitals overlapping.

Earlier we saw that covalent bonding requires the sharing of electrons between two atoms, so that each atom can complete its valence shell. But how does this sharing process occur? Remember that we can only estimate the likelihood of finding an electron in a certain area as a probability. In chemistry, valence bond (VB) theory is one of two basic theories—along with molecular orbital (MO) theory—that use quantum mechanics to explain chemical bonding.

According to this theory, a covalent bond is formed when two orbitals overlap (share the same space) to produce a new combined orbital containing two electrons of opposite spin.

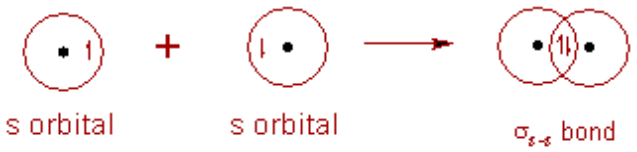
The valence bond theory was proposed by Heitler and London to explain the formation of covalent bond quantitatively using Quantum Mechanics. Later on, Linus Pauling improved this theory by introducing the concept of hybridization.

The main postulates of this theory are as follows:

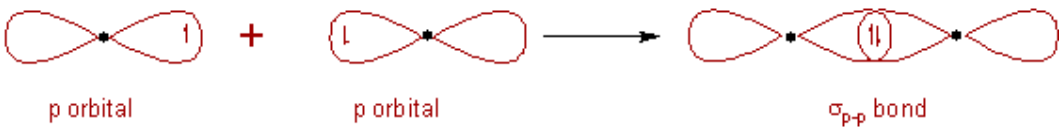
- A covalent bond is formed by the overlapping of two half filled valence atomic orbitals of two different atoms.
- The electrons in the overlapping orbitals get paired and confined between the nuclei of two atoms.
- The electron density between two bonded atoms increases due to overlapping. This confers stability to the molecule.
- Greater the extent of overlapping, stronger is the bond formed.
- The direction of the covalent bond is along the region of overlapping of the atomic orbitals i.e., covalent bond is directional.
- There are two types of covalent bonds based on the pattern of overlapping as follows:

  i. **σ-bond:** The covalent bond formed due to overlapping of atomic orbital along the inter nucleus axis is called σ-bond. It is a stronger bond and cylindrically symmetrical. Depending on the types of orbitals overlapping, the σ-bond is divided into following types:

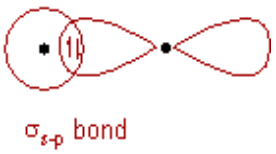
σ_{s-s} bond:



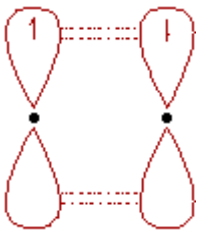
σ_{p-p} bond:



σ_{s-p} bond:



ii. π-bond: The covalent bond formed by sidewise overlapping of atomic orbitals is called π- bond. In this bond, the electron density is present above and below the inter-nuclear axis. It is relatively a weaker bond since the electrons are not strongly attracted by the nuclei of bonding atoms.



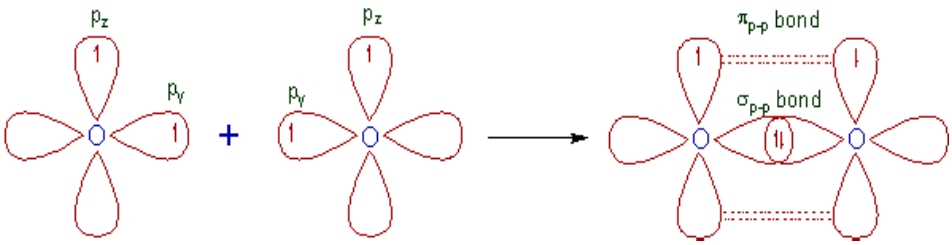
$\pi_{p\text{-}p}$ bond

Note: The 's' orbitals can only form σ-bonds, whereas the p, d& f orbitals can form both ***σ*** and ***π***-bonds.

Examples: Formation of covalent bonds in oxygen and hydrogen chloride molecules.

1. O_2 molecule:

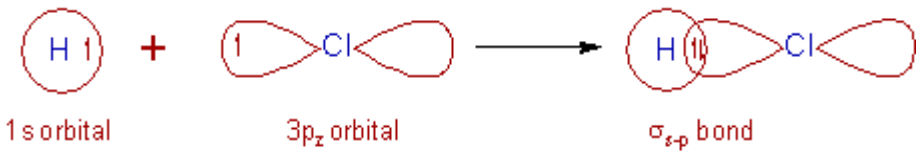
- The electronic configuration of O in the ground state is
  [He]2s^2 2p$_x^2$ 2p$_y^1$ 2p$_z^1$.
- The half filled **2p$_y$** orbitals of two oxygen atoms overlap along the inter-nuclear axis and form $\sigma_{p\text{-}p}$ bond.
- The remaining half filled **2p$_z$** orbitals overlap laterally to form a $\pi_{p\text{-}p}$ bond.
- Thus a double bond (**one** $\sigma_{p\text{-}p}$ and one $\pi_{p\text{-}p}$) is formed between two oxygen atoms.



2. HCl molecule:

- In the ground state, the electronic configuration of hydrogen atom is **1s**^{1}.
- And the ground state electronic configuration of Cl atom is
  [Ne]3s^2 3p$_x^2$ 3p$_y^2$ **3p$_z^1$**.

- The half filled 1s orbital of hydrogen overlap with the half filled $3p_z$ atomic orbital of chlorine atom along the inter-nuclear axis to form a $\sigma_{s\text{-}p}$ bond.



Need for modification of valence bond theory

The old version of Valence Bond Theory is limited to diatomic (like N_2) and binary (like HCl) molecules only. It could not explain the structures and bond angles of molecules with more than three atoms. For example: It could not explain the structures and bond angles of H_2O, NH_3 etc.

However, in order to explain the structures and bond angles of these molecules, **Linus Pauling** modified the Valence Bond Theory using **hybridization concept**.

Note: Both types of overlapping orbitals can be related to bond order.

- Single bonds have ***one sigma bond***,
- Double bonds consist of one σ and one π-bond,
- Triple bonds contain one σ and two π-bonds.

Application activity 19.4

1. In chemistry, valence bond (VB) theory is one of two basic theories—along with molecular orbital (MO) theory—that use quantum mechanics to explain chemical bonding. In one sentence, describe how the covalent bond is formed, according to this theory.
2. Explain why the Valence Bond Theory needed to be modified.
3. Describe the types of covalent bond basing on the pattern of overlapping.

19.5. Hybridisation and its types

Activity 19.5

Just as animals can be cross bred to form hybrids (having the characteristics of each of the intermixed animal), atomic orbitals also can be intermixed. Make a research to formulate a definition of hybridization and state the common types of hybridization of the atomic orbitals.

19.5.1. Definition of hybridisation

The concept hybridization involves the "*cross breeding*" of atomic orbitals to create "*new*" orbitals. Hence the use of the term "hybrid": Think of a *hybrid* animal which is a cross breed of two species.

Hybridization is the process of "intermixing of atomic orbitals of nearly same energies to form same number of identical and degenerate (having equivalent energies) new type of orbitals". Orbitals which are formed in hybridization process are called hybrid orbitals.

During hybridization, the atomic orbitals with different characteristics are mixed with each other. Hence there is no meaning of hybridization between same type of orbitals i.e., mixing of two '**s**' orbitals or two '**p**' orbitals is not called hybridization. However orbital of '**s**' type can mix with the orbitals of '**p**' type or of '**d**' type.

Keep in mind that only the orbitals of nearer energy values can participate in the hybridization. Based on the type and number of orbitals, the hybridization can be subdivided into following types.

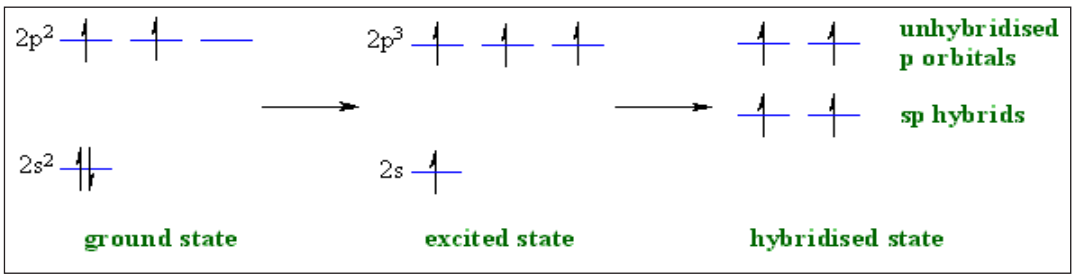
19.5.2. Types of hybridization

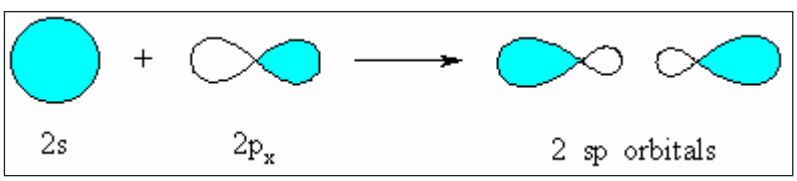
1. sp-Hybridization

*Intermixing of one '**s**' and one '**p**' orbitals of almost equal energy to give two identical and degenerate hybrid orbitals is called **'sp' hybridization**.*

These two sp-hybrid orbitals are arranged ***linearly*** at by making **180°** of angle.

They possess 50% '**s**' and 50% '**p**' character.



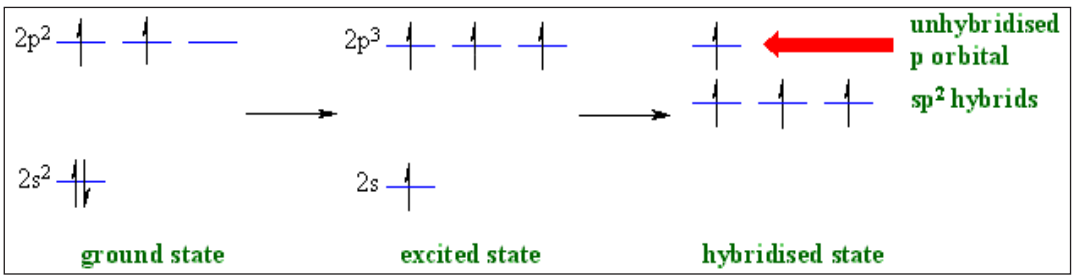


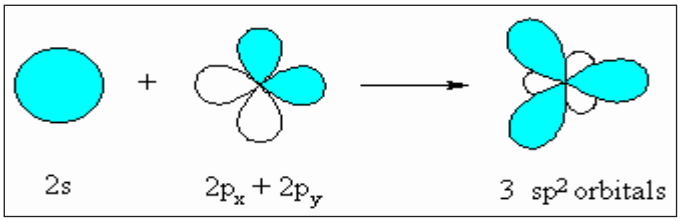
2. sp^2-Hybridization

*Intermixing of **one** '**s**' and **two** '**p**' orbitals of almost equal energy to give three identical and degenerate hybrid orbitals is known as **sp^2 hybridization**.*

The three sp^2 hybrid orbitals are oriented in ***trigonal planar*** symmetry at angles of **120°** to each other.

The sp^2 hybrid orbitals have 33.3% 's' character and 66.6% 'p' character.



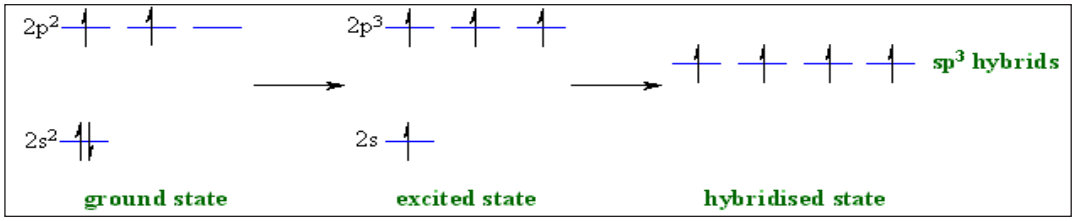


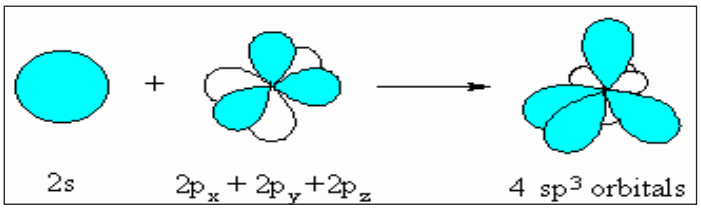
3. sp^3-Hybridization

*In sp^3 hybridization, **one 's'** and **three 'p'** orbitals of almost equal energy intermix to give four identical and degenerate hybrid orbitals.*

These four sp^3 hybrid orbitals are oriented in ***tetrahedral***symmetry with **109°28'** angle with each other.

The sp^3 hybrid orbitals have 25% '**s**' character and 75% '**p**' character.



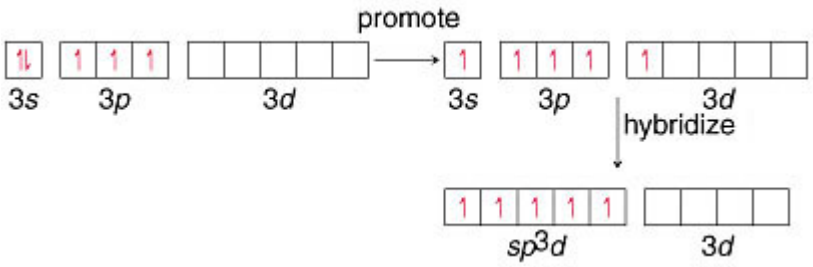


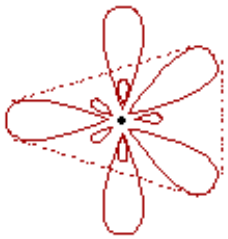
4. *sp^3d-Hybridization*

In sp^3d hybridization, ***one 's', three 'p'*** *and* ***one 'd'*** *orbitals of almost equal energy intermix to give five identical* and degenerate hybrid orbitals, which are arranged in ***trigonal bipyramidal*** symmetry.

Among them, three are arranged in trigonal plane and the remaining two orbitals are present above and below the trigonal plane at right angles.

The five sp^3d hybrid orbitals have 20% 's', 60% 'p' and 20% 'd' characters.



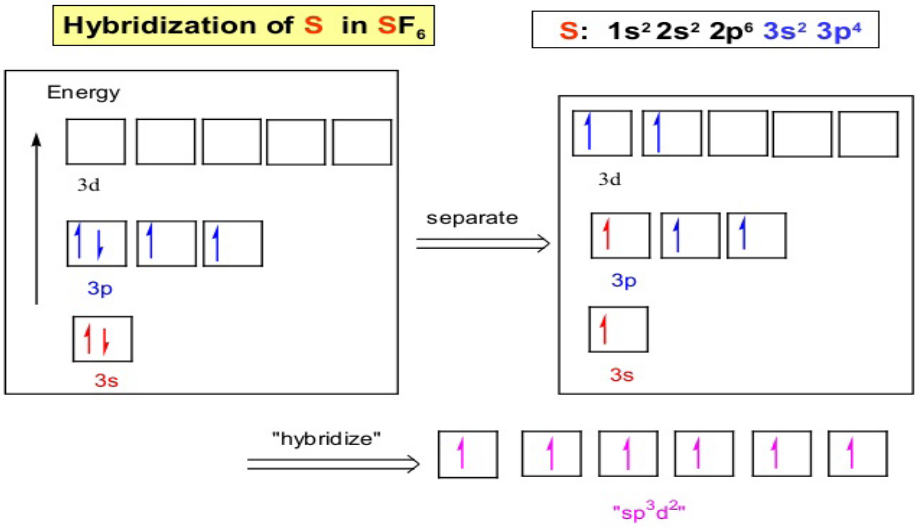


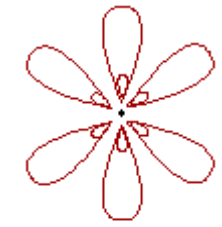
sp^3d hybrid orbitals

5. sp^3d^2-Hybridization

Intermixing of ***one 's', three 'p'*** *and* ***two 'd' orbitals*** *of almost same energy by giving six identical and degenerate hybrid orbitals is called* ***sp^3d^2 hybridization***.

These six sp^3d^2 orbitals are arranged in ***octahedral*** symmetry by making **90°** angles to each other. This arrangement can be visualized as four orbitals arranged in a square plane and the remaining two are oriented above and below this plane perpendicularly.





sp^{3}d^2 hybrid orbitals

Conditions for the hybridization

- Orbitals of same element should take part in the hybridization.
- There should be minimum difference between the orbitals undergoing hybridization.

Characteristics of Hybridization

- During hybridization the number of hybrid orbitals formed is equal to the number of atomic orbitals involved in hybridization.
- Hybrid orbitals form more stable and stronger bonds than pure atomic orbitals.
- Hybridization does not take place in isolated atoms and possible in those atoms which are prior to participate in chemical bonding.

Application activity 19.5

1. What is meant by hybridization?
2. Use a table to state five types of the common types of hybridization and for each; give the name of the shape and bond angles they present.

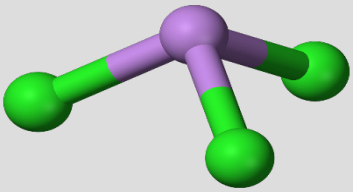
19.6. VSEPR theory

Activity 19.6

Find out the ball and stick molecular models in the laboratory and try to construct the following molecules: H_2O, HCl, CH_4, CO_2 and NH_3

- For each of the molecule constructed, try to draw it in your exercise book as it appears.

The model for PCl_3 is given for you as example

	Shape of PCl_3 (Trigonal pyramidal) Bond angles: 109°28’

- Try to name the shape of each of the molecule constructed.
- Give the approximate bond angles in each of the obtained molecule.

In order to predictthe geometry of molecules, ***Nyholm*** and ***Gillespie*** developed a qualitative model known as **Valence Shell Electron Pair Repulsion Theory (*VSEPR Theory*)**. The basic assumptions of this theory are summarized below.

1. The electron pairs in the valence shell around the central atom of a molecule repel each other and tend to orient in space so as to minimize the repulsions and maximize the distance between them.
2. There are two types of valence shell electron pairs namely, **(i)** Bond pairs and **(ii)** Lone pairs.

- Bond pairs are shared by two atoms and are attracted by two nuclei. Hence they occupy less space and cause less repulsion. It is also called “*sharing pair*”.
- Lone pairs are not involved in bond formation and are in attraction with only one nucleus. Hence they occupy more space. As a result, the lone pairs cause more repulsion.

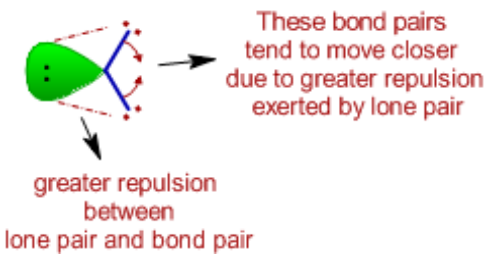
The *order of repulsion* between different types of electron pairs is as follows:

Lone pair - Lone pair > Lone Pair - Bond pair > Bond pair - Bond pair

Note: The bond pairs are usually represented by a solid line, whereas

the lone pairs are represented by a lobe with two electrons.

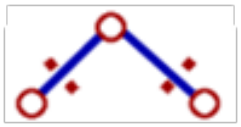
3. In VSEPR theory, the multiple bonds are treated as if they were single bonds. The electron pairs in multiple bonds are treated collectively as a single super pair. The repulsion caused by bonds increases with increase in the number of bonded pairs between two atoms i.e., a triple bond causes more repulsion than a double bond which in turn causes more repulsion than a single bond.
4. The shape of a molecule can be predicted from the ***number*** and ***type of valence shell electron pairs*** around the central atom. When the valence shell of central atom contains only bond pairs, the molecule assumes symmetrical geometry due to even repulsions between them. However the symmetry is distorted when there are also lone pairs along with bond pairs due to uneven repulsion forces.
5. Primary & Secondary effects on bond angle and shape:
   i. The bond angle decreases due to the presence of lone pairs, which cause more repulsion on the bond pairs and as a result the bond pairs tend to come closer.



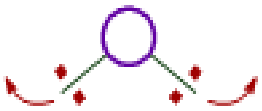
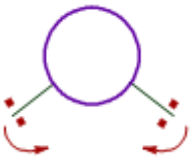
   ii. The repulsion between electron pairs increases with increase in electronegativity of central atom and hence the bond angle increases. The bond pairs are closer and thus by shortening the distance between them, which in turn increases the repulsion. Hence the bonds tend to move away from each other.

	The bond pairs tend to move away from each other since the distance between them is shortened as they are more localized on more electronegative central atom.

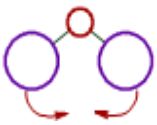
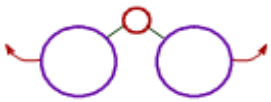
   However the bond angle decreases when the electronegativities of ligand atoms are more than that of central atom. There is increase in the distance between bond pairs since they are now closer to ligand atoms. Due to this, they tend to move closer resulting in the decrease in bond angle.

	The bond pairs tend to come closer since the distance between them is increased as they are more localized on more electronegative ligand atoms.

iii. The bond angle decreases with increase in the size of central atom.

	On smaller central atoms the bond pairs are closer and hence tend to move away from each other so as to minimize repulsion. Hence bond angle will be more.
	On bigger central atoms, the bond pairs are more distant from each other and hence there is less repulsion. Hence they tend to move closer and thus by decreasing the bond angle.

However the bond angle increases with increase in the size of ligand atoms, which surround the central atom.

	There is less repulsion between smaller ligand atoms and they can move closer to each other and thus decrease the bond angle.
	There is more repulsion between bigger ligand atoms and hence they tend to move away from each other. Thus bond angle increases.

iv. The bond angles are also changed when multiple bonds are present. It is due to uneven repulsions.

6. When there are two or more resonance structures, the VSEPR theory is applicable to any of such contributing structure.

The shape of molecule and also the approximate bond angles can be predicted from the **number** and **type** of electron pairs in the valence shell of central atom as tabulated below.

In the following table the molecule is represented by "**AXE**" notation, where

A = Central atom

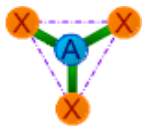
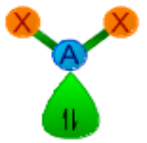
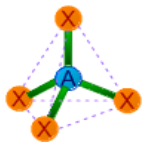
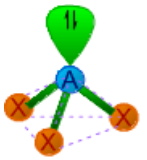
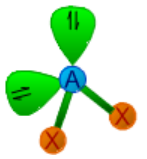
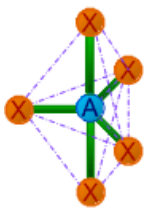
X = Ligand atom bonded to the central atom either by a single bond or by multiple bond; indicating a bond pair.

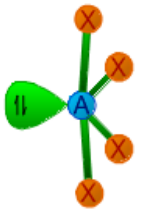
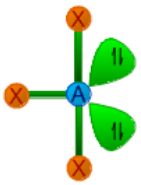
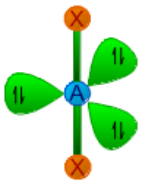
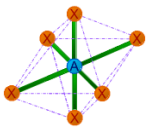
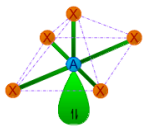
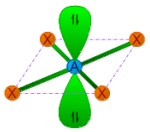
E = Lone pair

Note:

- The sum of number of ligand atoms (**X**) and number of lone pairs (**E**) is also known as **steric number.**
- The bond pairs are shown as green colored thick lines, whereas the lone pairs are shown as point charges using green colored lobes.

Table 19.1: Names of shapes and bond angles

Electron Arrange-ment	Formula	Hybri-dization	Bonding Pairs	Lone Pairs	Name of the Shape	Shape	Bond Angle
Linear	AX or AX2	Sp	1 or 2	0	Linear		180°
Trigonal Planar	AX_3	sp^2	3	0	Trigonal planar		120°
			2	1	Bent		116.8°
Tetra-hedral	AX_4	sp^3	4	0	Tetra-hedral		109.5°
			3	1	Trigonal pyra-midal		107°
			2	2	Bent		104.5°
Trigonal Bipy-ramidal	AX_5	sp^3d	5	0	Trigonal Bipy-ramidal		120° and 90°

			4	1	See-saw		<120° and 90°
			3	2	T-shaped		90° and 180°
			2	3	Linear		180°
Octa-hedral	AX6	sp^3d^2	6	0	Octa-hedral		90°
			5	1	Square pyra-midal		90° and 180°
			4	2	Square planar		90° and 180°

Determination of shape of a molecule

- The first step in determination of shape of a molecule is to write the Lewis dot structure of the molecule.
- Then find out the number of bond pairs and lone pairs in the valence shell of central atom.
- While counting the number of bond pairs, treat multiple bonds as if they were single bonds. Thus electron pairs in multiple bonds are to be treated collectively as a single super pair.
- Use the above table to predict the shape of molecule based on steric number and the number of bond pairs and lone pairs.

The following table shows some examples for each type of shapes:

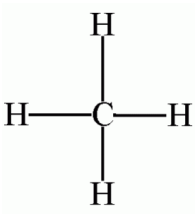
Table 19.2: Examples for each type of shapes

Steric Number	Number of Bond pairs	Number of Lone pairs	Formula	Examples
1	1	0	**AX**	ClF, BrF, BrCl, HF, O_2
2	2	0	**AX_2**	$BeCl_2$, $HgCl_2$, CO_2
3	3	0	**AX_3**	BF_3, CO_3^{2-}, NO_3^-, SO_3
	2	1	**AX_2E**	SO_2, $SnCl_2$, O_3, NSF, NO_2^-
4	4	0	**AX_4**	CH_4, $SiCl_4$, NH_4^+, PO_4^{3-}, SO_4^{2-}, ClO_4^-
	3	1	**AX_3E**	NH_3, PCl_3, XeO_3
	2	2	**AX_2E_2**	H_2O, SCl_2, Cl_2O, OF_2
5	5	0	**AX_5**	PCl_5, SOF_4
	4	1	**AX_4E**	SF_4, $TeCl_4$
	3	2	**AX_3E_2**	ClF_3, BrF_3, $BrCl_3$
	2	3	**AX_2E_3**	XeF_2, I_3^-
6	6	0	**AX_6**	SF_6
	5	1	**AX_5E**	ClF_5, BrF_5, ICl_5
	4	2	**AX_4E_2**	XeF_4

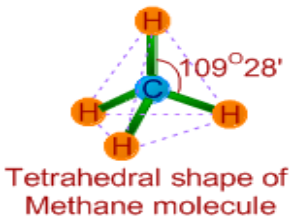
Worked examples

1. Methane (CH_4)

- The Lewis structure of methane molecule is:



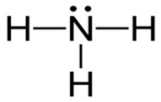
- There are 4 bond pairs around the central carbon atom in its valence shell. Hence it has tetrahedral shape with 109o28' of bond angles.



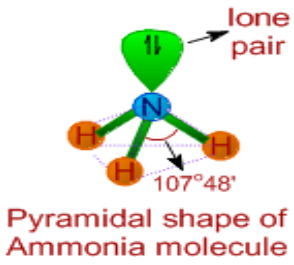
Tetrahedral shape of methane molecule

2. Ammonia (PCl_3)

- The Lewis structure of ammonia indicates there are three bond pairs and one lone pair around the central nitrogen atom.



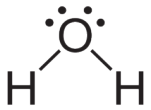
- Since the steric number is 4, its structure is based on tetrahedral geometry. However, its shape is pyramidal with a lone pair on nitrogen atom.
- The bond angle is decreased from 109o28' to 107o48' due to repulsion caused by lone pair on the bond pairs.

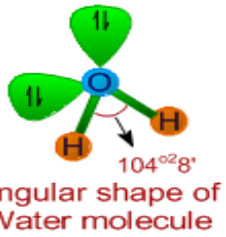


Tetrahedral shape of ammonia molecule

3. Water (H_2O)

- It is evident from the Lewis structure of water molecule; there are two bond pairs and two lone pairs in the valence shell of oxygen. Hence its structure is based on tetrahedral geometry. However its shape is angular with two lone pairs on oxygen.





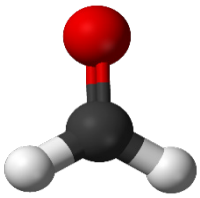
Tetrahedral shape of water molecule

- The bond angle is decreased to 104o28' due to repulsions caused by lone pairs on bond pairs. It can be noted that the bond angle decreases with increase in the number of lone pairs on the central atom.

4. Formaldehyde (HCHO)

There are three bond pairs around the central carbon atom. The double bond between C and O is considered as a single super pair. Hence the shape of the molecule is trigonal planar and the bond angles are expected to be equal to 120°.

However, the C=O exerts more repulsion on the C-H bond pairs. Hence the <H-C-H bond angle will be less than 120° and the <H-C-O is greater than 120°.

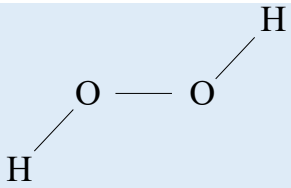


Application activity 19.6

1. There are two types of valence shell electron pairs such as Bond pairs and Lone pairs.
   a) State and explain these electron pairs.
   b) Show the increasing order of repulsion between different types of electron pairs.
2. Questions about covalent bonding.
   a) Sketch the shapes of each of the following molecules, showing any lone pairs of electrons. In each case, state the bond angle(s) present in the molecule and name the shape.

Molecule	Sketch of shape	Bond angle(s)	Name of shape
BF_3			
NF_3			
ClF_3			

   b) State the types of intermolecular force which exist, in the liquid state, between pairs of BF_3 molecules and between pairs of NF_3 molecules.
   c) Name the type of bond which you would expect to be formed between a molecule of BF_3 and a molecule of NF_3. Explain how this bond is able to form.
3. The following diagram shows a hydrogen peroxide molecule.



a) On the diagram above, draw the lone pairs, in appropriate positions, on the oxygen atoms.

b) Indicate, on the diagram, the magnitude of one of the bond angles.

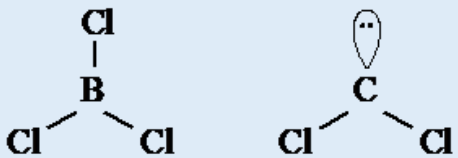
c) When considering electron pair repulsions in molecules, why does a lone pair of electrons repel more strongly than a bonding pair?

4. Phosphorus and nitrogen are in Group V of the Periodic Table and both elements form hydrides. Phosphine, PH_3, reacts to form phosphonium ions, , in a similar way to that by which ammonia, NH_3, forms ammonium ions,

a) Give the name of the type of bond formed when phosphine reacts with an H^+ion. Explain how this bond is formed.

b) Draw the shapes, including any lone pairs of electrons, of a phosphine molecule and of a phosphonium ion. Give the name of the shape of the phosphine molecule and state the bond angle found in the phosphonium ion.

5. The shape of the molecule BCl_3 and that of the unstable molecule CCl_2 are shown below.



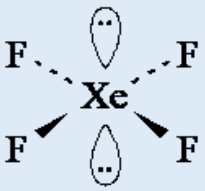
a) Why is each bond angle exactly 120° in BCl_3?

b) Predict the bond angle in CCl2 and explain why this angle is different from that in BCl_3

c) Give the name which describes the shape of molecules having bond angles of 109° 28'.

Give an example of one such molecule.

6. The shape of the XeF_4 molecule is shown below.



a) State the bond angle in XeF_4

b) Suggest why the lone pairs of electrons are opposite each other in this molecule.

c) Name the shape of this molecule, given that the shape describes the positions of the Xe and F atoms only.

7. Copy and complete the following table:

Compound	SiF_4	SF_4
Lewis Structure (with formal charges)		
Molecular shape (in words)		
Molecular shape (in words)		
Hybridization		
Number of π- bonds		

8. Boron, nitrogen and oxygen form fluorides with molecular formulae BF_3,NF_3 and OF_2

a) Draw the shape of each molecule showing the position of lone pairs if any.

b) Give the bond angle in each case, explaining your reasons.

19.7. Polarity of the covalent bond in relation to difference in electronegativity

Activity 19.7

The following figures show two types of covalent bond, namely, polar and non-polar.

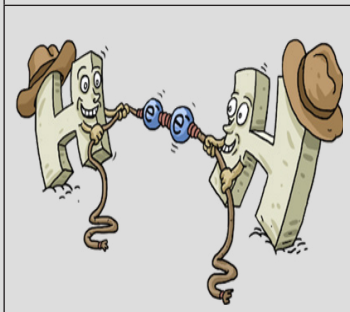
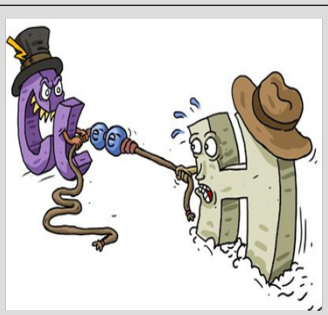
A	B
	

Figure 19.7: H-H and H-Cl interactions

1. Covalent bond is formed between two atoms with similar or close ability to attract electrons towards themselves, and this is the reason why they share electrons without being transferred.

    a) What is the name of the measure used to compare that ability?

    b) When the strengths of both atoms are equal, the covalent bond will be non-polar. Is figure A represents the polar or non-polar bond? Why?

    c) Look at the figure B. The atom, in the zone with more electrons, will have a partial negative charge. In which zone can we label the partial negative charges?

2. Experiment to study the behaviour of water on the charged species.

    - Fill a burette with water.
    - Open the tap and bring a charged ebonite rod close to the stream of water running from the jet.
    - Observe the direction of water and using a drawing, note the change, if there is any.

A quantity termed 'electronegativity' is used to determine the polarity of the covalent bond; whether a given bond will be non-polar covalent, polar covalent, or ionic.

Electronegativity is defined as the ability of an atom in a particular molecule to attract electrons to itself (the greater the value, the greater the attractiveness for electrons).

Fluorine is the *most* electronegative element (electronegativity = 4.0), the *least* electronegative is Caesium (electronegativity = 0.7).

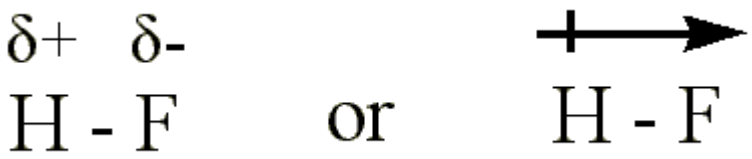
The bond pair is equally shared in between two atoms when the electronegativity difference between them is zero or nearer to zero. In this case, neither of the atoms gets excess of electron density and hence carry no charge. This is called **non-polar covalent bond**.

However, when there is a considerable difference in the electronegativity, the bond pair is no longer shared equally between the atoms. It is shifted slightly towards the atom with higher electronegativity by creating partial negative charge (represented by δ^-**)** over it, whereas, the atom with less electronegativity gets partial positive charge (represented by δ^+). This type of bond is also referred to as **polar covalent bond.**

We can use the difference in electronegativitybetween two atoms to gauge the polarity of the bonding between them.

In F_2 the electrons are ***shared equally*** between the atoms, the bond is ***non-polar covalent.***

In **HF** the fluorine atom has greater electronegativity than the hydrogen atom. The ***sharing of electrons in HF is unequal***: the fluorine atom attracts electron density away from the hydrogen (the bond is thus a ***polar covalent bond***). The H-F bond can thus be represented as:



The '+' and '-' symbols indicate ***partial*** positive and ***negative charge*** respectively.

The arrow indicates the "pull" of electrons off the hydrogen and towards the more electronegative atom, fluorine.

In **LiF** the *much greater relative* electronegativity of the fluorine atom completely strips the electron from the lithium and the result is an *ionic bond* (no sharing of the electron).

Note:The following is the general thumb rule for predicting the type of bond based upon electronegativity differences:

- If the electronegativities are equal (i.e. if the electronegativity difference is 0), the bond is **non-polar covalent**.
- If the difference in electronegativities between the two atoms is greater than 0, but less than 2.0, the bond is **polar covalent**.
- If the difference in electronegativities between the two atoms is 2.0, or greater, the bond is **ionic**.

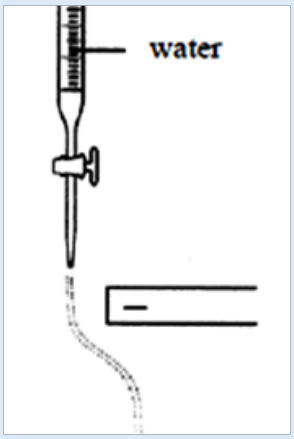
Using the examples used above, we can predict the type of bond as follows:

Compound	**F_2**	**HF**	**LiF**
Electronegativity Difference	4.0 - 4.0 = 0	4.0 - 2.1 = 1.9	4.0 - 1.0 = 3.0
According to the Rule: Type of Bond	$\Delta E=0$ Non-polar covalent	$0<\Delta E<2.0$ Polar covalent	$\Delta E\geq 2.0$ Ionic (non-covalent)

Note: A non-polar molecule is one in which the electrons are distributed more symmetrically and thus does not have an abundance of charges at the opposite sides. The charges all cancel out each other. Examples of non-polar molecules include diatomic molecules, CH4, CO2, C2H4, cyclohexene, CCl4, etc.

Application activity 19.7

1. State what is meant by the term polar bond.
2. Sulphuric acid is a liquid that can be represented by the formula drawn below.
3. The two substances CS_2 and HCN have linear molecules but CS_2 molecules are non-polar while HCN molecules are polar; explain why these molecules have different polarities. Support your explanation with appropriate diagrams.
4. A negatively charged rod was brought near a jet of water running out from a burette. The jet of water was deflected as shown:



Charged rod and jet of water interactions

a) With reference to the structure of water, explain why the jet of water was deflected.

b) State the effect on the jet of water if the negatively charged rod is replaced by a positively charged rod. Explain your answer.

c) If hexane is used instead of water and a negatively charged rod is brought near the liquid jet, would the liquid jet be deflected? Explain your answer.

19.8. Physical properties of covalent structures

Activity 19.8

1. You are provided with the following chemical compounds: PH_3, CO_2, H_2O and SiO_2, Cl_2, Br_2 and I_2 and graphite.
   a) Classify them into solid, liquid and gas.
   b) What type of bond which held together atoms in the given compounds they have in common?
   c) Deduce from **(b)** the name describing all those compounds.
   d) Some of them have very high melting points. Suggest the reason for this.
2. State some points that you know about diamond.

Covalent bonds involve the ***sharing of electrons*** so that all atoms have "*full outer shells*". Sometimes in a Covalent Bond, both shared electrons come from the same atom. This is known as a Dative Covalent Bond. This often results in the formation of charged molecules.

Covalently bonded substances fall into two main types:

- Simple molecular structures
- Giant covalent structures.

19.8.1. Simple molecular structures

Substances composed of relatively small covalently bonded structures are called Simple Molecular Structures. These contain only a few atoms held together by strong covalent bonds and can be further categorised into two types: Individual (which are usually gases like carbon dioxide) and molecular (which are usually solids like iodine).

The physical properties

1. Low melting and boiling points

   Simple Molecular Structures tend to have low melting and boiling points since the forces between molecules (intermolecular forces, which are van der Waals forces) are quite weak. Little energy is required to separate the molecules.

2. Poor electrical conductivity

   There are no charged particles (ions or electrons) delocalized throughout the molecular crystal lattice to conduct electricity. They cannot conduct electricity in either the solid or molten state.

3. Solubility

   They tend to be quite insoluble in water, but this depends on how polarized the molecule is. The more polar the molecules, the more water molecules will be attracted to them (some may dissolve in water as a result of forming hydrogen bonds with it). Molecular crystals tend to dissolve in non-polar solvents such as alcohol.

4. Soft and low density

   Van der Waals forces are weak and non-directional. The lattice is readily destroyed and the crystals are soft and have low density.

19.8.2. Giant covalent structures

Sometimes covalently bonded structures can form giant networks, known as Giant Covalent Structures or Macromolecular Structures.

Giant covalent structures contain very many atoms, each joined to adjacent atoms by covalent bonds. The atoms are usually arranged into giant regular lattices - extremely strong structures because of the many bonds involved.

Properties of giant covalent structures

- **Very high melting points**. This is because a lot of strong covalent bonds must be broken. Graphite, for example, has a melting point of more than 3,600°C.
- **Variable electrical conductivity.** Diamond does not conduct electricity, whereas graphite contains free electrons so it does conduct electricity. Silicon is a semi-conductor – it is midway between non-conductive and conductive.

1. Physical properties of diamond

   Diamond is a form (allotrope) of carbon in which each carbon atom is joined to four other carbon atoms, forming a giant covalent structure. As a result, diamond is very hard and has a high melting point. It does not conduct electricity. Diamond is tetrahedral face-centered cubic.

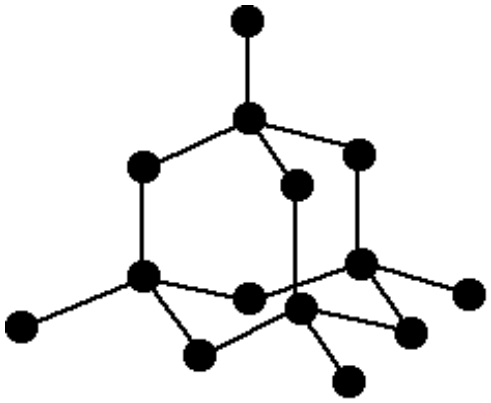


Figure 19.9.: Tetrahedral shape of diamond

The physical properties of diamond are outlined below.

- **It has a very high melting point** *(*almost 4,000°C). Very strong carbon-carbon covalent bonds have to be broken throughout the structure before melting occurs.
- **It is very hard**. This is again due to the need to break very strong covalent bonds operating in 3-dimensions.
- **It does not conduct electricity**. All the electrons are held tightly between the atoms, and are not free to move.
- **It is insoluble in water and organic solvents**. There are no possible attractions which could occur between solvent molecules and carbon atoms which could outweigh the attractions between the covalently bound carbon atoms.

2. Physical properties of graphite

   Graphite is another form (allotrope) of carbon in which the carbon atoms form layers. These layers can slide over each other, so graphite is much softer than diamond. Each carbon atom in a layer is joined to only three other carbon atoms in hexagonal rings.

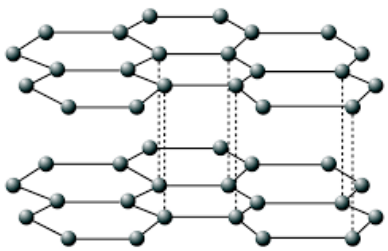


Figure 19.10: Hexagonal structure of graphite

The physical properties of graphite are oulined below.

- **It has a high melting point, similar to that of diamond**. In order to melt graphite, it isn't enough to loosen one sheet from another. You have to break the covalent bonding throughout the whole structure.
- **It has a soft, slippery feel, and is used in pencils and as a dry lubricant for things like locks**. You can think of graphite rather like a pack of cards - each card is strong, but the cards will slide over each other, or even fall off the pack altogether. When you use a pencil, sheets are rubbed off and stick to the paper.
- **It has a lower density than diamond**. This is because of the relatively large amount of space that is "wasted" between the sheets.
- **It is insoluble in water and organic solvents** - for the same reason that diamond is insoluble. Attractions between solvent molecules and carbon atoms will never be strong enough to overcome the strong covalent bonds in graphite.
- **It conducts electricity**. The delocalised electrons are free to move throughout the sheets. If a piece of graphite is connected into a circuit, electrons can fall off one end of the sheet and be replaced with new ones at the other end.

3. Physical properties of silicon dioxide

Silica, which is found in sand, has a similar structure to diamond. It is also hard and has a high melting point, but contains silicon and oxygen atoms, instead of carbon atoms. Silica or SiO2 is tetrahedral. The fact that it is a semi-conductor makes it immensely useful in the electronics industry: most transistors are made of silica.

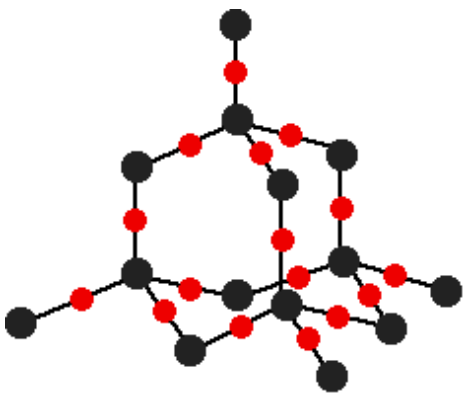


Figure 19.11: Tetrahedral shape of silicon dioxide

The physical properties of diamond are oulined below.

- **It has a high melting point** - varying depending on what the particular structure is (remember that the structure given is only one of three possible structures), but around 1700°C. Very strong silicon-oxygen covalent bonds have to be broken throughout the structure before melting occurs.
- **It is hard**. This is due to the need to break the very strong covalent bonds.
- **It does not conduct electricity**. There are not any delocalised electrons. All the electrons are held tightly between the atoms, and aren't free to move.
- **It is insoluble in water and organic solvents**. There are no possible attractions which could occur between solvent molecules and the silicon or oxygen atoms which could overcome the covalent bonds in the giant structure.

Application activity 19.8

1. What is meant by the statement that "*the electrons in diamond are localized, whereas graphite has delocalised electrons*".
2. Which electrons in graphite are delocalised? How do they affect the properties of graphite?
3. Why is it easy to rub away carbon atoms from graphite?
4. Why is graphite used as lubricant?
5. Diamond and graphite are both forms of carbon.
   - Diamond is able to scratch almost all other substances, whereas graphite may be used as a lubricant.
   - Diamond and graphite both have high melting points.

   a) Explain each of these properties of diamond and graphite in terms of structure and bonding.

   b) Give **one** other difference in the properties of diamond and graphite.
6. Iodine and diamond are both crystalline solids at room temperature.

   a) Identify one similarity in the bonding, and one difference in the structures, of these two solids.

   b) Explain why these two solids have very different melting points.

7. Silicon dioxide has a macromolecular structure. Draw a diagram to show the arrangement of atoms around a silicon atom in silicon dioxide. Give the name of the shape of this arrangement of atoms and state the bond angle.

19.9. Intermolecular forces

Activity 19.9

Experiment: ***To investigate boiling point and to determine the relation between boiling point and intermolecular forces.***

Apparatus

- Water, cooking oil (sunflower oil), Glycerine, nail polish remover, methylated spirits
- Test-tubes and a beaker
- Hot plate

Method

Methylated spirits and nail polish remover are highly flammable. They will easily catch fire if left near an open flame. For this reason they must be heated in a water bath. This experiment must be performed in a well ventilated room.

- Place about 20 mL of each substance given in separate test-tubes.
- Half-fill the beaker with water and place on the hot plate.
- Place the test-tubes in the beaker.
- Observe how long each substance takes to boil. As soon as a substance boils, remove it from the water bath.

Results and questions

1. Write down the order in which the substances boiled, starting with the substance that boiled first and ending with the substance that boiled last.
2. Suggest the explanation for the above order.

Now let us talk about the intermolecular forces that exist between molecules. Intermolecular forces are *much weaker* than the intramolecular forces of attraction but are important because they determine the physical properties of molecules like *their boiling point*, *melting point*, *density*, and *enthalpies of fusion* and *vaporization*.

19.9.1. Definition, types and origin of intermolecular forces

Intermolecular forces are the forces between molecules forces that bind them together.

Intermolecular forces are like the glue that holds molecules together. There are strong and weak forces; the stronger the force, the more energy is required to break those molecules apart from each other.

Intermolecular forces include (listed from weakest to strongest):

- Van der Waals dispersion forces
- Van der Waals dipole-dipole interactions
- Hydrogen bonding

So, if two molecules are only connected using van der Waals dispersion forces, then it would require very little energy to break those molecules apart from each other. On the other hand, if two molecules are connected using ionic bonds, it takes a whole lot more energy to break those two apart.

1. Van der Waals Dispersion Forces

   Van der Waals dispersion forces, also called ***London forces***, occur due to instantaneous dipoles. At any given moment the electrons in a molecule or atom may not be evenly distributed around the molecule. If more electrons are on the left side of the molecule than on the right side, then there will be a slight (partial) negative charge on the left side of the molecule. The side with fewer electrons will have a slight (partial) positive charge.

   These momentary, partial, positive and negative charges are attracted to each other (like the positive and negative ends on a magnet). This causes *momentary bonds* between molecules. We can already see why these bonds would be so weak, because *they only last for a little while*.

   Van der Waals dispersion forces increase as the atomic size increases. This means that larger molecules will feel more force, thus increasing the intermolecular forces. So if we have two molecules that are exactly the same except that one is bigger than the other (such as methane and ethane), then the intermolecular forces of the bigger one will be stronger than for the smaller one.

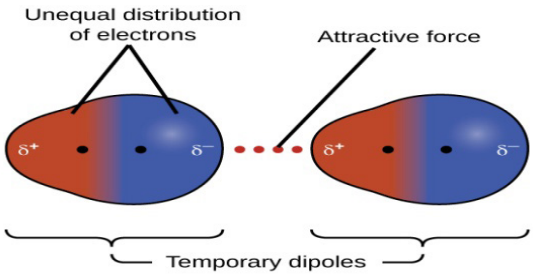


Figure 19.12: Dispersion forces result from the temporary dipoles for two non-polar diatomic molecules.

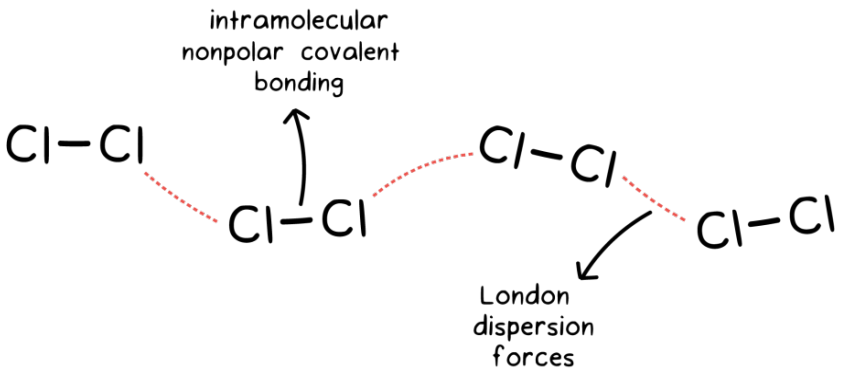


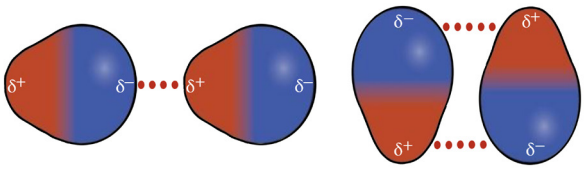
Figure 19.13.: Intramolecular non-polar covalent bonding between Cl-atoms and London dispersion forces between Cl-Cl molecules

2. Van der Waals Dipole-Dipole Interactions

A partial positive charge and a partial negative charge can be created between two atoms when there is a difference in electronegativity. These interactions are called *van der Waals dipole-dipole interactions*.

For example, carbon is less electronegative than oxygen, creating a partial positive on carbon and a partial negative on oxygen. The dipole interactions are stronger than the dispersion forces because the oxygen will almost always have slightly more electrons than the carbon, instead of constantly changing. There still is not a full negative charge on the oxygen, or a full positive charge on the carbon. But the partial positive and negative charges are still enough to attract opposite charges together.

The higher the difference in electronegativity, the strong the dipole-dipole interactions will be. So compounds with a higher electronegativity difference will have strong intermolecular forces.



This image shows two arrangements of polar molecules, such as HCl, that allow an attraction between the partial negative end of one molecule and the partial positive end of another.

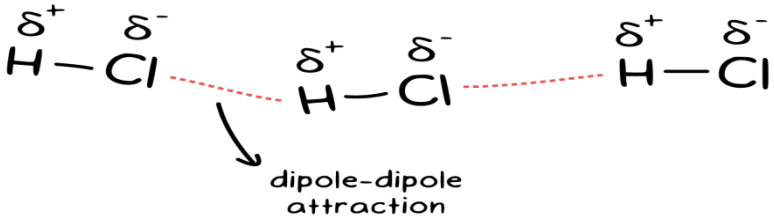


Figure 19.14: H-Cl to H-Cl dipole-dipole attraction

3. Hydrogen bonding

This is a special kind of dipole-dipole interaction that occurs between a hydrogen atom bonded to a high electronegative atom, specifically either an oxygen, nitrogen, or fluorine atom. The partially positive end of hydrogen is attracted to the partially negative end of the oxygen, nitrogen, or fluorine of another molecule. A hydrogen bond is usually represented as a dotted line between the hydrogen and the unshared electron pair of the other electronegative atom.

Hydrogen bonding is a relatively strong force of attraction between molecules, and considerable energy is required to break hydrogen bonds. This explains the exceptionally high boiling points and melting points of compounds like water and hydrogen fluoride.

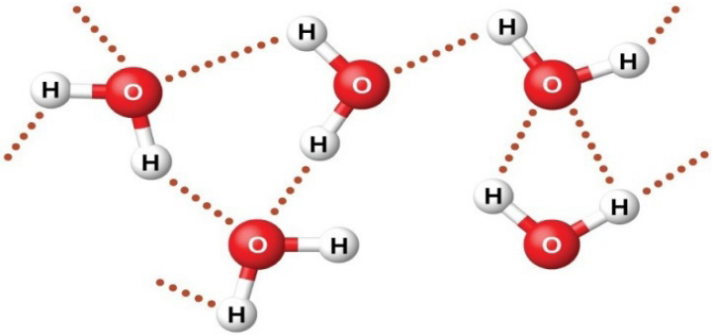


Figure 19.15: Water molecules participate in multiple hydrogen-bonding interactions with nearby water molecules

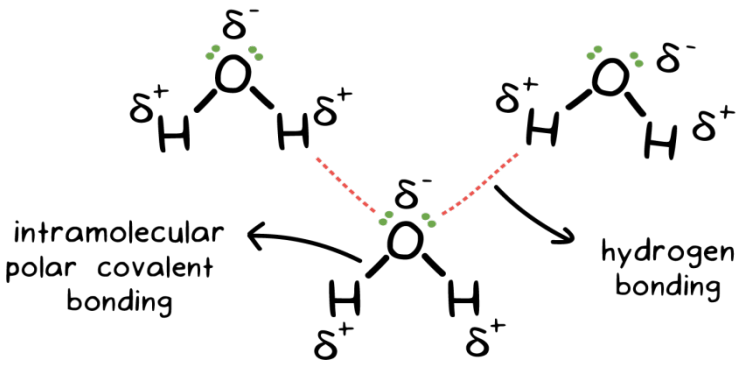


Figure 19.16: Intramolecular polar covalent bonding within H2O molecules and hydrogen bonding between O and H atoms

19.9.2. Effect of intramolecular forces on physical properties of certain molecules

Intermolecular forces control how well molecules stick together. This affects many of the measurable physical properties of substances:

- **Melting and Boiling Points**
  - If molecules stick together more, they will be tougher to break apart
  - Stronger intermolecular forces → higher melting and boiling points
- **Viscosity**
  - Viscosity is a measure of how well substances flow.
  - Stronger intermolecular forces → higher viscosity.
- **Surface Tension**
  - Surface tension is a measure of the toughness of the surface of a liquid
  - Stronger intermolecular forces → higher surface tension.
- **Vapour Pressure**
  - This is a small amount of gas that is found above all liquids.
  - Stronger intermolecular forces → Lower vapour pressure.

Note: If you are asked to rank molecules in order of melting point, boiling point, viscosity, surface tension or vapour pressure, what they are actually asking is for you to rank them by strength of intermolecular forces (either increasing or decreasing).Here is the strategy for this:

- Look for molecules with ***hydrogen bonding***. They will have the strongest intermolecular forces.

- Look for molecules with ***dipoles***. These will have the next strongest intermolecular forces.
- Larger molecules will ***have stronger London dispersion forces***. These are the weakest intermolecular forces but will often be the deciding factor in multiple choice questions.

If we use this trend to predict the boiling points for the ***lightest hydride for each group***, we would expect NH_3 to boil at about –120 °C, H_2O to boil at about –80 °C, and HF to boil at about –110 °C. However, when we measure the boiling points for these compounds, we find that they are dramatically higher than the trends would predict, as shown in the figure below. The stark contrast between our naïve predictions and reality provides compelling evidence for the strength of hydrogen bonding.

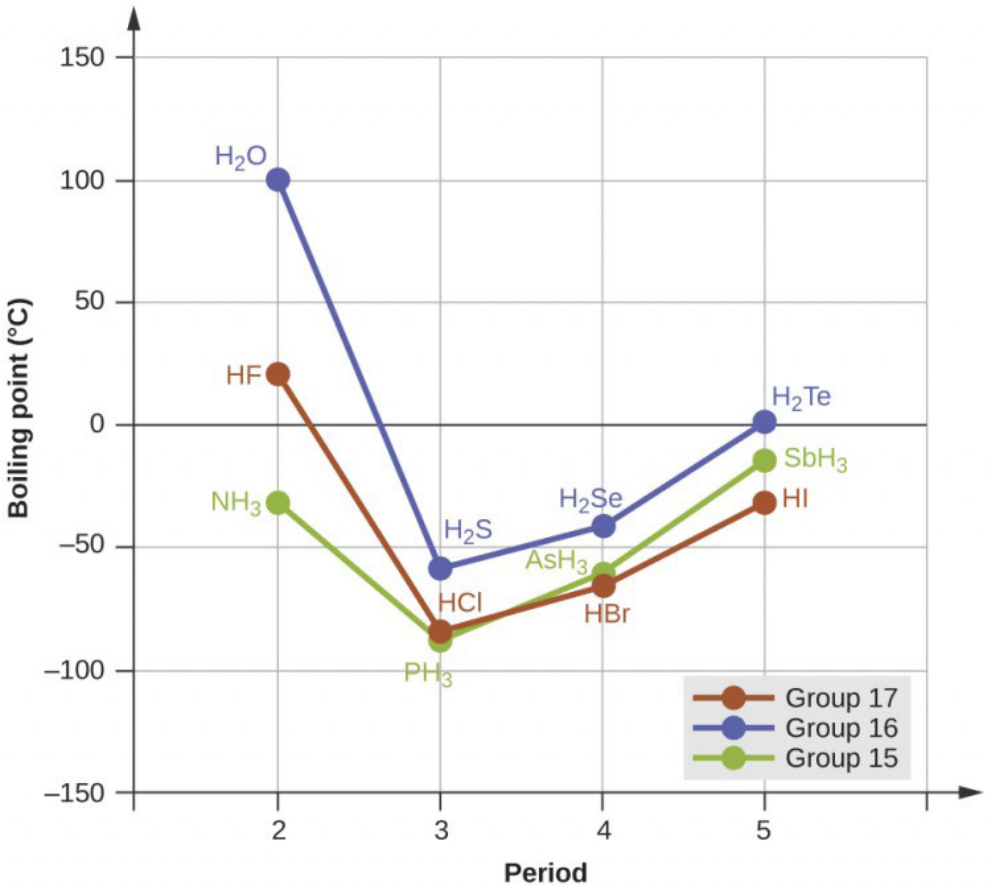


Figure 19.17: Comparison of the melting points of some hydrides of groups 17, 16 and 15

These exhibit anomalously high boiling points due to hydrogen bonding. Hydrogen bonding is important in many chemical and biological processes. It is responsible for water's unique solvent capabilities. Hydrogen bonds

hold complementary strands of DNA together, and they are responsible for determining the three-dimensional structure of folded proteins including enzymes and antibodies.

1. An Example: Water

Since oxygen is more electronegative than hydrogen, oxygen pulls the shared electrons more closely to itself. This gives the oxygen atom a slightly more negative charge than either of the hydrogen atoms. This imbalance is called a dipole, causing the water molecule to have a positive and negative side, almost like a tiny magnet. Water molecules align so the hydrogen on one molecule will face the oxygen on another molecule. This gives water a greater viscosity and also allows water to dissolve other molecules that have either a slightly positive or negative charge.

2. Protein Folding

Protein structure is partially determined by hydrogen bonding. Hydrogen bonds can occur between a hydrogen on an amine and an electronegative element, such as oxygen on another residue. As a protein folds into place, a series of hydrogen bond "zips" the molecule together, holding it in a specific three-dimensional form that gives the protein its particular function.

3. DNA

Hydrogen bonds hold complementary strands of DNA together. Nucleotides pair precisely based on the position of available hydrogen bond donors (available, slightly positive hydrogens) and hydrogen bond acceptors (electronegative oxygens). The nucleotide thymine has one donor and one acceptor site that pairs perfectly with the nucleotide adenine's complementary acceptor and donor site. Cytosine pairs perfectly with guanine through three hydrogen bonds.

Application activity 19.9

1. Define the following and give an example of each:
   a) Dispersion force
   b) Dipole-dipole attraction
   c) Hydrogen bond
2. The table below shows the boiling points of some other hydrogen halides.

	HCl	HBr	HI
Boiling point / K	188	206	238

a) Explain the trend in the boiling points of the hydrogen halides from HCl to HI.

b) Give **one** reason why the boiling point of HF is higher than that of all the other hydrogen halides.

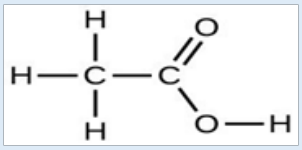
3. The types of intermolecular forces in a substance are identical whether it is a solid, a liquid, or a gas. Why then does a substance change phase from a gas to a liquid or to a solid?
4. Why do the boiling points of the noble gases increase in the order He < Ne < Ar < Kr < Xe?
5. Neon and HF have approximately the same molecular masses. Explain why the boiling points of Neon and HF differ.
6. Arrange each of the following sets of compounds in order of increasing boiling point temperature:

   a) HCl, H_2O, SiH_4

   b) F_2, Cl_2, Br_2

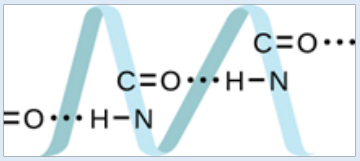
   c) CH_4, C_2H_6, C_3H_8

   d) O_2, NO, N_2
7. The molecular mass of butanol, C_4H_9OH, is 74.14; that of ethylene glycol, $CH_2(OH)CH_2OH$, is 62.08, yet their boiling points are 117.2 °C and 174 °C, respectively. Explain the reason for the difference.
8. On the basis of intermolecular attractions, explain the differences in the boiling points of *n*–butane (–1 °C) and chloroethane (12 °C), which have similar molar masses.
9. On the basis of dipole moments and/or hydrogen bonding, explain in a qualitative way the differences in the boiling points of acetone (56.2 °C) and 1-propanol (97.4 °C), which have similar molar masses.
10. The melting point of $H_2O(s)$ is 0 °C. Would you expect the melting point of $H_2S(s)$ to be –85 °C, 0 °C, or 185 °C? Explain your answer.
11. Explain why a hydrogen bond between two water molecules is weaker than a hydrogen bond between two hydrogen fluoride molecules.
12. Under certain conditions, molecules of acetic acid, CH_3COOH, form "dimers," pairs of acetic acid molecules held together by strong intermolecular attractions:



Draw a dimer of acetic acid, showing how two CH_3COOH molecules are held together, and stating the type of intermolecular force that is responsible.

13. Proteins are chains of amino acids that can form in a variety of arrangements, one of which is a helix. What kind of intermolecular force is responsible for holding the protein strand in this shape? On the protein image, show the locations of the intermolecular forces that hold the protein together:



14. Identify the intermolecular forces present in the following solids:
    a) CH_3CH_2OH
    b) CH_3CH2CH_3
    c) CH_3CH_2Cl
15. Explain why non-polar molecules usually have much lower surface tension than polar ones.

Skills lab 19

Can Water and Oil mix?

Figure 19.18.: Water and detergent

1. Pour water into a small jar that has a tight-fitting lid until the jar is about a third full.
2. Add an equal amount of vegetable oil to the jar. Cover the jar tightly.
3. Shake the jar vigorously for 20 seconds. Observe the contents.
4. Allow the jar to sit undisturbed for 1 minute. Observe again.
5. Remove the top and add 3 drops of liquid detergent.
6. Cover the jar and repeat Steps 3 and 4.

a) Based on your observations, write an operational definition of detergent.

b) How might your observations relate to chemical bonds in the detergent, oil and water molecules?

c) Demonstrate the action of soaps and detergents to your family.

d) Pour some vegetable oil on a clean cloth and show how a detergent solution can wash the oil away better than water alone can.

e) Explain to your family the features of soap and detergent molecules in terms of their chemical bonds.

End unit assessment 19

1. Which of the following statements is false about VSEPR theory?

a) The geometry of a molecule is determined by the number of electron groups on the central atom.

b) The geometry of the electron groups is determined by minimizing repulsions between them.

c) A lone pair, a single bond, a double bond, a triple bond and a single electron - each of these is counted as a single electron group.

d) Bond angles may depart from the idealized angles because lone pairs of electrons take up less space than bond pairs.

e) The number of electron groups can be determined from the Lewis structure of the molecule.

2. Choose the best answer. The correct dot formulation for nitrogen trichloride has:

a) 3 N-Cl bonds and 10 lone pairs of electrons.

b) 3 N=Cl bonds and 6 lone pairs of electrons.

c) 1 N-Cl bond, 2 N=Cl bonds and 7 lone pairs of electrons.

d) 2 N-Cl bonds, 1 N=Cl bond and 8 lone pairs of electrons.

e) 3 N-Cl bonds and 9 lone pairs of electrons.

3. Choose the molecule that is incorrectly matched with the electronic geometry about the central atom.

a) CF_4 - tetrahedral

b) $BeBr_2$ - linear

c) H2O - tetrahedral

d) NH_3 - tetrahedral

e) PF_3 - pyramidal

4. Choose the correct answer. A π(pi) bond is the result of the

a) Overlap of two s orbitals.

b) Overlap of an s and a p orbital.

c) Overlap of two p orbitals along their axes.

d) Sidewise overlap of two parallel p orbitals.

e) Sidewise overlap of two s orbitals.

5. Choose the missing answer. The F-S-F bond angles in SF_6 are _____.

a) 109°28'

b) 120° only

c) 90° and 120°

6. Draw a complete line-bond or electron-dot formula for acetic acid and then decide which statement is incorrect.

a) One carbon is described by sp2 hybridization.

b) The molecule contains only one πbond.

c) The molecule contains four lone pairs of valence electrons.

d) One carbon is described by sp3 hybridization.

e) Both oxygens are described by sp3 hybridization.

7. The equation below shows the reaction between boron trifluoride and a fluoride ion. $BF_3 + F^- \rightarrow BF_4^-$

a) Draw diagrams to show the shape of the BF_3 molecule and the shape of the BF_4^- ion. In each case, name the shape. Account for the shape of the BF4– ion and state the bond angle present.

b) In terms of the electrons involved, explain how the bond between the BF_3 molecule and the F^- ion is formed. Name the type of bond formed in this reaction.

8. Draw the shape of a molecule of $BeCl_2$ and the shape of a molecule of Cl_2O. Show any lone pairs of electrons on the central atom. Name the shape of each molecule.

9. Ammonia, NH_3, reacts with sodium to form sodium amide, $NaNH_2$, and hydrogen.

   a) Draw the shape of an ammonia molecule and that of an amide ion, NH_2^-. In each case show any lone pairs of electrons.

   b) State the bond angle found in an ammonia molecule.

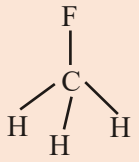
   c) Explain why the bond angle in an amide ion is smaller than that in an ammonia molecule.

10. The table below shows the electronegativity values of some elements.

	Fluorine	Chlorine	Bromine	Iodine	Carbon	Hydrogen
Electronegativity	4.0	3.0	2.8	2.5	2.5	2.1

   a) Define the term electronegativity.

   b) The table below shows the boiling points of fluorine, fluoromethane (CH_3F) and hydrogen fluoride.

	F–F		H–F
Boiling point/K	85	194	293

   Name the strongest type of intermolecular force present in:

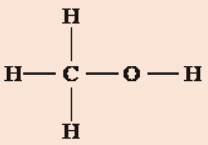
   i. F_2

   ii. CH_3F

   iii. HF

   c) Explain how the strongest type of intermolecular force in liquid HF arises.

11. Methanol has the structure



a) Explain why the O–H bond in a methanol molecule is polar.

b) The boiling point of methanol is +65 °C; the boiling point of oxygen is –183 °C. Methanol and oxygen each have a Mr value of 32. Explain, in terms of the intermolecular forces present in each case, why the boiling point of methanol is much higher than that of oxygen.

12. The table below gives the boiling points, T_b, of some hydrogen halides.

Hydrogen halide	HF	HCl	HBr	HI
T_b /K	293	188	206	238

a) By referring to the types of intermolecular force involved, explain why energy must be supplied in order to boil liquid hydrogen chloride.

b) Explain why the boiling point of hydrogen bromide lies between those of hydrogen chloride and hydrogen iodide.

c) Explain why the boiling point of hydrogen fluoride is higher than that of hydrogen chloride.

d) Draw a sketch to illustrate how two molecules of hydrogen fluoride interact in liquid hydrogen fluoride.

13. Figure below shows some data concerned with halogens.

Element	Electronegativity	Boiling point of hydride / K
Fluorine	4.0	293
Chlorine	3.0	188
Bromine	2.8	206
Iodine	2.5	238

a) Define the term *electronegativity.*

b) Explain the trend in boiling points from hydrogen chloride to hydrogen iodide.

c) Explain why hydrogen fluoride does not fit this trend.

14. The oxygen atoms in the sulphate ion surround the sulphur in a regular tetrahedral shape.

a) Write the formula of the ion.

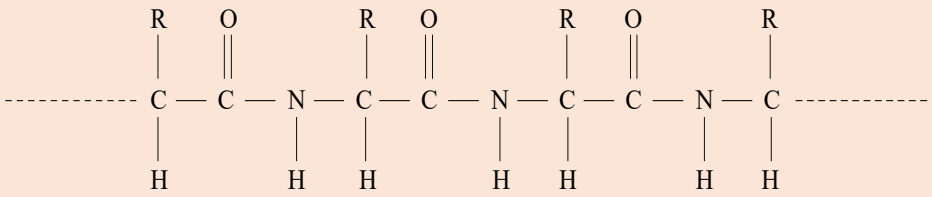
b) State the O–S–O. bond angle.

15. (a) State **one** feature which molecules must have in order for hydrogen bonding to occur between them.

b) Give the name of the type of intermolecular bonding present in hydrogen sulphide, H_2S, and explain why hydrogen bonding does not occur.

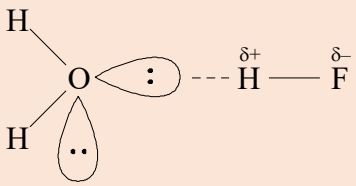
c) Account for the much lower boiling point of H_2S (–61 °C) compared with that of water (100 °C).

16. Protein molecules are composed of sequences of amino acid molecules that have joined together, with the elimination of water, to form long chains. Part of a protein chain is represented by the graphical formula given below.



Explain the formation of hydrogen bonding between protein molecules.

17. The diagram below shows how a water molecule interacts with a hydrogen fluoride molecule.



a) What is the value of the bond angle in a single molecule of water?

b) Explain your answer to part **(a)** by using the concept of electron pair repulsion.

c) Name the type of interaction between a water molecule and a hydrogen fluoride molecule shown in the diagram above.

d) Explain the origin of the **+** charge shown on the hydrogen atom in the diagram.

e) When water interacts with hydrogen fluoride, the value of the bond angle in water changes slightly. Predict how the angle is different from that in a single molecule of water and explain your answer.

18. Phosphorus exists in several different forms, two of which are white phosphorus and red phosphorus. White phosphorus consists of P_4molecules, and melts at 44°C. Red phosphorus is macromolecular, and has a melting point above 550°C.

a) Explain what is meant by the term *macromolecular*.

b) By considering the structure and bonding present in these two forms of phosphorus, explain why their melting points are so different.

UNIT 20:

MOMENTS AND EQUILIBRIUM OF BODIES

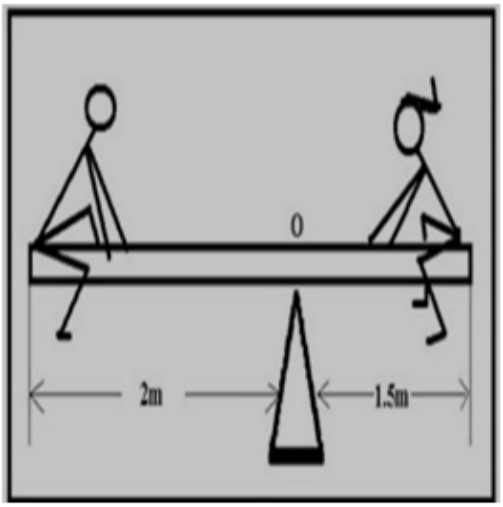
Key Unit Competence: Explain the principle of moments and apply it to the equilibrium of a body.

Introductory Activity 20

Study the diagrams (a), (b) and (c) below:

(a)

(b)



(c)

1. Referring to pictures (a) and (b), name the tools used.
2. What happens when force is applied:
   i. At the end of the tool used?
   ii. In the middle of the tool used?
3. Referring to (c) that shows two children balancing on a seesaw.
   i. If they have different weights can they balance? Explain
   ii. If they have the same weight do they balance? Explain

20.1. Vector and scalar quantities

Activity 20.1

Task 1

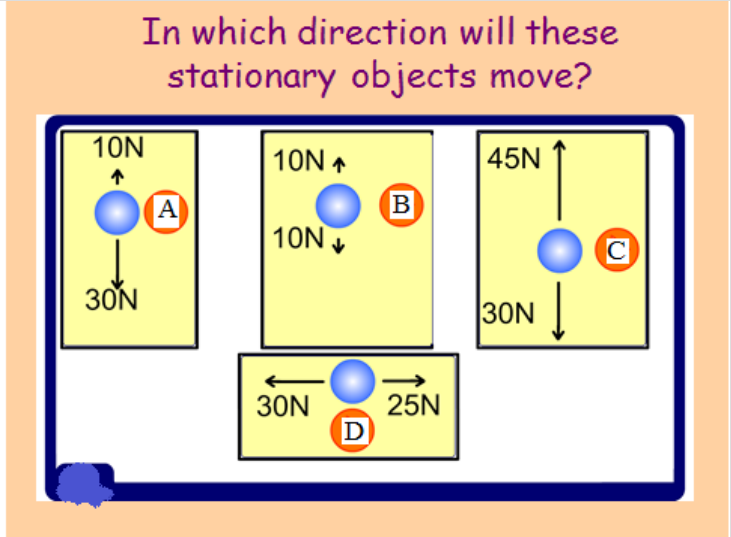
Try to stand in a line behind one another. Push your friend.

i. What happens to your friend?

ii. How do you feel?

iii. What if in the process one stops pushing, what would happen?

Task 2



Most quantities measured in science (Physics) are classified as either Scalar or vector.

A scalar quantity is a physical quantity that is defined by only magnitude (size).

Examples of scalar quantities are volume, mass, speed, and time intervals. The rules of ordinary arithmetic are used to manipulate scalar quantities.

A vector quantity is a quantity with both magnitude and direction.

Example of vector quantity: velocity, acceleration, force, weight, electric field, displacement and pressure.

20.1.1. Force as vector and moment of a force about a point

Force as vector

The force is vector quantity. We can think of force as that which causes an object to accelerate. What happens when several forces act simultaneously on an object? In this case, the object accelerates only if the net force acting on it is not equal to zero.

The net force acting on an object is defined as the vector sum of all forces acting on the object. (We sometimes refer to the net force as the total force, the resultant force, or the unbalanced force.)

Moment of a force or torque

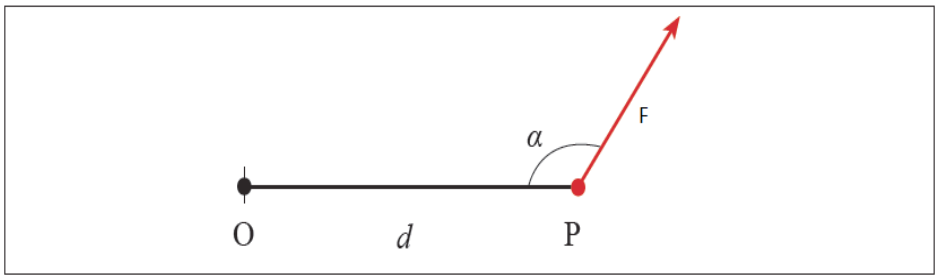


Figure 20.1: Force $\vec{F}$ acting on a body that is free to rotate about an axis

This figure shows a force $\vec{F}$ acting on a body that is free to rotate about an axis. The force is applied at the point P whose position is defined by the vector

The direction of $\vec{F}$ and make an angle α with each other.

We define the torque acting on the body from:

$\tau = F \times d \sin\alpha$

The perpendicular distance of the line of action of the force from the axis of rotation is called the moment arm of the force.

The S.I unit of torque is Newton-metre [N m] or metre-newton [m N]

20.1.2. Principles of moment

The principle of moments states that" when in equilibrium the total sum of the anticlockwise moment is equal to the total sum of the clockwise moment." When a system is stable or balance it is said to be in equilibrium as all the forces acting on the system cancel each other out.

In equilibrium:

$$Total\ anticlockwise\ moment = Total\ clockwise\ moment$$

The principle can be explained by considering two people on a seesaw.

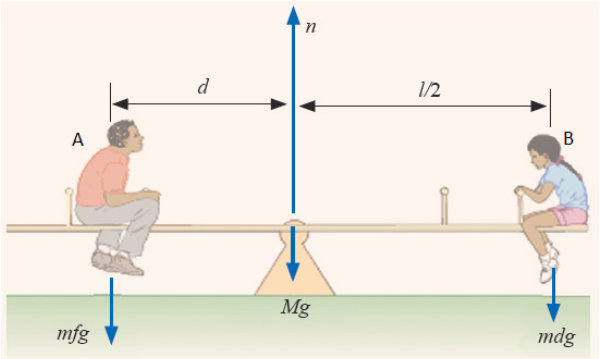


Figure 20.2: Balancing on a seesaw

Moments acting on a seesaw

Both people exert a downward force on the seesaw due to their weights.

Person A's weight is trying to turn the seesaw anticlockwise with person B's weight is trying to turn the seesaw clockwise.

Person A's moment = $force \times perpendicular\ distance\ from\ fulcrum$

$$mfg \times d$$

Person B's moment = $force \times perpendicular\ distance\ from\ fulcrum$

$$mdg \times \frac{1}{2}$$

$$Person\ A's\ moment = Person\ B's\ moment$$

$$Anticlockwise\ moment = clockwise\ moment$$

$$mfg \times d = mdg \times \frac{1}{2}$$

Therefore the seesaw is in equilibrium

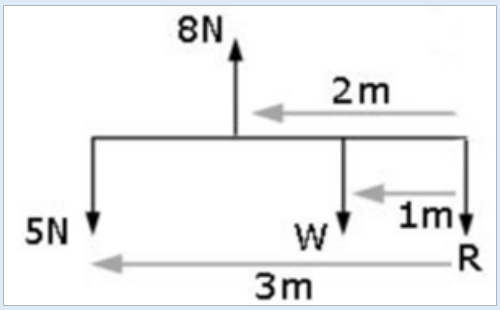
Application activity 20.1

1. Complete the following paragraph using the following words: forces, unbalanced , accelerating, slower, balanced, resultant, faster, stay still, and decelerating.

   A ball will stay still if the forces on it are _____. If the forces on it are unbalanced, the ball will get _____ or _____. The overall force is called the _____ force. If something gets faster, we say it is _____.

2. A beam of negligible weight is horizontal in equilibrium, with forces acting upon it, as shown in the diagram.

   Calculate the value of the weights R and W.



20.2. Types of equilibrium: stable, unstable and neutral

Activity 20.2

i. Displace the desk. What happens when you withdraw the force you applied?

ii. Place a bottle on a table so that it rests on its horizontal surface. Displace or roll it. What happens?

iii. Place a knife edge on a table resting on its tip. Give it a small displacement. What happens to it?

iv. From the observations made, how do you conclude?

Equilibrium has many different meanings, depending on what subject (chemistry or physics) or what topic (energy or forces). Dealing with energy, there are three types of equilibrium.

A body is in either Stable, Unstable or in neutral equilibrium depending how it behaves when subjected to a small displacement.

Stable is when any sort of movement will raise the object's centre of gravity. When objects in stable equilibrium are moved, they have a tendency to fall back to their original position. For instance, a skateboarder at the bottom, in the middle, of a ramp. Either way the skateboarder moves, his/her potential energy will increase because he/she will be rising in height. The boarder will also roll back to the bottom of the ramp if he/she doesn't exert any sort of energy to maintain the new position.

When a body returns to its original position on being slightly disturbed, it is said to be in stable equilibrium.

Unstableis when any sort of movement will lower the object's centre of gravity. When such objects are moved, they cannot return to their original position without some exertion of energy. For instance, when a coin is placed on its side, it exhibits unstable equilibrium. Any sort of push will cause the coin to fall flat, lowering its centre of mass. The coin will not return to its side unless someone picks it up and resets it.

If the position of a body is disturbed and the body does not return to its original position, it is in unstable equilibrium.

Neutral is when any sort of movement does not affect the object's centre of gravity. For instance, a ball on a table exhibits neutral equilibrium. If the ball rolls, the centre of mass stays at the same height and thus it maintains the same equilibrium.

A body is said to be in neutral equilibrium if it moves to a new position when it is disturbed.

Let us consider a cone to understand these states.

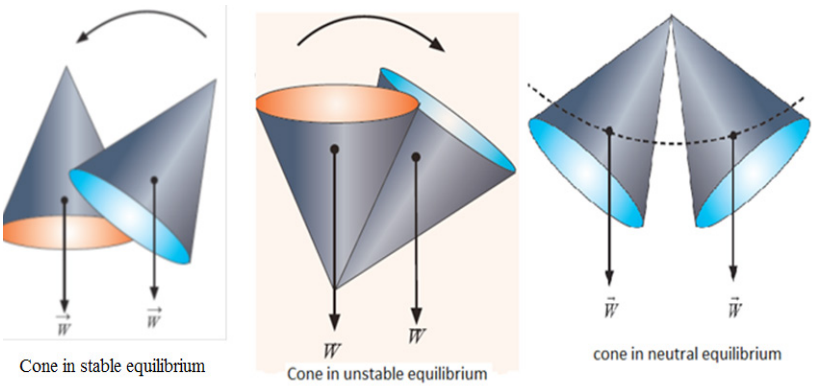


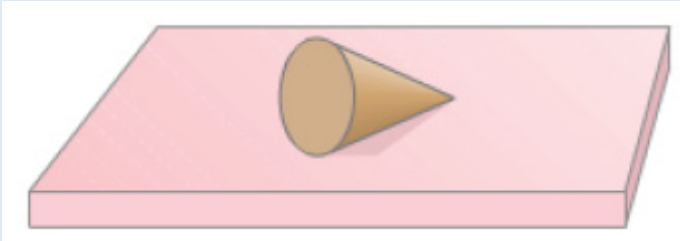
Figure 20.3: A cone in stable, unstable and neutral equilibrium.

From whatever we have done, we can conclude and say that a body is stable when:

1. The object's base is broad.
2. The centre of gravity is as low as possible.
3. The vertical line drawn from the centre of gravity should fall within the base. Lowering the centre of gravity of an object is important for stability.

Application activity 20.2

Figure below shows a cone. Explain how to lay it on a flat table so that it is in (a) stable equilibrium, (b) unstable equilibrium, (c) neutral equilibrium.



20.3. Condition for equilibrium of a body about an axis and Stevinus proof

Activity 20.3

As a class, let us move outside. With the help of a teacher outside the class:

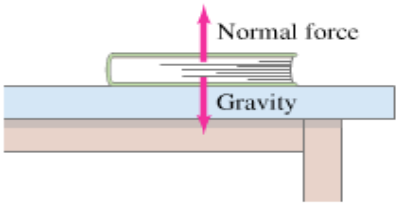
a). Are you seeing any object outside?

b). Are they stationary or in motion?

c). If at rest, what causes them to be at rest?

20.3.1. Condition for equilibrium of a body about an axis

Objects in daily life have at least one force acting on them (gravity). If they are at rest, then there must be other forces acting on them as well so that the net force is zero. A book at rest on a table, for example, has two forces acting on it, the downward force of gravity and the normal force the table exerts upward on it.



The book is in equilibrium: the net force on it is equal to zero.

Since the net force on the book is zero, the upward force exerted by the table on the book must be equal in magnitude to the force of gravity acting downward on the book. Such an object is said to be in equilibrium (Latin for "equal forces" or "balance") under the action of these two forces. The two conditions for equilibrium of a rigid object under the action of coplanar forces are:

1. The first or force condition: the vector sum of all forces acting on the body must be zero: Where the plane of the coplanar forces is taken to be the xy-plane. We must remember that if a particular force component points along the negative x or y axis, it must have a negative sign.

$$\sum F_x = 0 \qquad \sum F_y = 0$$

2. The second or torque condition: take an axis perpendicular to the plane of the coplanar forces. Call the torques that tend to cause clockwise rotation about the axis negative, and counterclockwise torques positive; then the sum of all the torques acting on the object must be zero:

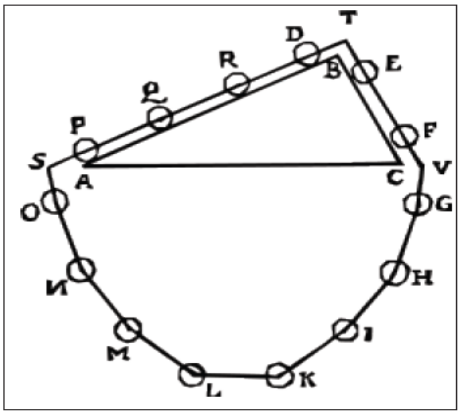
$$\sum \tau = 0$$

20.3.2. Stevinus proof

Stevinus (sometimes called Stevin) proof of the law of equilibrium on an inclined plane known as the "Epitaph of Stevinus".

He derived the condition for the balance of forces on inclined planes using a diagram with a "wreath" containing evenly spaced round masses resting on the planes of a triangular prism (see the illustration on the figure). He concluded that the weights required were proportional to the lengths of the sides on which they rested assuming the third side was horizontal and that the effect of a weight was reduced in a similar manner.

It's implicit that the reduction factor is the height of the triangle divided by the side (the sine of the angle of the side with respect to the horizontal). The proof diagram of this concept is known as the "Epitaph of Stevinus".



20.3.3. Free –Body Diagrams (FBD)

In physics and engineering, a free body diagram (force diagram, or FBD) is a graphical illustration used to visualize the applied forces, movements, and resulting reactions on a body in a given condition. They depict a body or connected bodies with all the applied forces and moments, and reactions, which act on the body (ies).

In educational environment, learning to draw a free body diagram is an important step in understanding certain topics in physics, such as statics, dynamics and other forms of classical mechanics.

Free body diagrams consist of:

- A simplified version of body (often a dot or a box)
- Forces shown as a straight arrows pointing in the direction they act on the body
- Moments shown as curved arrows pointing in the direction they act on the body
- Coordinate system
- Frequently reaction to applied forces are shown with hash marks through the stem of the arrow.

A number of forces and moments shown in a free body diagram depends on the specific problem and the assumptions made; common assumptions are neglecting air resistance, friction and assuming rigid bodies.

In statics all forces and moments must balance to zero; the physical interpretation of this is that if the forces and moments do not sum to zero the body is accelerating and the principle of statics do not apply. In dynamics the resultant forces and moments can be non-zero.

Example: Motion on a horizontal plane with frictional force

FBD

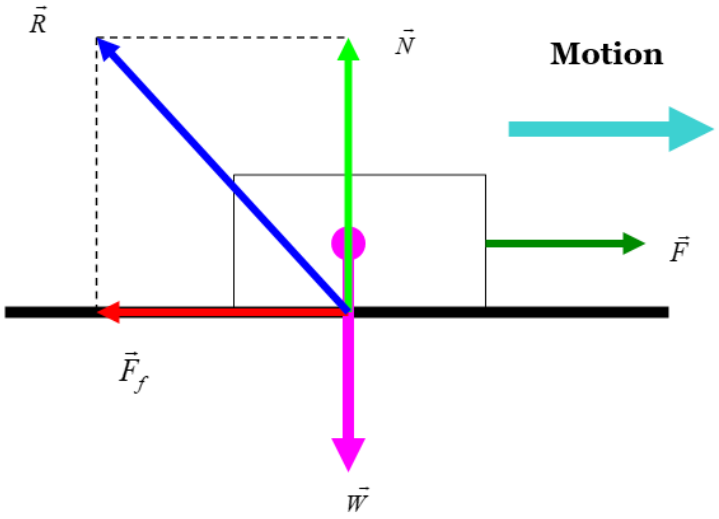


Figure 20.4: Free-body diagram of an object on horizontal plane

Example 20.1: Forces on a beam and supports.

A uniform 1500-kg beam, 20.0 m long, supports a 15,000-kg printing press 5.0 m from the right support column (Figure below). Calculate the force on each of the vertical support columns.

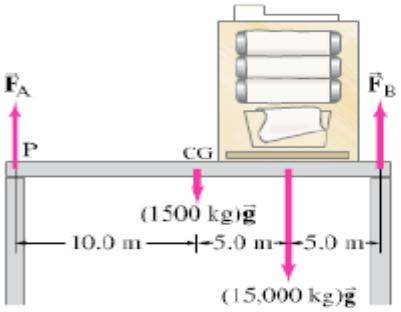


Figure: A 1500-kg beam supports a 15,000-kg machine.

APPROACH

We analyze the forces on a beam (the force the beam exerts on each column is equal and opposite to the force exerted by the column on the beam). We label these forces and in the figure. The weight of the beam itself acts at its center of gravity, 10.0 m from either end. We choose a convenient axis for writing the torque equation: the point of application of (labeled P), so will no enter the equation (its lever arm will be zero) and we will have an equation in only one unknown, .

SOLUTION

The torque equation, with the counterclockwise direction as positive gives

$\sum \tau = -(10.0\ m)(1500\ kg)g - (15.0\ m)(15{,}000\ kg)g + (20.0\ m)\ F_B = 0$

Solving for F_B, we find $F_B = (12{,}000\ kg)g = 118{,}000\ N$. To find F_A, we use ? $F_y = 0$, with + y upward:

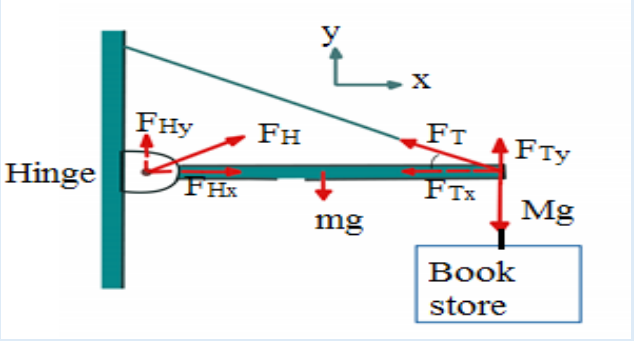
$\sum F_y = F_A - (1500\ kg)g - (15{,}000\ kg)g + F_B = 0$

Putting in $F_B = (12{,}000\ kg)g$ we find that $F_A = (4500\ kg)g = 44{,}100\ N$

Our next example involves a beam that is attached to a wall by a hinge and is supported by a cable or cord. It is important to remember that flexible cable can support a force only along its length. (if there were a component of force perpendicular to the cable, it would bend because it is flexible.) But for a rigid device, such as the hinge in figure below, the force can be in any direction and we can know the direction only after solving the problem. The hinge is assumed small and smooth, so it can exert no internal torque (about its center) on the beam.

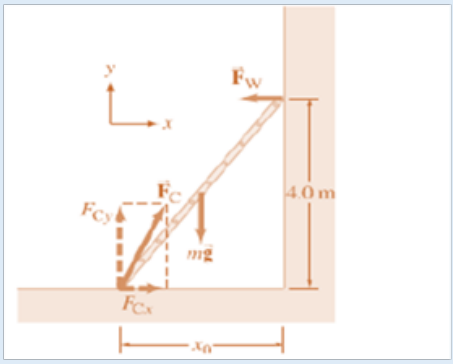
Application activity 20.3

1. A uniform beam, 2.20 m long with mass **m=25.0 kg**, is mounted by a hinge on a wall as shown in figure below. The beam is held in horizontal position by a cable that makes an angle ?=30.0^{0}as shown. The beam supports a sign of mass **M=28.0 kg** suspended from its end.



Determine the components of the force ($\vec{F}_H$) that the hinge exerts on the beam, and the tension F_T in supporting cable.

2. A 5.0-m-long ladder leans against a wall at a point 4.0 m above a cement floor as shown in figure below. The ladder is uniform and has mass m=12.0 kg. Assuming the wall is frictionless (but the floor is not), determine the forces exerted on the ladder by the floor and by the wall.



20.4. Forces and moments in equilibrium

Activity 20.4

- Try to lift your seat alone.
- How do you feel?
- Tell your friend to help you and you lift it together
- Are you feeling the same way as before when you were alone?
- What if your friend pulls in opposite direction to that of your force? What would happen?

20.4.1. Forces in equilibrium

When all the forces that act upon an object are balanced, then the object is said to be in a state of equilibrium. The forces are considered to be balanced if the rightward forces are balanced by leftward forces and the upward forces are balanced by the downward forces. This however does not necessarily mean that all the forces are equal to each other.

There are two cases of equilibrium that are often encountered; the first case deals with an object subjected to only two forces, and the second case is concerned with an object subjected to three force

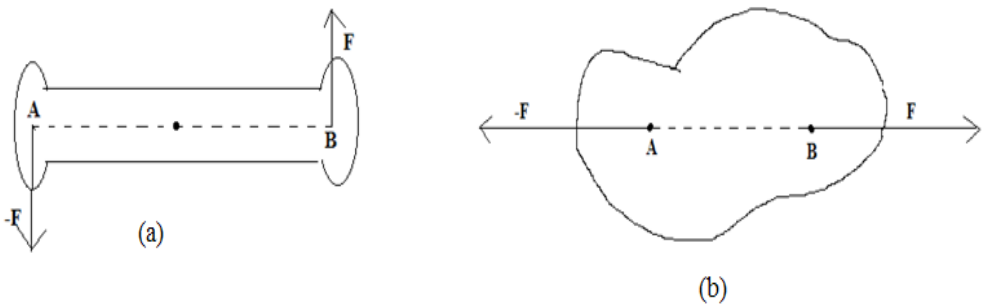


Figure 20.5 (a) The object is not in equilibrium since two forces do not have the same line of action; (b) the object is in equilibrium since two forces act along the same line.

Case I: if an object is subjected to two forces, the object is in equilibrium if and only if the two forces are equal in magnitude, opposite in direction and have the same line of action.

Figure (a) shows a situation in which the object is not in equilibrium because the two forces are not along the same line. Note that the torque about any axis such as one through A is not zero. Which violet the second condition of equilibrium.

Figure (b), the object is in equilibrium because the forces have the same line of action. In this situation, it is easy to see that the net torque about any axis is zero.

Case II: If an object subjected to three forces is in equilibrium, the line of action of the three forces must intersect at a common point; that is the force must be concurrent. One exception to this rule is the situation in which none of the lines of action intersect. In this situation, the forces must be parallel.

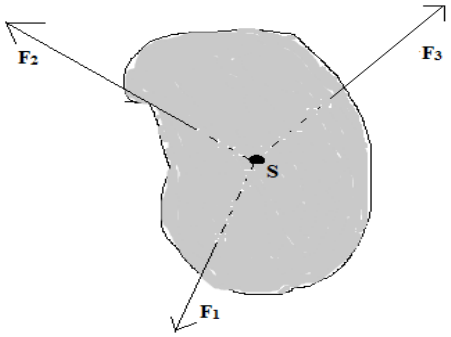


Figure 20.6: Concurrent forces

If three forces act on an object that is in equilibrium, their lines of action must intersect at a point S (or they must be parallel).

20.4.2. Couples and Torques

There are many examples in practice where two forces, acting together, exert a moment or turning effect on some object. As a very simple case, suppose two strings are tied to a wheel at X,Y and two equal and opposite forces, F, are exerted tangentially to the wheel, figure below. If the wheel is pivoted at its centre, O, it begins to rotate about O in anticlockwise direction. The total moment about O is then

$$(F \times OY) + (F \times OX) = F \times XY$$

Two equal and opposite forces whose lines of action do not coincide are said to form a **couple.** The two forces always has a turning effect, or moment, called a **torque**, which is given by

since XY is perpendicular to each of the forces F in figure below,

The moving coil or armature in electric motor is made to spin round by a couple. When it is working, two parallel and equal forces, in opposite directions, act on opposite sides of the coil and produce a torque on it.

In figure(ii), a wheel W is rotated about its center O by a tangential force F at A on one side. To find the whole effect of F on the wheel, we can put two other forces F at O which are parallel to F at A but opposite in direction. This does not disturb the mechanics because the two forces at O would cancel and leave F at A. But the upward force F at O and the opposite parallel force F at A form a couple of moment . This is actually the moment of F at A.

In addition, however, we are left with the downward force at O. so when F is applied to A to turn the wheel, it produces a couple plus an equal force F at O, the centre of the wheel.

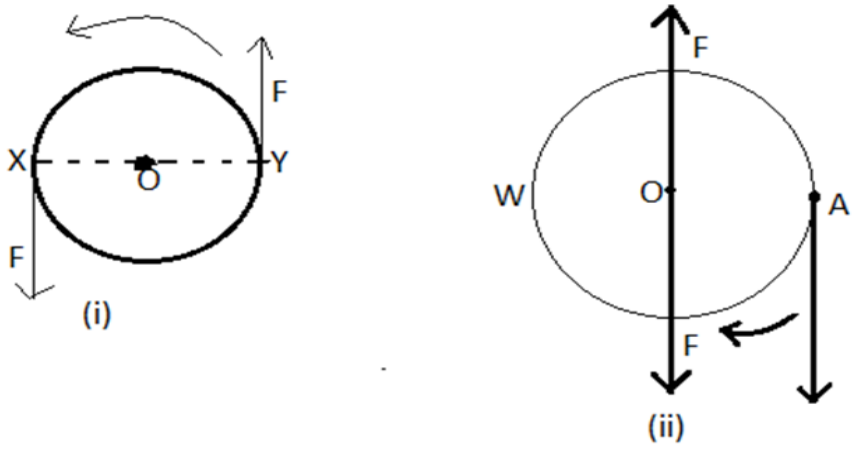


Figure 20.7: Couples and moments

20.4.3. Equilibrium of Coplanar forces

Forces in equilibrium mean that they are balanced.

Coplanar forces act in the same plane. Two balanced forces are equal in magnitude but opposite in direction to the other.

Parallelogram of forces

A force is a vector quantity. So it can be represented in size and direction by a straight line drawn to scale. The sum or resultant of two forces and can be added by one of two vector methods.

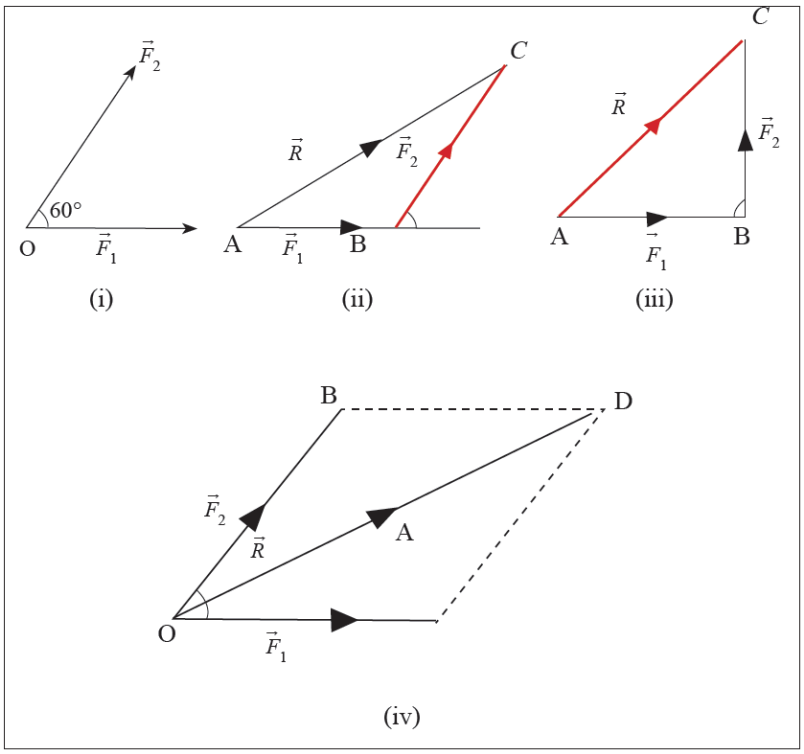


Figure 20.8: Vector addition of forces

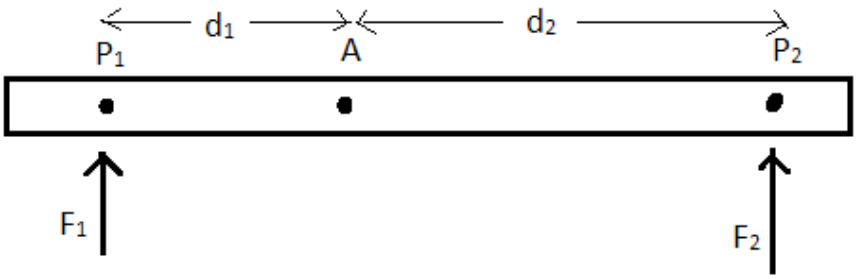
If three forces are in equilibrium, they can be represented in magnitude and direction by the three sides of a triangle taken in order. This is called the triangle of forces theorem. This result can be extended to many forces in equilibrium.

20.4.4. Equilibrium of moments of force

We consider the equilibrium of an object like a horizontal bar. In the lab it will be a meter stick. Equilibrium means that it does not have translation (motion in which all points on the body move with the same vector velocity) or rotation.

We can define the rotation by choosing any point on the body, calling that point the “axis”, and considering rotation about that axis.

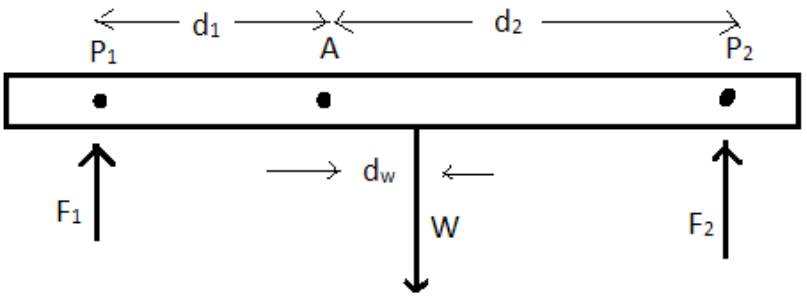
There may be several forces acting on the body, and each force acts at certain point. In the diagram, acts at point and acts at point; A is the axis.



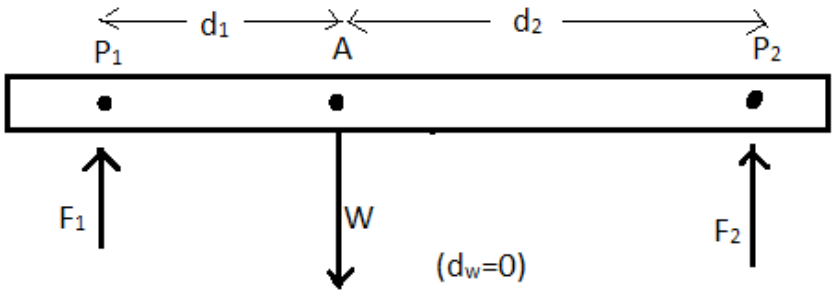
We consider only forces that act up or down. The distance from point where the force acts to the axis is called the moment arm, and in the diagram. The product of the force times the moment arm is called the torque (also called the moment), and is represented by a Greek letter tau

Torque also has a sign: it is positive (by convention) if it tends to rotate the object in counterclockwise direction around the axis. It is negative if it tends to rotate the object in a clockwise direction around the axis. In the diagram the torque due to is positive and the torque to is negative. The forces themselves have signs. We may take forces positive if directed up, and negative down.

The conditions for equilibrium are then stated simply: the sum of all forces must be zero, and the sum of all torques must be zero. One particular force requires further consideration: the weight of the bar. The weight (force of earth’s gravity) does not act at a point, but acts at all points along the object. However, for the purposes of determining equilibrium, the weight can be considered to be concentrated at a single point, called the center of mass. Thus a bar supported by two upward forces, and held in equilibrium, looks like the figure below.



Since the axis, for problem-solving purposes, can be chosen anywhere along the bar, it is convenient to choose it at the point where one of the forces acts. Then the moment arm of that force will be equal to zero, and so that force will not come into torque equation. For example, we may choose the axis to be at the center of mass.



In this diagram, suppose F_1=20 N,d_1=0.4 m,d_2=0.5 m, and F_2 is unknown. W is also not known, but , and so the torque due to the weight . The torque equation thus becomes,

$$\tau_1+\tau_2=0$$

$$\tau_1=-F_1\, d_1=-2.0\ \text{N}\times 0.4\ \text{m}=-0.8\ \text{Nm}$$

$$\tau_2=+F_2\, d_2=+F_2\times 0.5\ \text{m}$$

And so,

$$+F_2\times 0.5\ \text{m}-0.8\text{N m}=0$$

$$F_2=1.6\ \text{N}$$

Note that the weight is not given, or determined, in the discussion above. To find the weight we have to use the second equilibrium condition, that the sum of all forces is zero. Since F_1 and F_2 are up, and W is down, we have

$$2.0\ \text{N}+1.6\ \text{N}-W=0 \rightarrow W=3.6\ \text{N}$$

Application activity 20.4

Determine the mass of the object using the principle of moments.

Apparatus required

- 1 uniform meter rule
- A knife edge
- Unknown mass

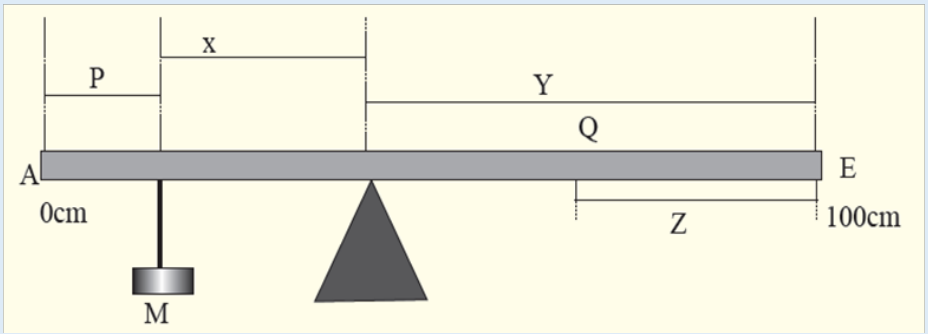


Figure 20.9: A balancing object

Instructions

Follow these steps to determine the mass *M* of the provided body (meter ruler).

a) Weigh the meter ruler provided to obtain its mass, *M.* Balance the meter ruler on a knife edge. Read Q the balance point. Find Z.

b) Balance the meter ruler with its graduated face upwards on a knife edge. With the mass *M* provided at P=10 cm from A as shown,

c) Measure and record distance X and Y.

d) Repeat procedure (b) and (c) for values of P=15, 20, 25, 30 and 35cm.

e) Tabulate your results including values of (Y-Z) and (X-P).

f) Plot a graph of (Y-Z) against (X-P).

g) Find the slope "*S*" of your graph.

20.5. Archimedes' principle of the lever and centre of gravity of bodies

Activity 20.5

Task 1

1. In group, discuss levers using knowledge of O' level Machines.
2. What is your friend telling you? Is it what you knew before?

Task 2

Given a pencil, a ruler, a notebook or other regular shaped material

Tray to balance the material using your finger. Is it balanced? What causes it to be balanced? How do you call the point of balance? How can you locate this point?

20.5.1. Archimedes and the principle of the lever

Building up from the earliest remaining writings regarding levers date from the 3rd century BC and were provided by Archimedes. "Give me a place tostand, and I shall move the Earth with it" is a remark of Archimedes who formally stated the correct mathematical principle of lever.

Force and levers, law of the lever

A lever is a beam connected to ground by a hinge or pivot called a fulcrum.

In other words we say that the lever is a movable bar that pivots on a fulcrum attached to a fixed point. The lever operates by applying forces at different distances from the fulcrum, or a pivot.

The ideal lever does not dissipate or store energy, which means there is no friction in the hinge or bending in the beam. In this case, the power into the lever equals the power out, and the ratio of output to input force is given by the ratio of the distances from the fulcrum to the points of application of these forces. This is known as the *law of the lever.*

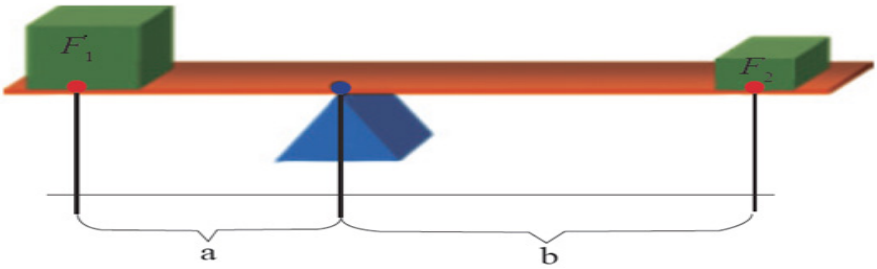


Figure 20.10: A lever in balance

The mechanical advantage of a lever can be determined by considering the balance of moments or torque, τ, about the fulcrum, and we write:

$$MA = \frac{F_2}{F_1} = \frac{a}{b}$$

where F_1 is the input force to the lever and F_2 is the output force. Then distances a and b are the perpendicular distances between the forces and the fulcrum.

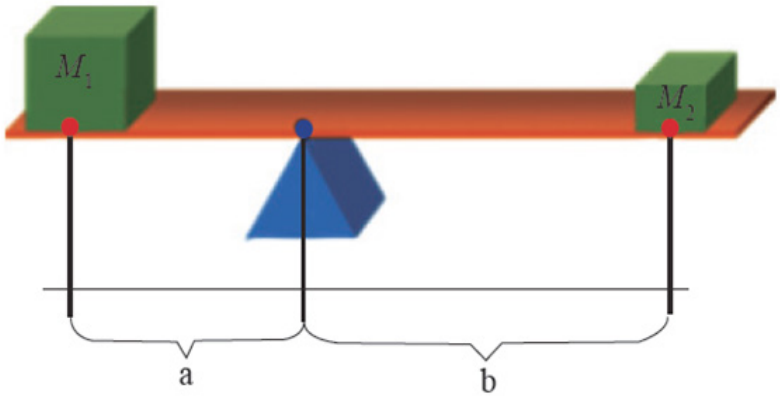


Figure 20.11: A lever in balance consisting masses

For the case of the figure 20.11, knowing that the forces considered are weights of objects and $W = Mg$, we can write the mechanical advantage of the lever as a ratio,

$$MA = \frac{M_2}{M_1} = \frac{a}{b}$$

$$W = 3.6\,N$$

20.5.2. Centre of gravity and the total weight centre of gravity of a flat object

In most equilibrium problems, one of the forces acting on the body is the weight. We need to be able to calculate the torque of this force. The weight doesn't act at a single point; it is distributed over the entire body. But we can always calculate the torque due to the body's weight by assuming that the entire force of gravity (weight) is concentrated at a point called the center of gravity (abbreviated " cg"). The acceleration due to gravity decreases with altitude; but if we can ignore this variation over the vertical dimension of the body, then the body's center of gravity is identical to its center of mass (abbreviated "cm"), let us prove it.

First let's review the definition of the center of mass. For a collection of particles with masses m_1, m_2,and coordinates $(x_1,y_1,z_1)(x_2,y_2,z_2)$ the coordinates and of the center of mass are given by

$$x_{cm} = \frac{m_1x_1 + m_2x_2 + m_3x_3 + \ldots}{m_1 + m_2 + m_3} = \frac{\sum_i m_i x_i}{\sum_i m_i}$$

$$y_{cm} = \frac{m_1y_1 + m_2y_2 + m_3y_3 + \ldots}{m_1 + m_2 + m_3} = \frac{\sum_i m_i y_i}{\sum_i m_i}$$

$$z_{cm} = \frac{m_1z_1 + m_2z_2 + m_3z_3 + \ldots}{m_1 + m_2 + m_3} = \frac{\sum_i m_i z_i}{\sum_i m_i}$$

Also, x_{cm},y_{cm} and z_{cm} are components of the position vector of center of mass, so this is equivalent to the vector equation

$$\vec{r}_{cm} = \frac{m_1\vec{r}_1 + m_2\vec{r}_2 + m_3\vec{r}_3 + \ldots}{m_1 + m_2 + m_3} = \frac{\sum_i m_i \vec{r}_i}{\sum_i m_i}$$

Now let's consider gravitational torque on a body of arbitrary shape

We assume that the acceleration due to gravity has the same magnitude and direction at every point in the body. Every particle in the body experiences a gravitational force, and the total weight of the body is a vector sum of a large number of parallel forces. A typical particle has mass and weight . If is the position vector of this particle with respect to an arbitrary origin O, then the torque vector of the weight with respect to O is,

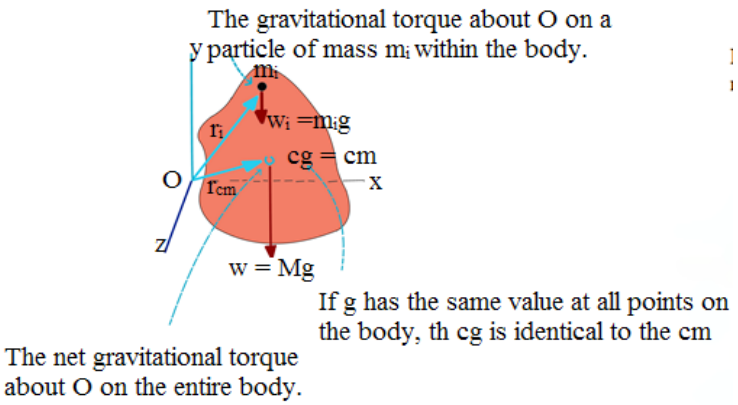


Figure 20.12: Center of gravity of a body

$$\vec{\tau}_i = \vec{r}_i \times \vec{w}_i = \vec{r}_i \times m_i \vec{g}$$

The total torque due to the gravitational forces on all the particles is

$$\vec{\tau} = \sum_i \vec{\tau}_i = \vec{r}_1 \times m_1 \vec{g} + \vec{r}_2 \times m_2 \vec{g} + \ldots$$
$$= \left(m_1 \vec{r}_1 + m_2 \vec{r}_2 + \ldots\right) \times \vec{g}$$
$$= \left(\sum_i m_i \vec{r}_1\right) \times \vec{g}$$

When we multiply and divide this by the total mass of the body,

$$M = m_1 + m_2 + \ldots = \sum_i m_i$$

We get

$$\vec{\tau} = \frac{m_1 \vec{r}_1 + m_2 \vec{r}_2 + \ldots}{m_1 + m_2 + \ldots} \times M\vec{g} = \frac{\sum_i m_i \vec{r}_i}{\sum_i m_i} \times M\vec{g}$$

The fraction in this equation is just the position vector of the center of mass, with components x_{cm}, y_{cm} and z_{cm}, and $M\vec{g}$ is equal to the total weight $\vec{w}$ of the body. Thus

$$\vec{\tau} = \vec{r}_{cm} \times M\vec{g} = \vec{r}_{cm} \times \vec{w}$$

The total gravitational torque is the same as though the total weight $\vec{w}$ were acting on the position $\vec{r}_{cm}$ of the center of mass, which we also call the center of gravity. If $\vec{g}$ has the same value at all points on a body, its center of gravity is identical to its center of mass. Note, however, that the center of mass is defined independently of any gravitational effect. While the value of $\vec{g}$ does vary somewhat with elevation, the variation is extremely slight. Hence we will assume that the center of gravity and the center of mass are identical unless explicitly stated otherwise.

20.5.3. Equilibrium of a system of objects

Balancing turning forces

An object in static equilibrium is one that has no acceleration in any direction. While there might be motion, such motion is constant.

Two children on a seesaw: the system is in static equilibrium, showing no acceleration in any direction.

When an object is balanced on a pivot the turning effect of the forces on one side of the pivot must balance the turning effect of the forces on the other side of the pivot - if they didn't it would not balance. In the picture (Figure 20.13(a)) two girls are sitting on a see saw. They have moved until it is balanced. They are the same weight and so to balance the see saw they must sit the same distance from the pivot.

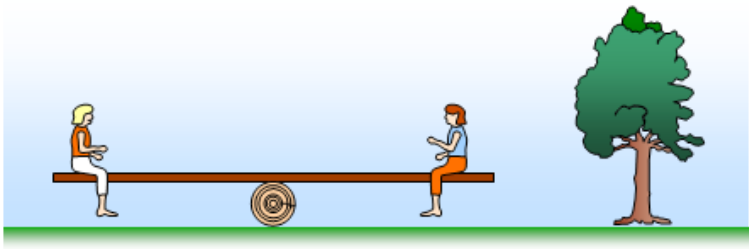
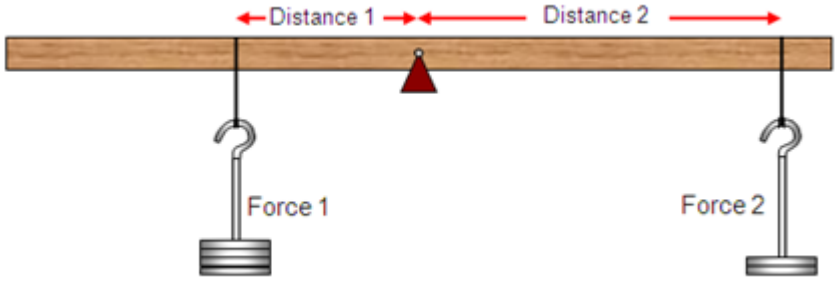


Figure 20.13 (a): A see saw in balance with equal weights

Figure 20.13 (b): A see saw in balance with different weights

In the picture (Figure 20.13(b)) one of the girls gets off and a man sits on instead. They move until the see saw is balanced. The girl is much lighter than the man and so she has to sit further away from the pivot then he does so that she can balance his extra weight.

You should remember that the turning effect of a force is called the moment of the force and is found by multiplying the force by its distance from the pivot. When the see saw is balanced we say that the anticlockwise moments (those trying to turn the object anticlockwise) equal the clockwise moments (those trying to turn the object clockwise). In our example the man's weight tries to turn the see saw clockwise and the girl's weight tries to turn it anticlockwise. You can investigate this in the lab by using sets of weights hanging on a wooden ruler (Figure 20.14)



The rule for something to be balanced is called the principle of moments and is written as follows: *when an object is balanced (in equilibrium) the sum of the clockwise moments is equal to the sum of the anticlockwise moments.*

Force 1 x perpendicular distance 1 = Force 2 x perpendicular distance 2 from pivot

$$F_1 x d_1 = F_2 x d_2$$

Example 20.2: Balancing a seesaw.

A board of mass M = 2.0 kg serves as a seesaw for two children, as shown in figure below. Child A has a mass of 30 kg and sits 2.5m from the pivot point, P (his center of gravity is 2.5m from the pivot).

At what distance x from the pivot must child B, of mass 25 kg, place herself to balance the seesaw? Assume the board is uniform and centered over the pivot.

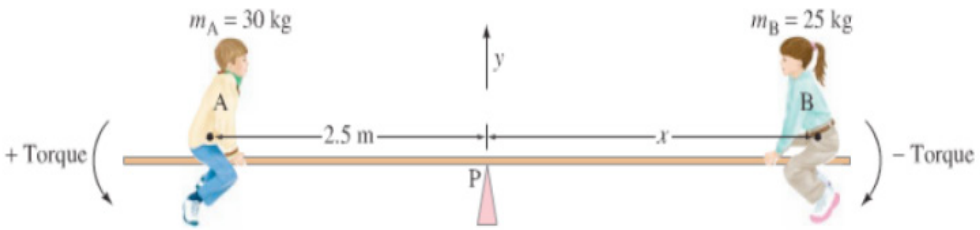


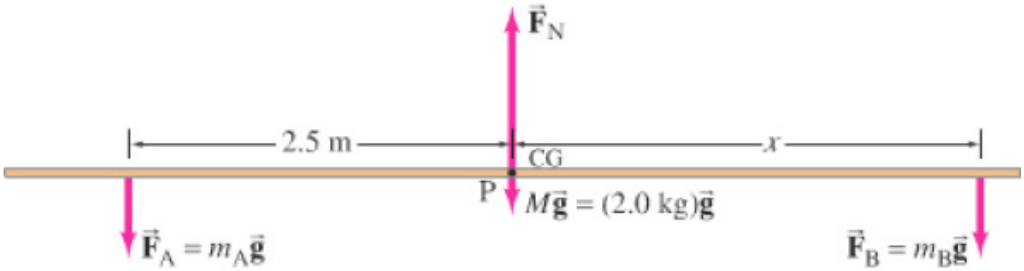
Figure 20.14: Sets of weights hanging on a wooden ruler

Approach

We follow the steps of the problem solving box explicitly.

Solution

1. Free-body diagram. We choose the board as our object, and assume it is horizontal. Its free-body diagram is



The forces acting on the board are the forces exerted downward on it by each child, $\vec{F}_A$ and $\vec{F}_B$, the upward force exerted by the pivot $\vec{F}_N$ and the force of gravity on the board (= Mg) which acts at the center of the uniform board.

2. Coordinate system. We choose y to be vertical, with positive upward, and x horizontal to the right, with origin at the pivot.
3. Force equation. All the forces are in the y (vertical) direction, so

   $\Sigma F_y = 0$

   $F_N - m_A g - m_B g - Mg = 0$

   Where $F_A = m_A g$ and $F_B = m_B g$ because each child is in equilibrium when the seesaw is balanced.
4. Torque equation. Let us calculate the torque about an axis through the board at the pivot point, P. then the lever arm for F_N and for weight of the board are zero , and they will contribute zero torque (about point P) in our torque equation. Thus the torque equation will involve only the forces $\vec{F}_A$ and $\vec{F}_B$, which are equal to the weights of children. The torque exerted by each child will be mg times the appropriate lever arm, which here is the distance of each child from the pivot point. Hence the torque equation is

   $\Sigma\tau = 0$

   $m_A g(2.5\ m) - m_B gx + Mg(0\ m) + F_N(0\ m) = 0$

   Or

   $m_A g(2.5\ m) - m_B gx = 0,$

   where two terms were dropped because their lever arms were zero.
5. Solve. We solve the torque equation for x and find

$$x = \frac{m_A}{m_B}(2.5\,m) = \frac{30\,kg}{25\,kg}(2.5\,m) = 3.0\,m$$

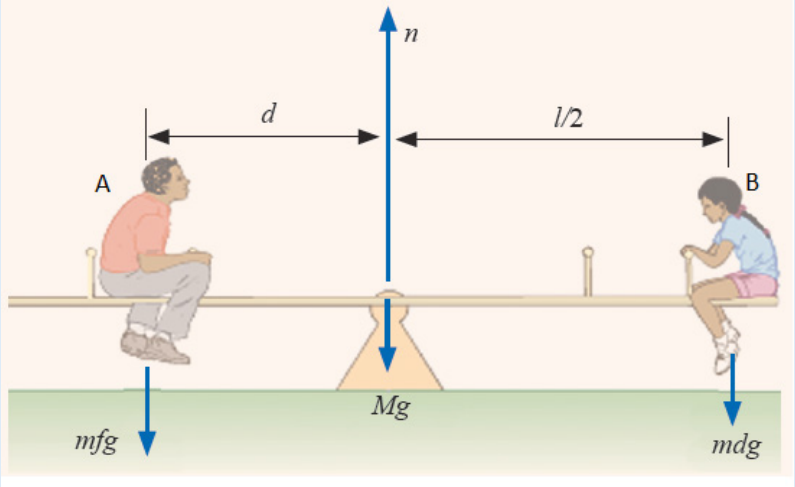
To balance the seesaw, child B must sit so that her CM is 3.0 m from the pivot point. This makes sense: since she is lighter, she must sit farther from the pivot than the heavier child.

Application activity 20.5

A seesaw consisting of a uniform board of mass M = 10 kg and length l = 2 m supports a father and a daughter with mass mf and md, 50 kg and 20 kg respectively as shown in the figure. The support (called the fulcrum) is under the centre of gravity of the board. The father is a distance d from the center, and the daughter is at distance l/2 from the center.

a) Determine the magnitude of the upward force (reaction) n exerted by the support on the board.

b) Determine where the father should sit to balance the system.



Skills lab 20

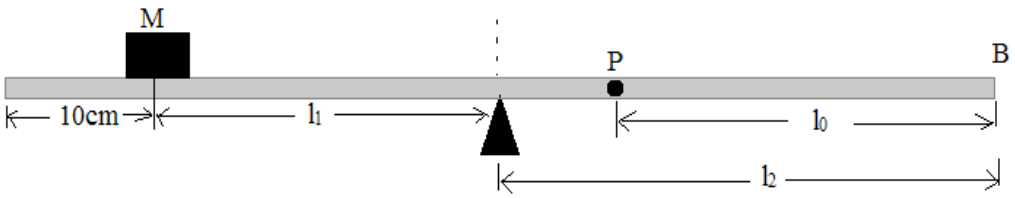
Title: determination of the mass of the meter rule

Apparatus: 2 meter rules, knife edge, 6 masses of 10 g,

Procedure

1. Balance the meter rule provided on a knife edge with the graduated side facing upwards.

2. Note the balance point P and record its distance from B.
3. Place a mass M of 10 g on the top of mater rule at the 10 cm mark and balance the arrangement as shown in figure below.
4. Read and record distance l_1 and l_2
5. Repeat the procedures (3) and (4) from the values of M = 20,30,40,50, and 60g.
6. Record your results in a suitable table including the values of $l_2 - l_0$ and $\frac{l_2 - l_0}{l_1}$



7. Plot a graph of M against $\frac{l_2 - l_0}{l_1}$.
8. Find the slope of your graph.

End unit assessment 20

1. ..... is an example of a scalar quantity

   a) Velocity. b) Force. c) Volume. d) Acceleration.
2. ..... is an example of a vector quantity

   a) Mass. b) Force. c) Volume. d) Density.
3. A scalar quantity:

   a) Always has mass.

   b) Is a quantity that is completely specified by its magnitude.

   c) Shows direction.

   d) Does not have units.

4. A vector quantity

   a) Can be a dimensionless quantity.

   b) Specifies only magnitude.

   c) Specifies only direction.

   d) Specifies both a magnitude and a direction.

5. A boy pushes against the wall with 50 kilograms of force. The wall does not move. The resultant force is:

   a) -50 kilograms.

   b) 100 kilograms.

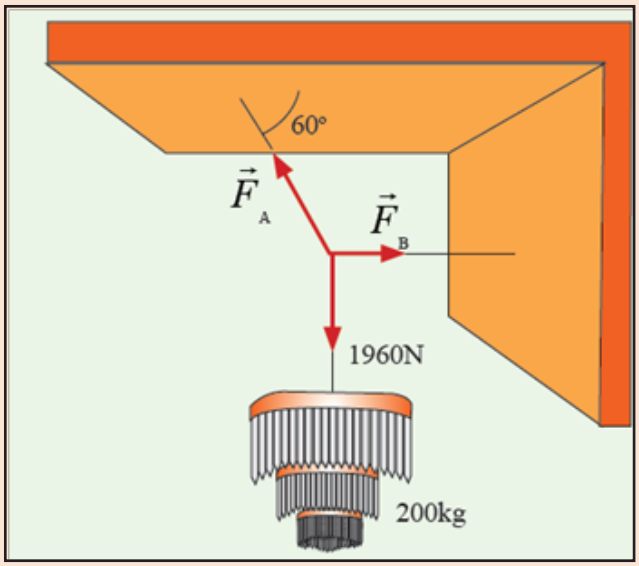
   c) 0 kilograms.

   d) -75 kilograms.

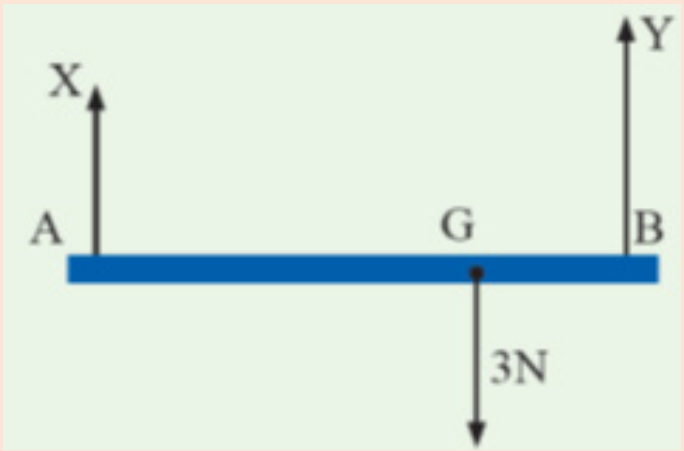
6. a) State the principle of moment

   b) What are the conditions for equilibrium?

7. Calculate the magnitudes FA and FB of the tensions in the two cords that are connected to the vertical cord supporting the 200kg chandelier in the following figure.



8. A horizontal rod AB is suspended at its ends by two strings. (See the figure below). The rod is 0.6m long and its weight of 3N acts at G where AG is 0.4m and BG is 0.2m. Find the tensions X and Y


X
Y
A
G
B
3N

UNIT 21:

PROJECTILE AND UNIFORM CIRCULAR MOTION

Key Unit competence: Analyse and solve problems related to projectile and circular motion.

Introductory Activity 21

1. a) Throw a ball, a stone or any object upwards.
   b) State what happens.
   c) Hold a ball in your hands and release it to fall.
   d) Is the motion of the ball same as in the first case?
   e) Note down your observation.
   f) Relate your observations for bodies moving linearly.
2. Observe the following image thereafter answer the question below

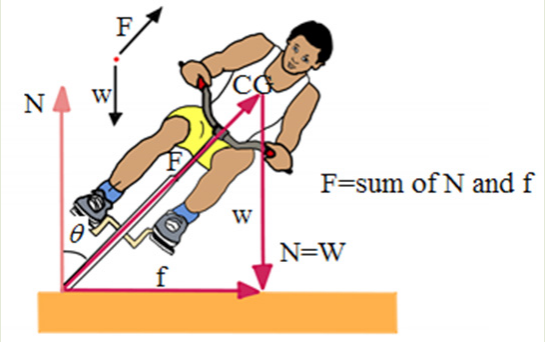


Figure 21.1 A cyclist negotiating a turn on level ground

a) What happen for this bicyclist to turn on level ground easily?

21.1. Projectile Motion

Activity 21.1

1. a) Kick the football individually.

   b) By observing, the flight of the ball state whether it will cover a longer horizontaldistance when it is projected at a large angle or a small angle.

   c) Explain your observation and note down any key points in your book.

2. Observe the image below and thereafter answer the next questions:

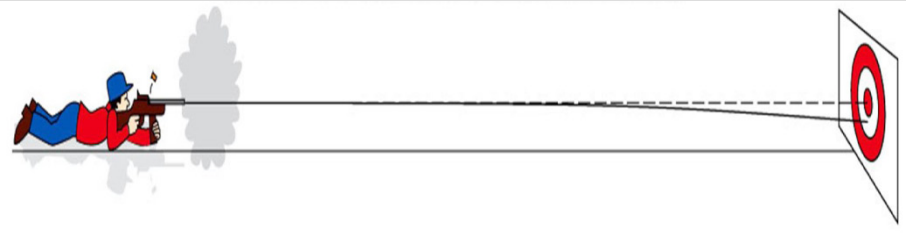


Figure 21.2: Firing a spot

   a) If the rifle is fired directly at the target in a horizontal direction, will the bullet hit the center of the target?

   b) Does the bullet fall during its flight?

Projectile motion is a predictable path traveled by an object that is influenced only by the initial launch speed, launch angle, and the acceleration due to gravity. You can try it out from where you're sitting. Pick up an object, and gently toss it up and away from you. It will rise as it flies away from you, reach a maximum height, and then start falling down to the floor. Toss a few more objects while you're at it. As long as you're not tossing pieces of paper or feathers, the projectile paths should be similar

Airplanes and birds are not considered projectiles because they are self-propelled. Neglecting air resistance, it is relatively easy to calculate the height reached, the horizontal distance travelled, and the object's velocity at any point in time, given only the initial velocity and the angle at which the projectile was propelled. The path taken is called a trajectory. We have three cases: oblique projection, vertical projection and horizontal projection.

The best angle to kick the football, If I am kicking the ball really slow, a 45° would be close to the best angle. At an angle of 45°, you get the best of both horizontal velocity and time for the ball to be in the air.

A projectile, in other words, travels the farthest when it is launched at an angle of 45 degrees.

The greatest distance is achieved using an angle close to 45° if the effects of air resistance are negligible.

Key terms in projectile

- **A projectile** is any object that is given an initial velocity and then follows a path determined entirely by gravitational acceleration. Regardless of whether you›re launching a balloon, a baseball, or an arrow, all projectiles follow a very predictable path, making them a great tool for studying kinematics.
- **Trajectory**: The path of aprojectile in motion, a parabola upward and across space.
- **Maximum height**:The highest vertical position along its trajectory. The object is flying upwards before reaching the highest point and it's falling after that point. It means that at the highest point of projectile motion, the vertical velocity is equal to 0 (Vy = 0).
- **The horizontal range of a projectile**:The distance along the horizontal plane it would travel, before reaching the same vertical position as it started from.
- **Time of flight**: is the amount of time spends in the air. If the ground from which the projectile is launched is level, the time of flight only depends on the initial velocity V0, the launch angle , and the acceleration due to gravity.
- **Angle of projection** is the angle between the initial velocities of a body from a horizontal plane through which the body is thrown.

Application activity 21.1

1. Make a research on other terms used in projectile motion
2. Draw an image of an object which can apply different terms used in projectile motion thereafter interpret it.

21.2. Applications of projectile motion

Activity 21.2

Task 1

1. Analyze the picture below and answer those questions:

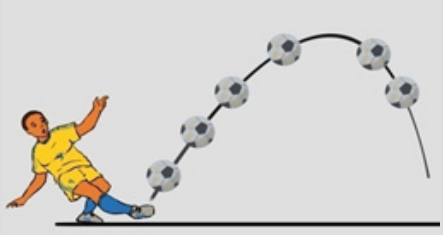


Figure 21.3: Kicked ball

a) What is the cause turning back of this kicked ball?

b) Through your analysis, if the ball is kicked at 45^0 from horizontal for first session and at 60^0 from horizontal for second session. What will happen? Explain your reason.

Task 2

You may have seen soldiers in the television news firing their rifles into the air to celebrate some victory. The bullets travel as high speed projectiles and, despite air resistance, return to Earth at high enough speeds to be dangerous. Reports indicate that from time to time people are injured by the returning bullets.

a) Explain your reasoning about the above statement

b) Plot a graph of the motion of rifles they flown

Task 3

At the instant horizontally pointed cannon is fired, a cannonball held at the cannon's side is released and drops to the ground. Which cannonball strikes the ground first, the one fired from the cannon or the one dropped?

The first important area which comes to mind is **ball games and sports**. Footballs are heavy enough to follow a nearly parabolic trajectory, without spin, with the effect of spin often being spectacular. Footballers have to develop a feel for such trajectories.

A cricket ball is small and dense enough to follow a nearly parabolic path, and it is up to the batsman to judge this, in playing his shot. The fielder must firstly get himself into the same plane as the flight, and then needs to judge

where to place himself. The symmetry of the flight means that when it comes down to a catchable height, it will be moving at the same speed that it left the bat.

With tennis and table tennis, the constant use of spin, even with lob shots, makes parabolic trajectories less important. This is also true in games such as golf, where spin predominates, and in rugby, where the shape of the ball affects its motion.

Scoring from distance in basketball is an example of the pure judgement of parabolic flight, and can only be mastered by constant practice.

There are the obvious throwing events, such as javelin, shot put and hammer, where there is obviously pure projectile motion, but this is also the case in jumping events, where the centre of mass of the body follows a purely parabolic trajectory after take off.

Fundamentally, we can say that whenever anything moves freely, without support, we have an example of projectile motion, and there are many that are not mentioned, but perhaps we ought to end with satellites and space stations. These orbit the earth in circles, but they are still in constant freefall, and they are still technically projectiles, and an example of projectile motion.

21.2.1. Horizontal projection

Along Y-axis the object moves as in a free fall which means that the motion is accelerated by g in the opposite direction of the axis. On the X-axis there is no acceleration and the object moves as in a URM.

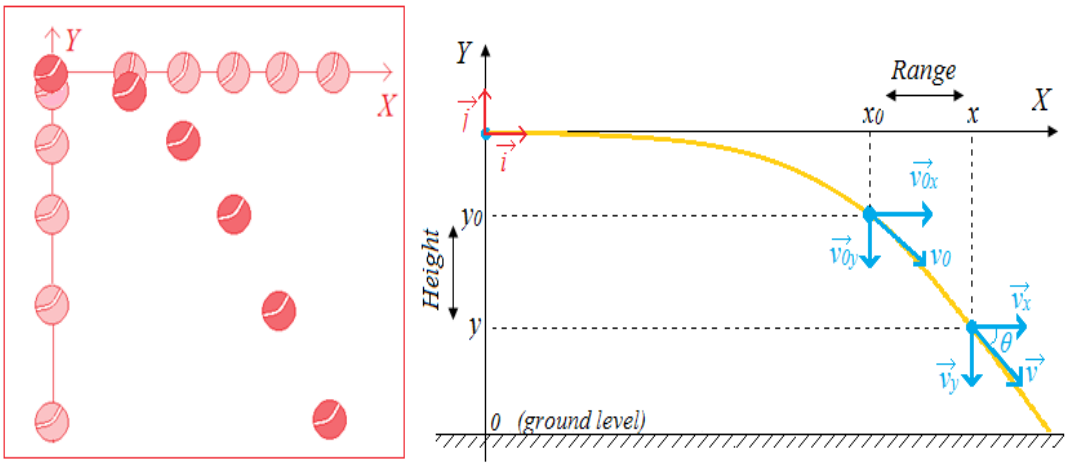


Figure 21.4: Objects thrown horizontally

a. General condition

Velocity

At t_0, the components of the initial velocity are

$$v_{0x} = v_0 \cos\theta$$
$$v_{0y} = v_0 \sin\theta$$

We can write the velocity at t_0

$$\vec{v}_0 = v_{0x}\vec{i} + v_{0y}\vec{j} = v_0 \cos\theta\,\vec{i} + v_0 \sin\theta\,\vec{j}$$

At every point of the path, the components of the velocity after the time $\Delta t = t - t_0$ are respectively:

$$v_x = v_0 \cos\theta$$

$$v_y = g(t - t_0) + v_0 \sin\theta$$

The velocity at the moment t

$$\vec{v} = v_x\vec{i} + v_y\vec{j} = (v_0 \cos\theta)\vec{i} + (g(t - t_0) + v_0 \sin\theta)\vec{j}$$

The magnitude of the velocity at the moment t is

$$v = \sqrt{v_x^2 + v_y^2}$$

$$v = \sqrt{v_0^2 + 2gtv_0 \sin\theta + g^2(t - t_0)^2}$$

At any moment, the angle made by the velocity with the horizontal is deduced from

$$\tan\theta = \frac{v_y}{v_x}$$

Position

At every point of the path, the position and the speed of the object relatively to the y direction after a time $\Delta t = t - t_0$ are respectively:

$$y = \frac{1}{2}g(t - t_0)^2 + (v_0 \sin\theta)(t - t_0) + y_0)$$

The relation between y, v and g is:

$$\left(v^2 - v_0^2\right)\sin^2\theta = 2g\left(y - y_0\right)$$

At every point of the path, the velocity v_{0x} doesn't change in the x direction.

The range after a time $\Delta t = t - t_0$ is:

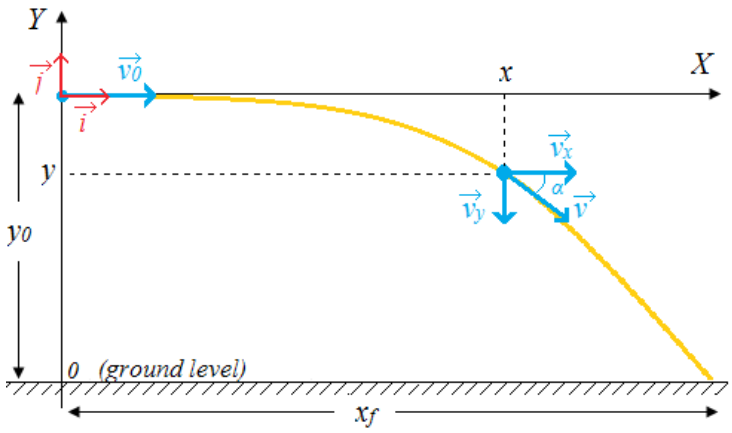
$$x = v_0 \cos\theta\left(t - t_0\right) + x_0$$

The vector position is given by

$$\vec{r} = x\vec{i} + y\vec{j} = \left(v_0 \cos\theta\left(t - t_0\right) + x_0\right)\vec{i} + \left(\frac{1}{2}g\left(t - t_0\right)^2 + v_0 \sin\theta\left(t - t_0\right) + y_0\right)\vec{j}$$

Initial condition

Consider, for example, a projectile thrown horizontally from the top of a cliff of height y_0, with the initial velocity v_0.



Velocity

In the beginning the object is thrown horizontally along X-axis $\left(\theta = 0\right)$. The components of the initial velocity at $t_0 = 0$ are

$$v_{0x} = v_0 \cos\theta = v_0$$
$$v_{0y} = v_0 \sin\theta = 0$$

The components of the velocity become at the moment t

$$v_x = v_0$$
$$v_y = gt$$

This leads to the velocity

$$\vec{v} = v_x\vec{i} + v_y\vec{j} = v_0\vec{i} + g\vec{t}\vec{j}$$

We deduce the speed of the projectile at the moment t:

$$v = \sqrt{v_x^2 + v_y^2} = \sqrt{v_0^2 + g^2t^2}$$

Position

The projectile coordinates at the moment t becomes

$$x = v_{0x}t$$
$$\Rightarrow x = v_0 t \qquad (1)$$
$$y = \frac{1}{2}gt^2 + v_{0y}t + y_0$$
$$\Rightarrow y = \frac{1}{2}gt^2 + y_0 \qquad (2)$$

The position vector of the projectile at the moment t is

$$\vec{r} = x\vec{i} + y\vec{j} = v_0 t\vec{i} + \left(\frac{1}{2}gt^2 + y_0\right)\vec{j}$$

The projectile motion is a parabola. Solving (1) for $(t - t_0)$

$$t = \frac{x}{v_0}$$

Replacing in (2) the equation of the parabola becomes

$$y(x) = -\frac{g}{2v_0^2}x^2 + y_0$$

Duration and final range of the motion

When the projectile reaches the ground, $y = 0$ and equation (2) yields

$$\frac{1}{2}gt^2 + y_0 = 0$$

We deduce the final time it takes the projectile to reach the ground:

$$t_f = t = \sqrt{-\frac{2y_0}{g}}$$

Replacing t_f in (1), the final range from the cliff is given by

$$x_f = v_0\sqrt{-\frac{2y_0}{g}}$$

21.2. Motion of a projectile thrown obliquely

Now, we are going to examine the more general motion of object moving through the air in two dimensions near the Earth's surface, such as a golf ball, athletes doing the long jump or high jump. Specifically, in the absence of air resistance, there is no acceleration in the horizontal or x direction and the object moves as in the URM. In the y direction, the object moves as in the UDRM until when the object reaches the maximum height y_m. After, the motion proceeds as when the projectile is thrown horizontally.

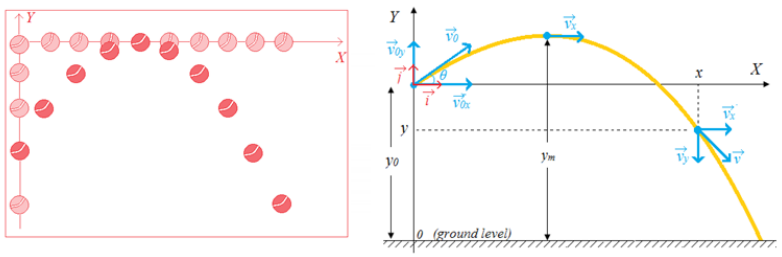


Figure 21.5: Object thrown oblically

Consider now, a projectile fired obliquely with initial velocity $\vec{v}_0$ at an angle θ above the horizontal axis.

a. General condition

When an object is fired obliquely, in the upward motion, it goes decelerating. After reaching the maximum height, it goes downward accelerating. We still can refer to the general conditions of the motion of a projectile thrown horizontally

b. Initial condition

- **Position**

The projectile coordinates at the moment t are

$$x = (v_0 \cos\theta)t$$

$$y = \frac{1}{2}gt^2 + (v_0 \sin\theta)t + y_0$$

The position vector of the projectile at the moment t is

$$\vec{r} = x\vec{i} + y\vec{j} = (v_0 \cos\theta t)\vec{i} + \left(\frac{1}{2}gt^2 + (v_0 \sin\theta)t + y_0\right)\vec{j}$$

Maximum height

When the object reaches its maximum height, $v_y = 0$.

$$v_y = gt + v_0 \sin\theta$$

$$v_y = 0 \Rightarrow gt + v_0 \sin\theta = 0$$

We deduce the time it takes for the projectile to reach the maximum height:

$$t_m = -\frac{v_0 \sin\theta}{g}$$

The maximum height y_m can be found affirming that $y_m = y(t_m)$. Replacing t_m in equation for y we get

$$y_m = -\frac{v_0^2 \sin^2\theta}{2g}$$

Note that the maximum height for a given initial velocity v_0 will be reached if the object is launched at $\theta = 90^0$, which the motion of a projectile is thrown vertically.

Motion duration

The motion of the object from the maximum height is similar to the motion of a projectile thrown horizontally from the equivalent height .The time t_g it takes the object to reach the ground from the maximum height is also deduced similarly.

$$t_g = \sqrt{-\frac{2y_m}{g}}$$

Then, the total duration of the motion will be the sum of the time to reach the maximum height t_m and the time to reach the ground from the maximum height t.

$$t_f = t_m + t = -\frac{v_0 \sin\theta}{g} + \sqrt{-\frac{2y_m}{g}}$$

c. Oblique motion launched from the ground level

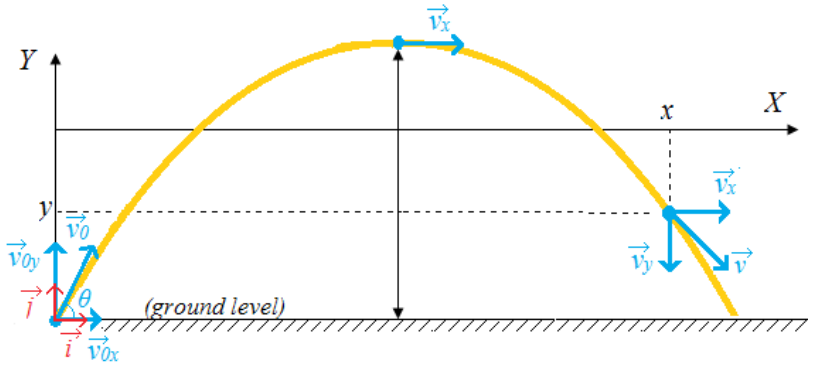


Figure 21.6: object thrown oblically from the ground level

Motion duration

If the object is launched from the ground $(y_0 = 0)$, it takes the same time to reach the maximum height as to fall back to the ground.

$$t_f = 2t_m = -\frac{2v_0 \sin\theta}{g}$$

Range

We can find how far away the projectile hits the ground as we know that $x_f = x(t_f)$:

$$x_f = xt_f = v_0 \cos\theta t_f = v_0 \cos\theta\left(\frac{2v_0 \sin\theta}{g}\right) = v_0^2\left(\frac{2\sin\theta\cos\theta}{g}\right)$$

Since $2\sin\theta\cos\theta = \sin 2\theta$,

$$x_m = \frac{v_0^2 \sin 2\theta}{g}$$

The maximum range, for a given initial velocity v_0, is obtained when the sine takes on its maximum value of 1, which occurs for $\theta = 45^0$. In this case

$$x_{\max} = \frac{v_0^2}{g}$$

Notice: The magnitude of g is $g = -9.8\, m/s^2$

Motion of projectiles thrown obliquely

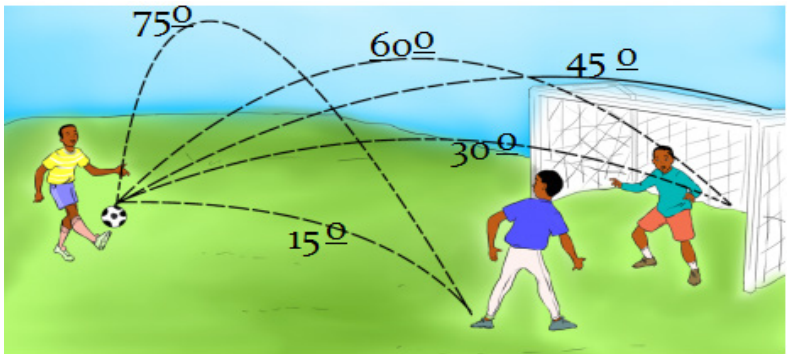


Figure 21.7: Motion of projectiles

The paths of projectiles launched at the same speed but at different angles. The paths neglect air resistance. On the figure above the paths of several projectiles all having the same initial speed but different projection angles. The figure neglects the effects of air resistance, so the paths are all parabolas. Notice that these projectiles reach different heights above the ground. They also travel different horizontal distances, that is, they have different horizontal ranges. The remarkable thing to note form the figure above is that the same range is obtained for two different projection angles, angles that add up to 90^0.

For example, an object thrown into the air at an angle of 60^0 will have the same range as if it were thrown at 30^0 with the same speed. Of course, for the smaller angle the object remains in the air for a shorter time.

We have emphasized the special case of projectile motion for negligible air resistance.

Example

A football is kicked at an angle $\Theta_0 = 37.0^0$ with a velocity 20.0 m/s. calculate

a). The maximum height,

b). The time of travel before the football hits the ground,

c). How far away it hits the ground.

d). The velocity vector at the maximum height,

e). The acceleration vector at the maximum height. Assume the ball leaves the foot at ground level, and ignore air resistance and rotation of the ball.

Solution

We resolve the initial velocity into its components

$v_{x0}=v_0 \cos 37.0^0 = (20.0\ m/s)(0.799) = 16.0\ m/s$

$v_{y0}=v_0 \sin 37.0^0 = (20.0\ m/s)(0.602) = 12.0\ m/s$

a). We consider a time interval that begins just after the football loses contact with the foot until it reaches its maximum height. During this time interval, the acceleration is g downward. At the maximum height, the velocity is horizontal so, $v_y=0$ and this occurs at a time given by $v_y=v_{y0}-gt$ with $v_y=0$. Thus

$$t=\frac{v_{y0}}{g}=\frac{12.0\,m/s}{9.80\,m/s^2}=1.22\,s$$

With $y_0=0$, we have

$$y=v_{y0}t-\frac{1}{2}gt^2=(12.0\,m/s)(1.22\,s)-\frac{1}{2}(9.80\,m/s^2)(1.22\,s)^2=7.35\,\text{m}$$

The maximum height is 7.35 m

b) To find the time interval it takes for the ball to return to the ground, we consider a different time interval, starting at the moment the ball leaves the foot ($t=0$, $y_0 = 0$) and ending just before the ball touches the ground ($y = 0$).

$$y = y_0 + v_{y0}t - \frac{1}{2}gt^2$$

$$0 = 0 + (12.0\,m/s)\,t - \frac{1}{2}(9.80\,m/s^2)\,t^2$$

This equation can be easily factored:

$$\left[\frac{1}{2}(9.80\,m/s^2)\,t - (12.0\,m/s)\right]t = 0$$

There are two solutions, t (which corresponds to the initial point, y_0), and

$$t = \frac{2(12.0\,m/s)}{9.80\,m/s^2} = 2.45\,s$$

Which is the total travel time of the football.

c) The total distance traveled in the x-direction is found by:

$$x_0 = 0, a_x = 0, v_{x0} = 16.0\,m/s$$
$$x = v_{x0}t = (16.0\,m/s)(2.45s) = 39.2\,m$$

d). At the highest point, there is no vertical component to the velocity. There is only the horizontal component (which remains constant through the flight), so

$$v = v_{x0} = v_0 \cos 37.0^0 = (20.0\,m/s)(0.799) = 16.0\,m/s$$

The acceleration vector is the same at the highest point as it is throughout the flight, which is 9.80 m/s^2 downward.

Application activity 21.2

1. There is more than one way to prove that the horizontal and vertical components of a projectile's motion are independent of each other. Describe two or three ways that you could use to analyze the motion of the two balls.
2. A projectile is launched at an angle into the air. Neglecting air resistance, what is its vertical acceleration? Its horizontal acceleration? At what point in its path does a projectile have minimum speed?
3. A cannon ball is fired with an initial velocity of 100. m/s at an angle of 45° above the horizontal. What maximum height will it reach and how far will it fly horizontally?

21.3. Circular motion: Meaning and related key terms

Activity 21.3

Study carefully the motion of the ball shown below.

1. From your observation of below figure, how can a body move in that circular path? Explain your reason

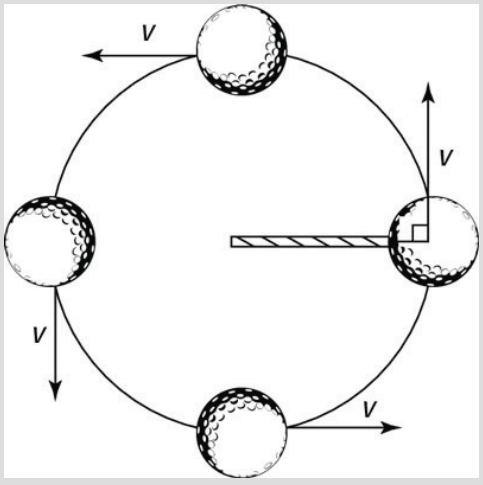


Figure 21.8: Ball moving in a circular path

Circular motion: Movement of an object along the circumference of a circle, or rotation along a *circular* path. It is a *motion* with constant velocity.

Angular displacement: The *angle* in radian through which a point revolves around a centre or line has been rotated in a specified sense about a specified axis.

Linear velocity: The measure of how fast the dot is moving along the circumference of the circle

Angular velocity: The rate at which the central angle swept out by the object changes as the object moves around the circle, and it is thus measured in radian per unit time.

Period: The period is the time taken for a full revolution of the motion

Frequency: The number of rotation per unit time.

Angular acceleration (rotational acceleration): Quantitative expression of the change in angular velocity that a spinning object undergoes per unit time. It is a vector quantity, consisting of a magnitude component and either of two defined directions or senses.

Linear acceleration: The rate of change of linear velocity

Centripetal acceleration: The acceleration directed toward the center of a circle, which causes uniform circular motion.

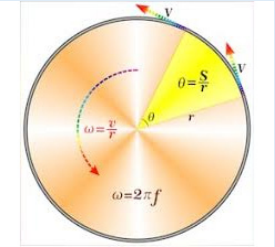
Centripetal force: The force acting towards the centre of a circle. It is a force that is needed (not caused by) a circular motion, for example when you swing a ball on a string round in a circle, the tension of the string is the centripetal force. If the string is cut then the ball will travel in a straight line at a tangent to the circle at the point where the string was cut (Newton's first law)

Centrifugal force (nonexistent force): The force acting away from the centre of a circle. This is what makes a slingshot go outwards as you spin it. The centrifugal force is the reaction to the centripetal force (Newton's third law). It has the same magnitude but opposite direction to the centripetal force ("equal but opposite").

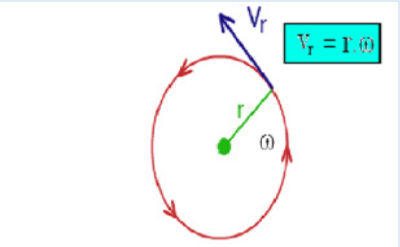
Critical velocity: Minimum velocity for an object to travel in vertical circle and maintain its circular path against the force of gravity.

Application activity 21.3

A



B



C

After observing the above pictures, what A, B, C and D represent in circular motion?

21.4. Relationship between angular and linear parameters

Activity 21.4

Write your observation on the image below

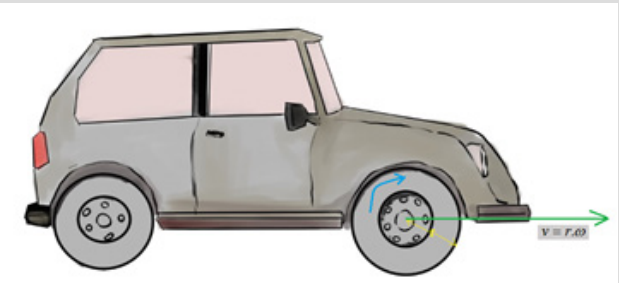


Figure 21.9 A car moving at a velocity, v, to the right has a tire rotating with angular velocity ω

1. What do variables v, ω and r mean in the expression illustrated on the figure 21.12
2. What does ω show in the movement of the wheel in the illustration above.

21.4.1. Angular Velocity

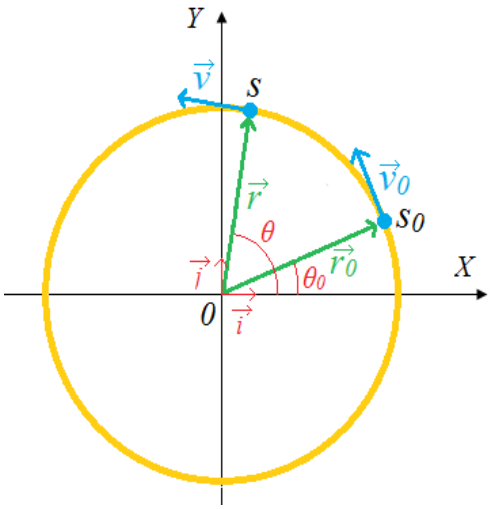


Figure 21.10 Uniform circular motion

To indicate the angular position of a particle, or how far it has rotated we specify the angle θ of a line joining the centre of the particle and its position with respect to some reference line, such as the x axis. Consider an object moving in a circle of radius r with a uniform speed v round a fixed point 0 as centre.

When an object rotates the angular displacement $\Delta\theta = \theta - \theta_0$, the average angular velocity is defined as

$$\omega = \frac{\Delta\theta}{\Delta t} = \frac{\theta - \theta_0}{t - t_0} \qquad (1)$$

The linear displacement or the Arc of the object along the circle is

$$\Delta s = s - s_0 = r\Delta\theta$$

The linear average velocity

$$v = \frac{\Delta s}{\Delta t} = r\frac{\Delta\theta}{\Delta t} = r\varpi \qquad (2)$$

We define the instantaneous angular velocity as the very small angle $\Delta\theta$, through which the object turns in the very short time interval Δt:

$$\omega = \lim_{\Delta t \to 0} \frac{\Delta \theta}{\Delta t} = \frac{d\theta}{dt} \quad (3)$$

The instantaneous linear velocity

$$v = \lim_{\Delta t \to 0} \frac{\Delta \theta . r}{\Delta t} = r \frac{d\theta}{dt} \quad (4)$$

The angular velocity is generally specified in radian per second (rad/s) whereas the instantaneous linear velocity is expressed in (m/s).

21.4.2. Periodic time and frequency

The period T is the time needed for the object to make one complete revolution. During this time, the object travels a distance equal to the circumference of the circle. The frequency f is referred to as the number of revolutions made by an object in one second. The unit of frequency is Hertz.

$$f = \frac{1}{T} \quad (5)$$

The object's angular speed is then represented by

$$\omega = \frac{2\pi}{T} = 2\pi f \quad (6)$$

21.4.3. The average angular acceleration

The average angular acceleration is defined as the change in angular velocity divided by the time required to make this change:

$$\beta = \frac{\Delta\omega}{\Delta t} = \frac{\omega - \omega_0}{t - t_0} \quad (7)$$

The average linear acceleration

$$a = \frac{\Delta s}{\Delta t} = \frac{r\Delta\omega}{\Delta t} = r\beta \quad (8)$$

The instantaneous angular acceleration is

$$\beta = \lim_{\Delta t \to 0} \frac{\Delta \omega}{\Delta t} = \frac{d\omega}{dt} \qquad (9)$$

The instantaneous linear acceleration

$$a = \lim_{\Delta t \to 0} \frac{\Delta v}{\Delta t} = r \frac{\Delta \omega}{\Delta t} = r \frac{d\omega}{dt} \qquad (10)$$

The angular acceleration is generally specified in radian per second $\left(rad/s^2\right)$ whereas the instantaneous linear acceleration is expressed in $\left(m/s^2\right)$.

Application activity 21.4

1. A ball at the end of a string is revolving uniformly in a horizontal circle of radius 2 meters at constant angular speed 10 rad/s. Determine the magnitude of the linear velocity of a point located:
   a) 0.5 meters from the center
   b) 1 meter from the center
   c) 2 meters from the center
2. The blades in a blender rotate at a rate of 5000 rpm. Determine the magnitude of the linear velocity :
   a) a point located 5 cm from the center
   b) a point located 10 cm from the center
3. A point on the edge of a wheel 30 cm in radius, around a circle at constant speed 10 meters/second.
   What is the magnitude of the angular velocity?
4. The angular speed of wheel 20 cm in radian is 120 rpm. What is the distance if the car travels in 10 seconds.

21.5. Acceleration in circular motion: Constant and tangential acceleration

Activity 21.5

A car executing a turn, after a certain speed limit, a car will start drifting.

Why at that limitation of the speed the car start travelling? Explain your reasoning

21.5.1. Tangential and Centripetal acceleration

In the circular motion, the tangential acceleration a_T always points in the direction tangent to the circle and changes the rate of velocity in terms of magnitude because their vectors are always in the same or opposite direction. The tangential acceleration can be considered as the linear acceleration.

The centripetal acceleration (normal acceleration or radial acceleration) a_N changes the velocity in terms of direction and its vector is perpendicularly directed inward the circle.

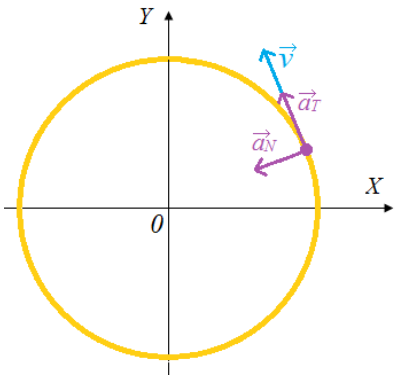


Fig 21.11: Tangential and Centripetal acceleration

Since the velocity is constant, the tangential component of acceleration doesn't exist in UCM:

$$a_T = \frac{dv}{dt} = 0$$

Consequently, the tangential component of the acceleration is also zero. Only, the normal component of the acceleration exists in UCM. Thus, the velocity v changes in direction but not in magnitude. The figure below shows the representation of the angular velocity $\vec{\omega}$ using a distant reference frame $\vec{i}, \vec{j}, \vec{k}$.

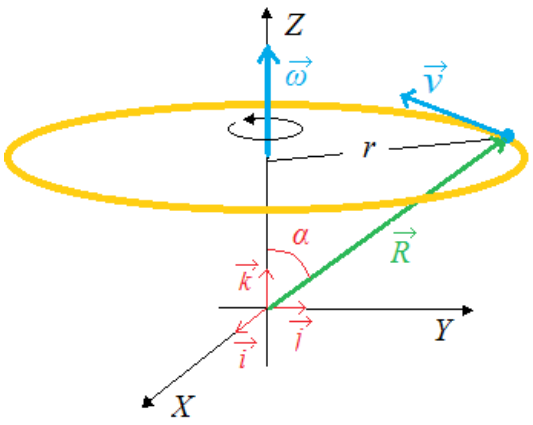


Fig 21.12: Tangential and Centripetal acceleration

The centripetal component of the acceleration is directed along the radius. From the above figure we can notice that $r = R\sin\alpha$.

$$v = r\omega = \omega R\sin\alpha$$

This relation indicates that the vector $\vec{v}$ can be expressed in the vector form by the equation:

$$\vec{v} = \vec{\omega}\times\vec{R} \qquad (1)$$

It follows that

$$a_N = \frac{dv}{dt} = \omega\sin\alpha\frac{dR}{dt}$$
$$\Rightarrow \vec{a}_N = \frac{d\vec{v}}{dt} = \vec{\omega}\times\frac{d\vec{r}}{dt} \qquad (2)$$
$$\vec{a}_N = \vec{\omega}\times\vec{v}$$

Introducing the equation (1) into (2)

$$\vec{a}_N = \vec{\omega}\times\vec{v} = \vec{\omega}\times\left(\vec{\omega}\times\vec{R}\right) \qquad (3)$$

The magnitude of the centripetal acceleration is given by

$$a_N = \omega v\sin\alpha$$

Since $\vec{\omega}\perp\vec{v}$, its magnitude is

$$a_N = \omega v$$

As $v = r\omega$, we also find that

$$a_N = \frac{v^2}{r} = r\omega^2 \qquad (4)$$

21.5.2. Acceleration in a non-uniform circular motion

Circular motion at a constant speed occurs when the acceleration of the object is directed toward the centre of the circle (Only the centripetal component is available). If the acceleration is not directed toward the centre but is at an angle, as shown in figure below, the acceleration has two components; the centripetal and the tangential component.

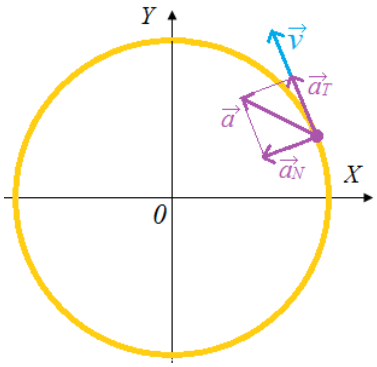


Fig 21.13: Acceleration in a non-uniform circular motion

The tangential component of the acceleration is the rate that changes the magnitude of the velocity

$$a_T = \frac{dv}{dt} = \frac{d(r\omega)}{dt} = r\frac{d\omega}{dt} = r\beta$$

The centripetal acceleration arises from the change in direction of the velocity and, as we have seen, is given by

$$a_N = r\omega^2 = v\omega = \frac{v^2}{r}$$

If the tangential acceleration is in the direction of motion i.e. parallel to $\vec{v}$, the speed is increasing and if it is antiparallel to $\vec{v}$ is the speed is decreasing.

In either case, $\vec{a}_T$ and $\vec{a}_N$ are always perpendicular to each other; and their directions change continually as the object moves along its circular path. The total vector acceleration is the sum of these two:

$$\vec{a} = \vec{a}_T + \vec{a}_N$$

the magnitude of $\vec{a}$ at any moment is

$$a = \sqrt{a_N^2 + a_T^2}$$

21.5.3. Distance time graph of circular motion

When an object executes a circular motion of constant radius R, its projection on an axis executes a motion of amplitude a that repeats itself back and forth, over the same path.

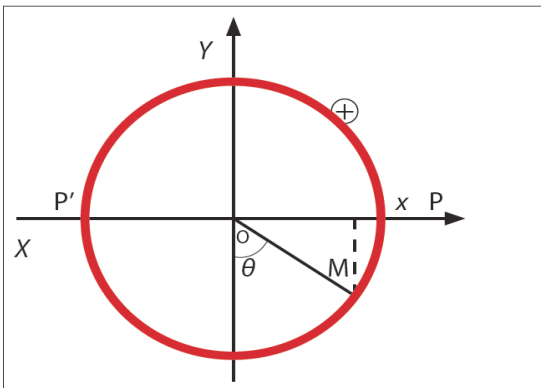


Figure 21.14. A circular motion can be projected on an axis

When M executes a uniform circular motion, its projection on X-axis executes a back and forth motion between positions P and P' about O.

Considering the displacement and the time, we find the following graph

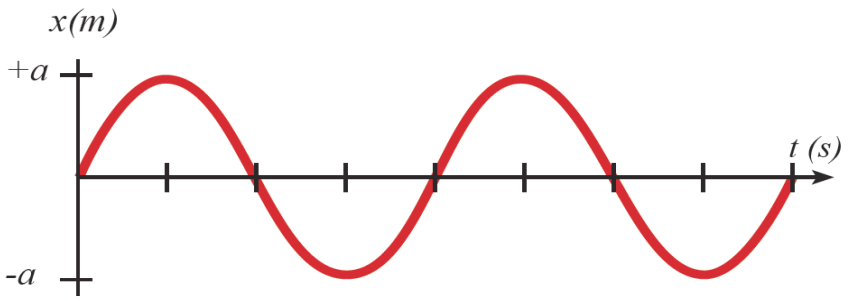


Fig 21.15. Displacement time graph of circular motion

Application activity 21.5

1. A ball is attached to a string that is 1.5m long. It is spun so that it completes two full rotations every second. What is the centripetal acceleration felt by the ball?
2. Imagine a car driving over a hill at a constant speed. Once the car has reached the apex of the hill, what is the direction of the acceleration?

21.6. Centripetal force

Activity 21.6

Observe the following figure and answer the question below:

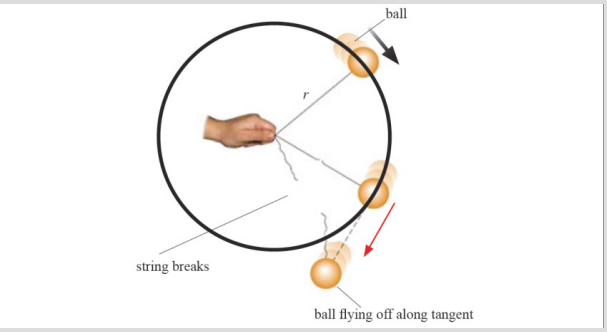


Fig 21.15: Ball swung around horizontal circle

What would happen when a ball is attached to a string and is swung round in horizontal circle? Explain your reasoning.

If you try to move in a circular path, you will finally notice that you keep moving in a circle even when you try to stop. There is a force that keeps you more in a circular path called centripetal force. Since a body moving in a circle (or a circular arc) is accelerating, it follows

from Newton's first law of motion that there must be force acting on it to cause the acceleration.

This force, like the acceleration, will also be directed toward the centre and is called the centripetal force. The value F of the centripetal force is given by

Newton's second law, that is: $F = ma_N = \frac{mv^2}{r}$

Wherem is the mass of the body and v is its speed in circular path of radius

R. If the angular velocity of the body is ω, we can also say, since $V = R\omega$,

$F = mR\omega^2$

Application activity 21.6

1. A 3.0 kg mass is tied to a rope and swung in a horizontal circle. If the velocity of the mass is 4.0 m/s and the radius of the circle is 0.75 m, what is the centripetal force and centripetal acceleration of the mass?
2. A 200-gram ball, attached to the end of a cord, is revolved in a horizontal circle with an angular speed of 5 rad s-1. If cord's length is 60 cm, what is the centripetal force?
3. A student of mass 50kg decides to go on the ride. The coefficient of static friction between the student and wall is 0.8. If the diameter of the ride is 10m, what is the maximum period of the ride's rotation that will keep the student pinned to the wall once the floor drops? $g=10m/s^2$

21.7. Applications of circular motion

Activity 21.7

Task 1

1. What is the difference between horizontal and vertical circular motion?
2. Does gravity affect horizontal circular motion? Explain your reason

Task 2

a) Tie a thread of about 50cm on retort stand.

b) On a thread, tie a pendulum bob.

c) Displace the bob through a certain angle. What do you observe?

d) Release the bob to move through a certain angle so that it moves in a horizontal circle.

e) Try to investigate forces acting in the bob.

Task 3

Observe the following image and answer the related questions

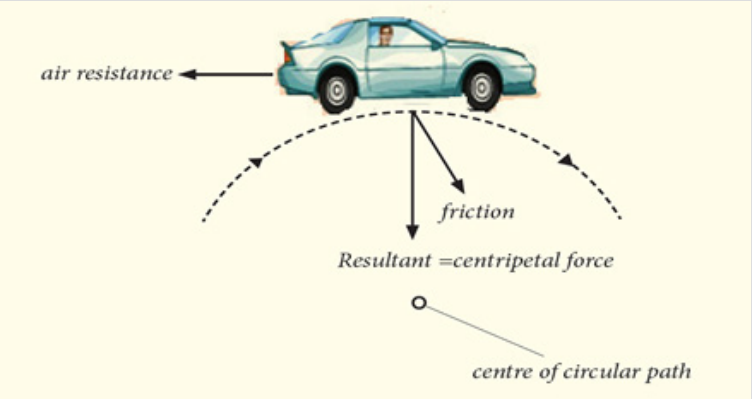


Fig 21.17: A car negotiating a corner

a) Explain the way the driver will use for not falling down at that corner of the above figure

b) Identify different factors needed to travel properly in the corner

21.7.1. Vertical and horizontal circle

1. Vertical circle

Taking the approach that the ball moves in a vertical circle and is not undergoing uniform circular motion, the radius is assumed constant, but the speed v changes because of gravity. The free-body diagram is shown in the figure below for the positions 1 and 2.

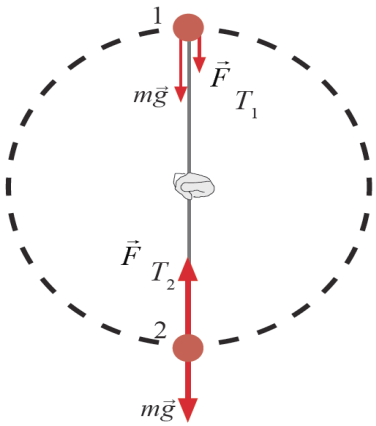


Figure 21.18: Free-body diagrams for position 1 and 2

a) At the top (point 1), two forces act on the ball: the force of gravity and the tension force the cord exerts at point 1. Both act downward and their vector sum acts to give the ball its centripetal acceleration. We apply Newton's second law, for t a vertical direction, choosing downward as positive since the acceleration is downward (toward the centre):

$$\sum \vec{F} = m\vec{a}$$

$$F_{T1} + mg = m\frac{v_1^2}{r}$$

From this equation, we can see that the tension force F_{T1} at point 1 will get larger if V_1 (ball's speed at top of circle) is made larger, as expected. But we are asked for the minimum speed to keep the ball moving in a circle.

The cord will remain taut as long as there is tension in it. But the tension disappears (because V_1 is too small), the cord can go limp and the ball will fall out of its circular path. Thus, the minimum speed will occur if $F_{T1} = 0$, for which we have (minimum speed at top)

We solve for V_1:

$$v_1 = \sqrt{gr}$$

This is the minimum speed at the top of the circle if the ball is to continue moving in a circular path.

b) When the ball is at the bottom of the circle (point 2 in the figure 21.9), the cord exerts its tension force upward, whereas the force of gravity, still acts downward. So we apply Newton's second law, this time choosing upward as positive since the acceleration is upward (toward the centre):

$$F_{T2} - mg = m\frac{v_2^2}{r}$$

We solve for

$$F_{T2} = m\frac{v_2^2}{r} + mg$$

2. Horizontal circle

The second case is the case of a force on revolving ball (horizontal) which is:

Estimate the force a person must exert on a string attached to a ball to make the ball revolve in a horizontal circle of radius r.

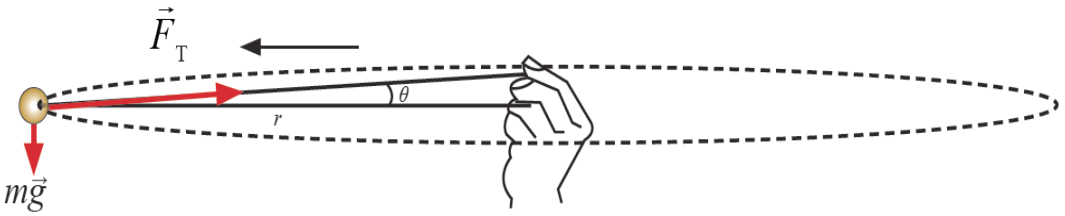


Figure 21.19: An object in a horizontal circular motion

The forces acting on the ball are the force of gravity, downward, and the tension that the string exerts toward the hand at the centre. The free-body diagram for the ball is as shown in the figure 21.22. The ball's weight complicates matter and makes it a little difficult to revolve a ball with the cord perfectly horizontal.

We assume the weight is small and put in the figure 21.22. Thus the tension will act nearly horizontally and, in any case, provides the force necessary to give the ball its centripetal acceleration. We apply Newton's second law to the radial direction $\sum F=ma_N$ where

$$a_N = \frac{V^2}{r} \rightarrow F_T = m\frac{V^2}{r}$$

21.7.2. Conical pendulum

A small object of mass m is suspended from a string of length L. The object revolves with constant speed v in a horizontal circle of radius r, as shown in Figure 21.9.2. Because the string sweeps out the surface of a cone, the system is known as a conical pendulum.

Let us find an expression for v.

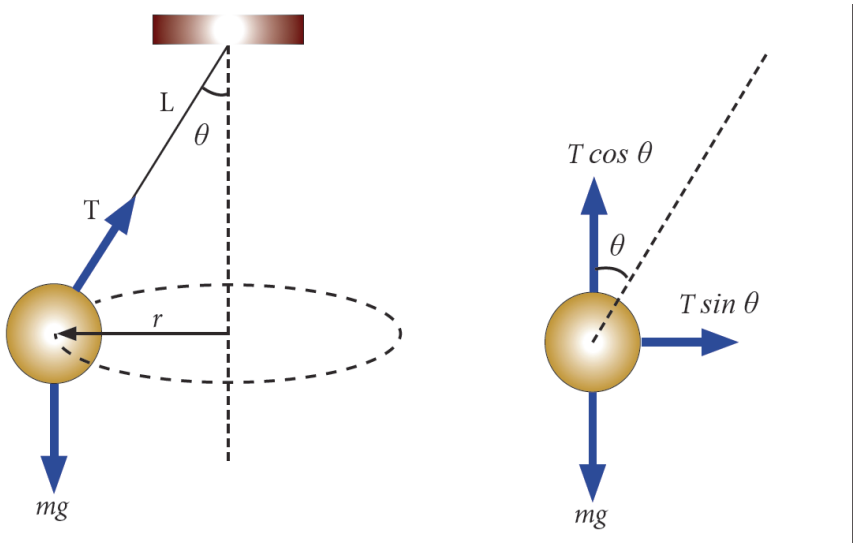


Figure 21.20: The conical pendulum and its free-body diagram

To analyse the problem, begin by letting θ represent the angle between the string and the vertical. In the free-body diagram shown, the force $\vec{T}$ exerted by the string is resolved into a vertical component $T\cos\theta$ and a horizontal component $T\sin\theta$ acting toward the centre of revolution. Because the object does not accelerate in the vertical direction, $\sum \vec{F}_y = m\vec{a}_y = 0$ and the upward vertical component of $\vec{T}$ must balance the downward gravitational force. Therefore;

$$T\cos\theta = mg \qquad (1)$$

Because the force providing the centripetal acceleration in this example is the component , we can use the formula of centripetal acceleration to obtain

$$\sum F = T\sin\theta = ma_N = m\frac{v^2}{r} \qquad (2)$$

21.7.2. Conical pendulum

Dividing (2) by (1) and using $\frac{\sin\theta}{\cos\theta} = \tan\theta$; we eliminate T and find that

$$\tan\theta = \frac{v^2}{rg} \rightarrow v = \sqrt{rg\tan\theta}$$

From the geometry in Figure 21.23, we see that $r = L\sin\theta$; therefore,

$$v = \sqrt{\mathrm{Lg}\sin\theta\tan\theta}$$

21.7.3. Road banking

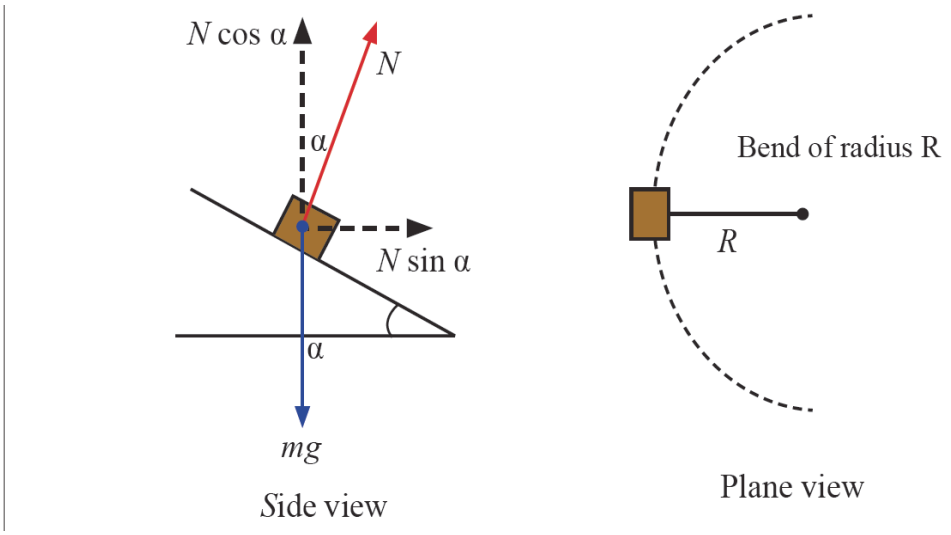


Figure 21.21: Rounding a bend

Treating a car as a particle and resolving $\vec{N}$ vertically and horizontally we have; since $N\sin\alpha$ is the centripetal force $N\sin\alpha = m\dfrac{v^2}{r}$

Where m and v are the mass and the speed respectively of the car and R is the radius of the bend.

Also, the car is assumed to remain in the same horizontal plane and so has no vertical acceleration, therefore $N\cos\alpha = \mathrm{mg}$

Hence by division:

$$\tan\alpha = \frac{v^2}{gR} \rightarrow v^2 = Rg\tan\alpha$$

$$v = \sqrt{Rg\tan\alpha}$$

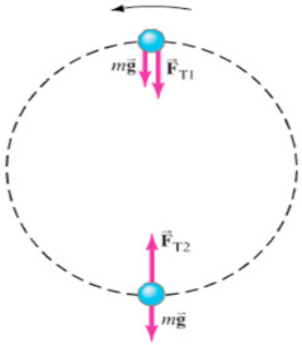
The equation shows that for a given radius of bend, the angle of banking is only correct for one speed.

Example

A ball on the end of a string is revolved at a uniform rate in a vertical circle of radius 72.0 cm, as shown in figure below. If its speed is 4.00 m/s and its mass is 0.300 kg; calculate the tension in the string when the ball is

a). At the top of its path,

b). At the bottom of its path



Solution

a). Take positive to be downward. Write Newton's 2nd law in the downward direction.

$$\sum F_R = F_{T1} + mg = m\frac{v_1^2}{r}$$

$$F_{T1} = m\left(v^2/r - g\right) = \left(0.300kg\right)\left[\frac{\left(4.00\,m/s\right)^2}{0.720m} - 9.8\,m/s^2\right] = 3.73N$$

This a downward force, as expected.

b). Take positive to be upward. Write Newton's 2nd law in the upward direction.This is an upward force, as expected.

$$\sum F_R = F_{T2} - mg = m\frac{v^2}{r}$$

$$F_{T2} = m\left(v^2/r + g\right) = \left(0.300kg\right)\left[\frac{\left(4.00\,m/s\right)^2}{0.720m} + 9.8\,m/s^2\right] = 9.61N$$

Application activity 21.7

1. Why do we give banking to curved roads?
2. A particle is revolving in a circle increasing its speed uniformly. Which of following is constant? why? and why not? Explain the all three of option.

   a) Centripetal force b) Tangential acceleration c) Angular acceleration
3. A steel ball of mass 0.5kg is suspended from a light inelastic string of length 1m. The ball swings in a horizontal circle of radius 0.5 m. Find
   i. The centripetal force and tension in the string.
   ii. The angular Speed of the ball.
4. A student twirls a 75 gram weight at the end of a string in a horizontal circle of radius 50 cm. If the speed of the weight is 7.0 m/s, what is the tension in the string?

Skills lab 21

Go on youtube.com, watch video about “Trucks fail to Negotiate Dangerous Bend in Road ”. It is a social issue which relationship with Circular motion.

1. What did you see on the video and what happen?
2. If you are engineer, how you will solve this problem?
3. Brainstorm with your classmates about the property needed to design a safety road by used scientific knowledge in Circular motion.

End unit assessment 21

1. At the instant a ball is thrown horizontally over a level range, a ball held at the side of the first is released and drops to the ground. If air resistance is neglected, which ball strikes the ground first?
2. Neglecting air resistance, if you throw a ball straight up with a speed of 20m/s, how fast will it be moving when you catch it?
3. A body describes a circumference of radius 2m and the motion is uniform. It does 2 rotations in 6s. If $\pi^2 = 9.86$, find the centripetal acceleration.
4. A moving body is in uniform circular motion. The radius of the circle is 25m. Assuming that the acceleration equals $9m/s^2$, find the angular velocity.
5. How many rotations a wheel of 3.20m diameter does in one minute. Assuming that the linear speed is equal to 16m/s?
6. A ball at the end of a string is swinging in a horizontal circle of radius 1.15m. The ball makes exactly 2.00 revolutions in a second. What is its centripetal acceleration?
7. The wheel of an engine of 4m diameter does 90 rotations per minute. Calculate:

   h) The linear speed.

   i) The angular speed.

   j) The centripetal acceleration.
8. What is the angular velocity of the earth around its axis? What is the linear velocity of a point situated at the equator? The radius of the earth is supposed to be 6400km.

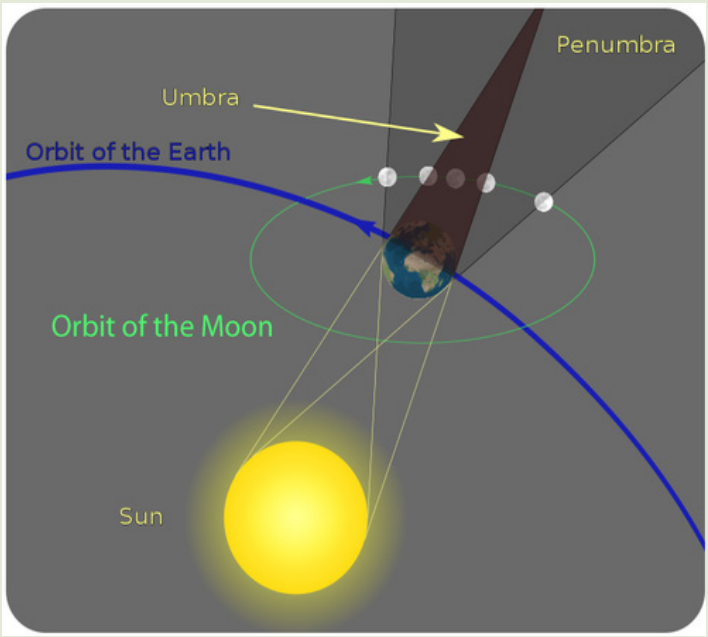
UNIT 22:

GENERAL STRUCTURE OF SOLAR SYSTEM

Key Unit competence: To illustrate and describe the general structure of the solar system.

Introductory Activity 22

Observe the following illustration and answer the questions



Questions:

a) What do you see in the illustration above?

b) Can you suggest how the bodies identified in the illustration above move? Explain to support your decision?

c) Explain how the earth receives light during the night?

d) What causes day and night? Explain your answer with possible illustrations.

e) How do Kepler talk about the general structure of the solar system? Justify your arguments with possible illustrations.

22.1. Astronomical scales

Activity 22.1

1. Student-teachers interpret reality from their perspective of the world around them. As a result their everyday thinking about space and time is often limited to local conditions; often at most perhaps hundreds of kilometre or decades of years. Vast distances and times are central ideas in 'the changing Earth and its place in space'.

   Questions: Read the paragraph above and try to answer the following questions;

   a) Explain how these ideas are very difficult for student-teachers to grasp.

   b) Explain why it has very important implications for their learning of ideas about distances in space or very long spans of time.

22.1.1. Origin of the solar system

Our solar system originated from a giant cloud of gas and debris left from the explosion of stars five billion years ago. Everything in the universe and on Earth is made of this material. Scientific evidence implies that some rock near the Earth's surface is several billion years old.

The Earth's surface is shaped by water (including ice) and wind over very long times. The change is so slow that it is hard to observe rock erosion and soil formation. Biological evolution is also difficult to observe due to the very slow changes that occur.

22.1.2. Astronomical scale

Astronomy is the study of the universe, and when studying the universe, we often deal with unbelievable sizes and unfathomable distances. To help us get a better understanding of these sizes and distances, we can put them to scale.

Scale is the ratio between the actual object and a model of that object. Some common examples of scaled objects are maps, toy model kits, and statues. Maps and toy model kits are usually much smaller than the object it represents, whereas statues are normally larger than its analog.

Our solar system is **immense** in size. We think of the planets as revolving around the sun but rarely consider how far each planet is from the sun or from each other. Furthermore, we fail to appreciate the even greater distances to

the other stars. Astronomers refer to the distance from the sun to the Earth as one“astronomical unit” or AU = approximately 150 million kilometres. This unit provides an easy way to calculate the distances of the other planets from the sun and build a scale model with the correct relative distances.

Viewed from Earth it is difficult to gauge the scale of the universe but astrophysicists have developed techniques to help to do this. Stars and galaxies are so far away than a new unit of distance measurement, **the light-year** (ly),is often used. For light travelling at 3 x 108m/s, the distance traveled in oneyear is:1 ly = (3 x 108 m/s) x (365 x 24 x 60 x 60 s) = 9.46 x 1015 m.

For specifying distances to the Sun and Moon, we usually use metres or kilometres, but we could specify them in terms of light. The Earth-Moon distance is 384 000 km, which is 1.28 light-seconds. The Earth-Sun distance is 1.5 x 1011 m, or 150,000,000 km; this is equal to 8.3 light-minutes. Far out in our solar system, the ninth planet, Pluto, is about 6 x 109km from the Sun, or 6 x 10-4 ly. The nearest star tous, other than the sun, is Proxima Centauri, about 4.3 ly away. (Note that the nearest star is about 10,000 times farther from us than the outer reaches of our solar system.)

The Milky Way or our Galaxy is about 100 000 ly across; our sun is located on one

of the spiral arms of the galaxy at a distance of 28,000 ly from the galactic centre.

Application activity 22.1

1. Construct a time line to gain perspective of the vastness of time and our lifetime in reference to it. Examples of timelines are ‘the origins of the solar system’ and ‘life on Earth’. If a helpful scale is used a time line representing the age of the Earth can be displayed around the walls of their classroom.
2. Construct a scale model of the solar system with both distances and size of planets to scale. If an orange is used to represent the size of the sun then the model of the solar system should fit into the size of an average school oval.

22.2. Sun-earth-moon system: eclipses and phases of the moon

Activity 22.2

From our perspective, the three objects that have the greatest impact on our lives are the Earth, Sun, and Moon. The Earth, of course, is the planet beneath our feet. Without it, well, we wouldn't have anything at all. The Sun warms our planet, and with the Moon, creates the tides. Interpret the figure below and try to answer the questions that follow:

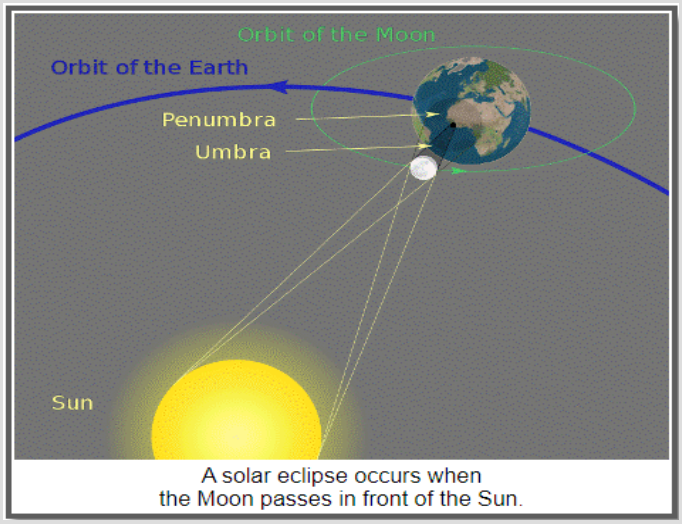


Fig 22.1: A solar eclipse occurs when the Moon passes in front of the Sun.

Questions:

c) Describe how Earth's movements affect seasons and cause day and night.

d) Explain solar and lunar eclipses.

e) Describe the phases of the Moon and explain why they occur.

f) Explain how movements of the Earth and Moon affect Earth's tides

Eclipse, in astronomy is the obscuring of one celestial body by another, particularly that of the sun or a planetary satellite. Two kinds of eclipses involve the earth: those of the moon, or lunar eclipses; and those of the sun, or solar eclipses. A lunar eclipse occurs when the earth is between the sun and the moon and its shadow darkens the moon. A solar eclipse occurs when the moon is between the sun and the earth and its shadow moves

across the face of the earth.

22.2.1.Lunar Eclipses

The earth, lit by the sun, casts a long, conical shadow in space. At any point within that cone the light of the sun is wholly obscured.

A total lunar eclipse occurs when the moon passes completely into the umbra. If it moves directly through the centre, it is obscured for about 2 hours. If it does not pass through the centre, the period of totality is less and may last for only an instant if the moon travels through the very edge of the umbra.

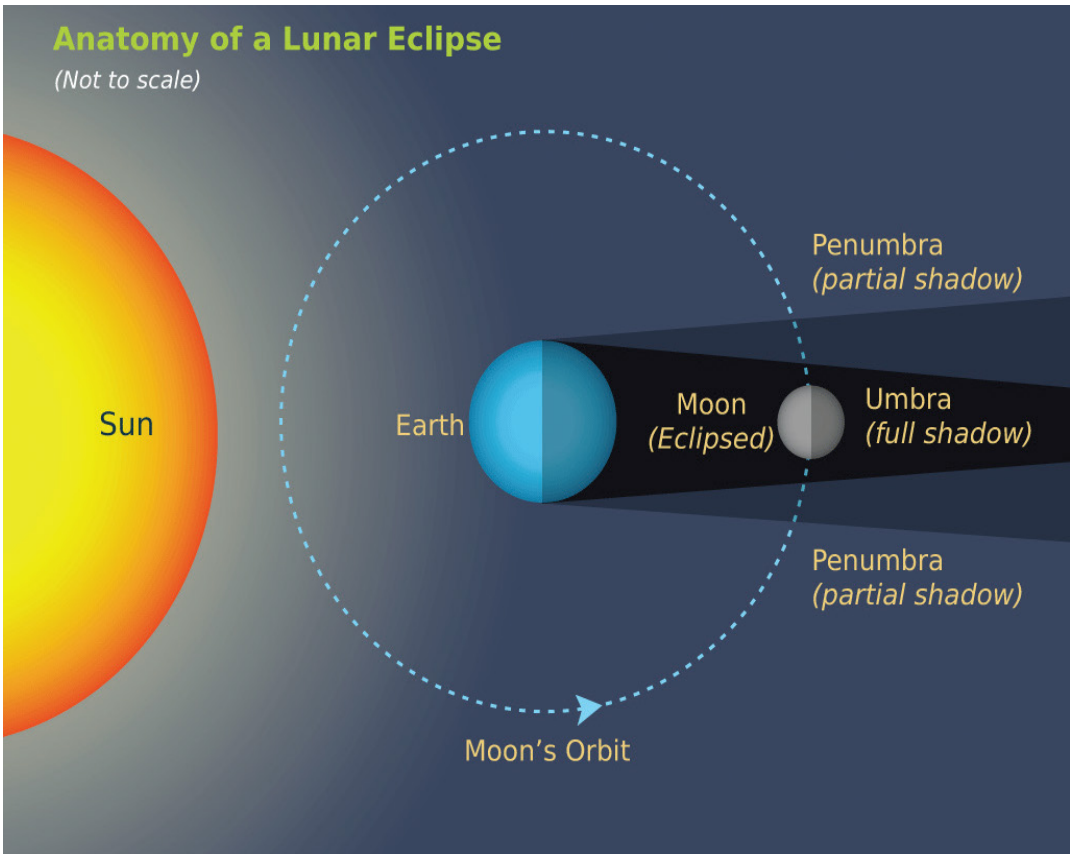


Figure 22.2: Lunar eclipse

A partial lunar eclipse occurs when only a part of the moon enters the umbra and is obscured. The extent of a partial eclipse can range from near totality, when most of the moon is obscured, to a slight or minor eclipse, when only a small portion of the earth's shadow is seen on the passing moon. Historically,the view of the earth's circular shadow advancing across the face of the moon was the first indication of the shape of the earth.

22.2.2.Solar eclipses

A solar eclipse occurs when the Moon passes in front of the Sun causing a shadow to fall on certain portions of the Earth. The eclipse is not seen from every place on Earth, but only from the locations where the shadow falls.

In areas outside the band swept by the moon's umbra but within the penumbra, the sun is only partly obscured, and a partial eclipse occurs.

Types of Solar Eclipses
Depending on what part of the shadow you are located in, there are three types of eclipses:

- **Total** - A total eclipse is where the Sun is covered completely by the Moon. The portion of the Earth that is in the umbra experiences a total eclipse.
- **Annular** - An annular eclipse is when the Moon covers the Sun, but the Sun can be seen around the edges of the Moon. An annular eclipse occurs when the viewer is within the antumbra.
- **Partial** - A partial eclipse is when only a portion of the Sun is blocked by the Moon. It occurs when the observer is within the penumbra.

Caution! Don't Look at A Solar Eclipse. We should warn you here to never look directly at a solar eclipse. Even though it appears darker, the harmful rays of the Sun can still damage your eyes.

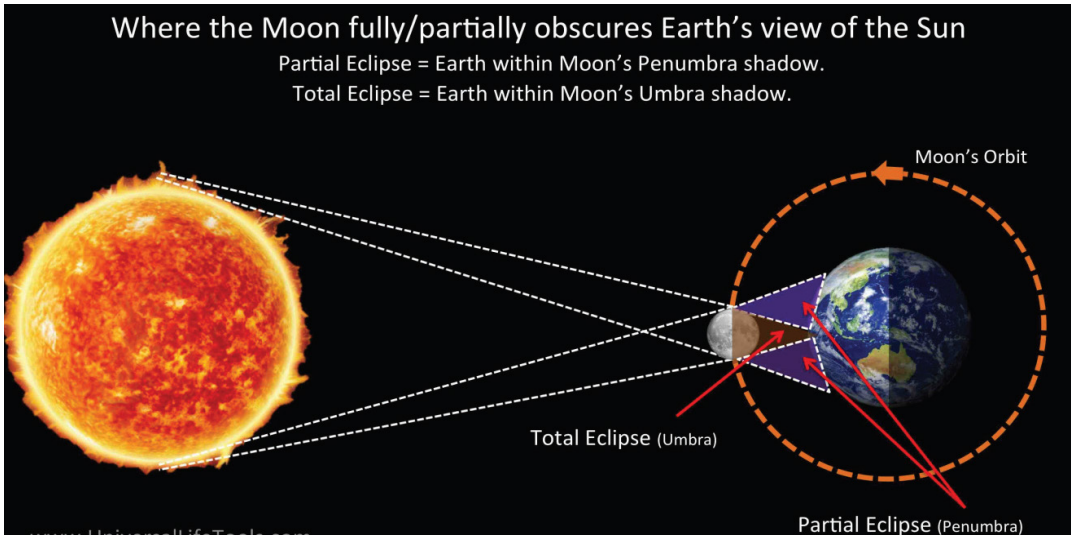


Figure 22.3: Solar eclipse

22.2.3. Phases of the moon

Like everything in the solar system except the Sun, the Moon does not produce any light of its own. It only reflects sunlight. As the Moon moves around Earth, different portions of the satellite are illuminated. This causes the phases of the Moon, so that our view of the Moon goes from fully lit to completely dark and back again.

There are different phases of the Moon that make it appear a little different every day, but it looks the same again about every four weeks. The Moon can sometimes be seen at night and sometimes during the day.

Phases of the Moon

- The Moon is full when Earth is between the Moon and the Sun and the Moon's nearside is entirely lit.
- The Moon is at first quarter phase about one week later, when the Moon appears as a half-circle. Only half of the Moon's lit surface is visible from Earth.
- The Moon is in a new moon phase when the Moon moves between Earth and the Sun and the side of the Moon facing Earth is completely dark. Earth observers may be able to just barely see the outline of the new moon because some sunlight reflects off the Earth and hits the moon.
- Before and after the quarter-moon phases are the gibbous and crescent phases. During the **gibbous** moon phase, the moon is more than half lit but not full. During the **crescent** moon phase, the moon is less than half lit and is seen as only a sliver or crescent shape.

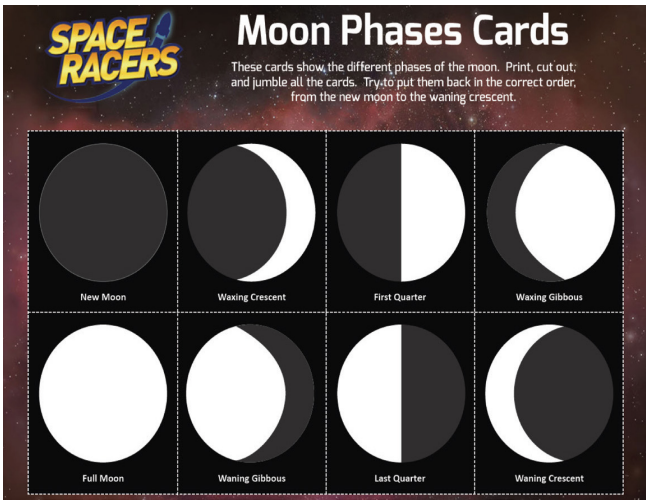


Figure 22.4: Moon phase cards

One revolution of the Moon around Earth takes a little over 27 days 7 hours. The Moon rotates on its axis in this same period of time, so the same face of the Moon is always presented to Earth. Over a period, a little longer than 29 days 12 hours, the Moon goes through a series of phases, in which the amount of the lighted half of the Moon we see from Earth changes. These phases are caused by the changing angle of sunlight hitting the Moon. (The period of phases is longer than the period of revolution of the Moon, because the motion of Earth around the Sun changes the angle at which the Sun's light hits the Moon from night to night).

Application activity 22.2

1. What is a solar eclipse?

   a) When the Moon passes in front of the Sun

   b) When the Earth casts a shadow on the Moon

   c) When the Sun is blocked by another planet

   d) All of the above

   e) None of the Above

2. What do we call the area of a solar eclipse where only a portion of the Moon is in front of the Sun?

   a) Umbra b) Antumbra c) Penumbra.

3) What do we call the area of a solar eclipse where the Moon covers the Sun, but the outline of the Sun can still be seen?

   a) Umbra b) Antumbra c) Penumbra

4) What do we call the area of a solar eclipse where the Moon completely covers the Sun?

   a) Umbra b) Antumbra c) Penumbra

5) What type of solar eclipse occurs when only a portion of the Sun is blocked by the Moon?

   a) Total b) Lunar c) Special d) Annular e) Partial.

6) What type of solar eclipse occurs when the Sun is completely covered by the Moon?

   a) Total b) Lunar c) Special d) Annular e) Partial.

7) What is a lunar eclipse?

a) When the Moon passes in front of the Sun.

b) When the Earth casts a shadow on the Moon.

c) When the Sun is blocked by another planet.

d) All of the above.

e) None of the Above.

8) What color will the moon sometimes appear during a lunar eclipse?

a) Green b) Blue c) Yellow d) Red e) purple.

9) Around how long can a solar eclipse last?

a) 1.5 minutes b) 7.5 minutes c) 30minutes d) 1.5 hours e) 7.5 hours.

10) True or False: Looking directly at the Sun during a solar eclipse can damage your eyes.

22.3. Solar system

Activity 22.3

Observe and interpret the image below and try to explain the structure of the solar system based on the inner and outer planets.

Figure 22.5: Solar system planets and dwarf planets

Solar System is constituted by the Sun and everything that orbits the Sun,including the planets and their satellites, the dwarf planets, asteroids, and comets, and interplanetary dust and gas...

22.3.1. Inner planets and outer planets

In our Solar System, astronomers often divide the planets into two groups: **the inner planets and the outer planets**. The inner planets are closer to the Sun and are smaller and rockier. The outer planets are further away, larger and made up mostly of gas.

The inner planets (in order of distance from the sun, closest to furthest) are Mercury, Venus, Earth and Mars. After an asteroid belt come the outer planets, Jupiter, Saturn, Uranus and Neptune. The interesting thing is, in some other planetary systems discovered, the gas giants are actually quite close to the sun. This makes predicting how our Solar System formed an interesting exercise for astronomers. Conventional wisdom is that the young Sun blew the gases into the outer fringes of the Solar System and that is why there are such large gas giants there. However, some extra-solar systems have “hot Jupiters” that orbit close to their Sun.

a. The Inner Planets

The four inner planets are called terrestrial planets because their surfaces are solid (and, as the name implies, somewhat similar to Earth — although the term can be misleading because each of the four has vastly different environments). They’re made up mostly of heavy metals such as iron and nickel, and have either no moons or few moons. Below are brief descriptions of each of these planets based on this information from National Aeronautic and Space Authority of the USA (NASA).

Mercury

Mercury is the smallest planet in our Solar System and also the closest. It rotates slowly (59 Earth days) relative to the time it takes to rotate around the sun (88 days). The planet has no moons, but has a tenuous atmosphere (exosphere) containing oxygen, sodium, hydrogen, helium and potassium. The NASA MESSENGER (Mercury Surface, Space Environment, Geochemistry, and Ranging) spacecraft is currently orbiting the planet.

Figure 22.6: The terrestrial planets of our Solar System at approximately relative sizes. From left: Mercury, Venus, Earth and Mars

Venus

Venus was once considered a twin planet to Earth, until astronomers discovered its surface is at a lead-melting temperature of 900 degrees Fahrenheit (480 degrees Celsius). The planet is also a slow rotator, with a 243-day long Venusian day and an orbit around the sun at 225 days. Its atmosphere is thick and contains carbon dioxide and nitrogen. The planet has no rings or moons and is currently being visited by the European Space Agency's Venus Express spacecraft.

Earth

Earth is the only planet with life as we know it, but astronomers have found some nearly Earth-sized planets outside of our solar system in what could be habitable regions of their respective stars. It contains an atmosphere of nitrogen and oxygen, and has one moon and no rings. Many spacecraft circle our planet to provide telecommunications, weather information and other services.

Mars

Mars is a planet under intense study because it shows signs of liquid water flowing on its surface in the ancient past. Today, however, its atmosphere is a wispy mix of carbon dioxide, nitrogen and argon. It has two tiny moons (Phobos and Deimos) and no rings. A Mars day is slightly longer than 24 Earthhours and it takes the planet about 687 Earth days to circle the Sun. There's a small fleet of orbiters and rovers at Mars right now, including the large NASA Curiosity rover that landed in 2012.

b. The Outer Planets

Sometimes called **Jovian planets** or gas giants are huge planets swaddled in gas. They all have rings and all of plenty of moons each.

Despite their size, only two of them are visible without telescopes: Jupiter and Saturn. Uranus and Neptune were the first planets discovered since antiquity, and showed astronomers the solar system was bigger than previously thought. Below are brief descriptions of each of these planets based on this information from NASA.

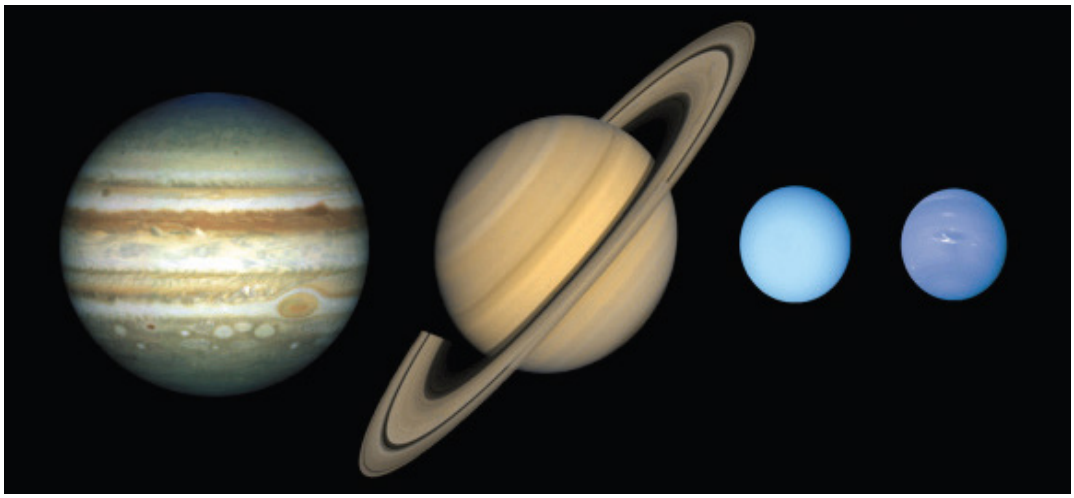


Figure 22.7: The outer planets of our Solar System at approximately relative sizes. From left: Jupiter, Saturn, Uranus and Neptune

Uranus was first discovered by William Herschel in 1781. The planet's day takes about 17 Earth hours and one orbit around the Sun takes 84 Earth years.

Its mass contains water, methane, ammonia, hydrogen and helium surrounding a rocky core. It has dozens of moons and a faint ring system. There are no spacecraft slated to visit Uranus right now; the last visitor was Voyager 2 in 1986.

Jupiter

Jupiter is the largest planet in our Solar System and spins very rapidly (10 Earth hours) relative to its orbit of the sun (12 Earth years). Its thick atmosphere is mostly made up of hydrogen and helium, perhaps surrounding a terrestrial core that is about Earth's size. The planet has dozens of moons, some faint rings and a Great Red Spot, a raging storm happening for the past 400 years at least (since we were able to view it through telescopes). NASA's Juno spacecraft is en route and will visit there in 2016.

Saturn

Saturn is best known for its prominent ring system, seven known rings with well-defined divisions and gaps between them.

How the rings got there is one subject under investigation. It also has dozens of moons. Its atmosphere is mostly hydrogen and helium, and it also rotates quickly (10.7 Earth hours) relative to its time to circle the Sun (29 Earth years). Saturn is currently being visited by the Cassini spacecraft, which will fly closer to the planet's rings in the coming years.

Uranus

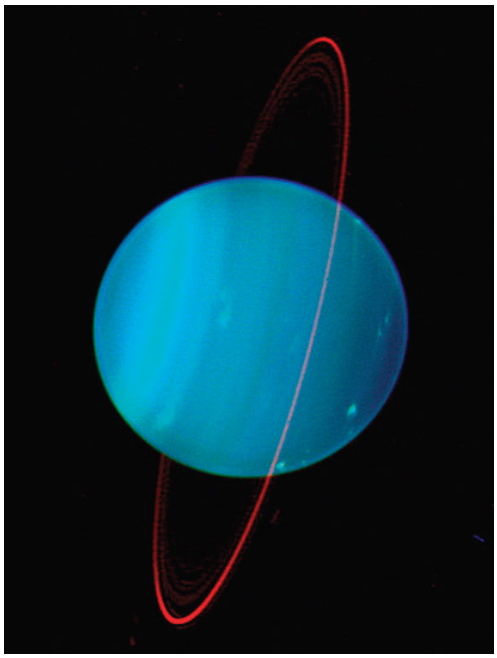


Figure 22.8: Near-infrared views of Uranus reveal its otherwise faint ring system, highlighting the extent to which it is tilted

Neptune

Neptune is a distant planet that contains water, ammonia, methane, hydrogen and helium and a possible Earth-sized core. It has more than a dozen moons and six rings. The only spacecraft to ever visit it was NASA's Voyager 2 in 1989.

Comets

Comet, small icy body in space that sheds gas and dust. Like rocky asteroids, icy comets are ancient objects left over from the formation of the solar system about 4.6 billion years ago. Some comets can be seen from Earth with the unaided eye.

Figure 22.9: Comet

Comets typically have highly elliptical (oval-shaped), off-centre orbits that swing near the Sun. When a comet is heated by the Sun, some of the ice on the comet's surface turns into gas directly without melting. The gas and dust freed from the ice can create a cloud (coma) around the body (nucleus) of the comet.

More gas and dust erupt from cracks in the comet's dark crust. High-energy charged particles emitted by the Sun, called the solar wind, can carry the gas and dust away from the comet as a long tail that streams into space. Gas in the tail becomes ionized and glows as bluish plasma, while dust in the tail is lit by sunlight and looks yellowish. This distinctive visible tail is the origin of the word comet, which comes from Greek words meaning "long-haired star."

Humans have observed comets since prehistoric times. Comets were long regarded as supernatural warnings of calamity or signs of important events.

Astronomers and planetary scientists now study comets for clues to the chemical makeup and early history of the solar system, since comets have been in the deep-freeze of outer space for billions of years. Materials in comets may have played a major role in the formation of Earth and the origin of life. Catastrophic impacts by comets may also have affected the history of life on Earth, and they still pose a threat to humans.

Meteorites

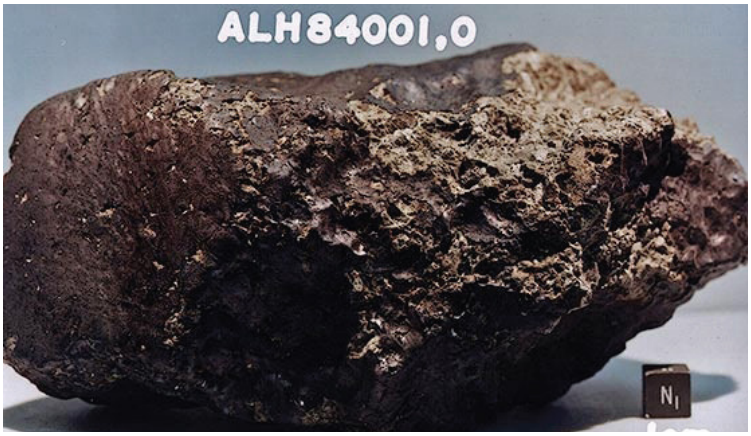


Figure 22.10: Meteorite from Mars

A meteorite is a **rock from outer space**; it's a piece of rock that has reached Earth from outer space. It can also be defined as **a fiery mass of rock fromspace**, a mass of rock from space that burns up after entering the Earth's atmosphere.

Meteorite, meteor that reaches the surface of Earth or of another planet before it is entirely consumed. Meteorites found on Earth are classified into types, depending on their composition: **irons**, those composed chiefly of iron, a small percentage of nickel, and traces of other metals such as cobalt; **stones**, stony meteors consisting of silicates; and **stony irons**, containing varying proportions of both iron and stone.

Although most meteorites are now believed to be fragments of asteroids or comets, recent geochemical studies have shown that a few Antarctic stones came from the Moon and Mars, from which they presumably were ejected by the explosive impact of asteroids. Asteroids themselves are fragments of planetesimals, formed some 4.6 billion years ago, while Earth was forming. Irons are thought to represent the cores of planetesimals, and stones (other than the aforementioned Antarctic ones) the crust. Meteorites generally have a pitted surface and fused, charred crust. A meteorite that landed in Texas in 1998 was found to have water trapped in its rock crystals. The discovery helped scientists theorize about whether water exists in other parts of the solar system.

The largest known meteorite, estimated to weigh about 60 metric tons, is situated at Hoba West near Grootfontein, Namibia. The next largest, weighing more than 31 metric tons, is the Ahnighito (the Tent); it was discovered, along with two smaller meteorites, in 1894 near Perlernerit (Cape York), Greenland, by American explorer Robert Edwin Peary.

Composed chiefly of iron, the three masses had long been used by the Inuit as a source of metal for the manufacture of knives and other weapons.

Asteroids

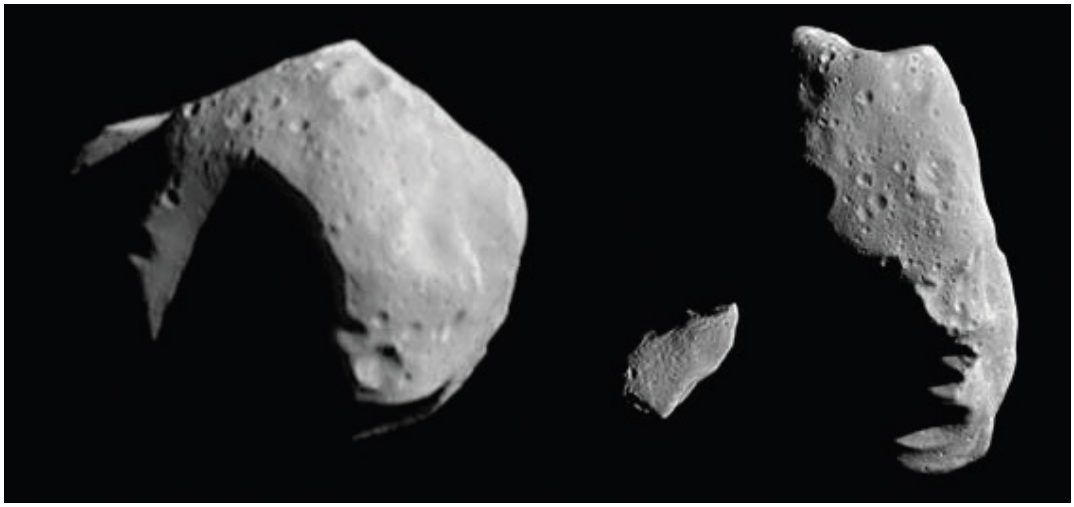


Figure 22.11: Three asteroids

Asteroid, small rocky or metallic body that orbits the Sun. Hundreds of thousands of asteroids exist in the solar system. Asteroids range in size from a few metres to over 500km wide. They are generally irregular in shape and often have surfaces covered with craters. Like icy comets, asteroids are primitive objects left over from the time when the planets formed, making them of special interest to astronomers and planetary scientists.

On the figure 12.10, Asteroid Mathilde, left, is the third and the largest asteroid ever to be viewed at close range. The Near Earth Asteroid Rendezvous(NEAR) spacecraft flew by Mathilde in late June 1997. Asteroids Gaspra and Ida, centre and right, photographed by the Galileo orbiter in 1991 and 1993, respectively, are smaller and more oblong-shaped than Mathilde. The three asteroids are partially obscured by shadows.

Most asteroids are found between the orbits of the planets Mars and Jupiter in a wide region called the **asteroid belt**. Scientists think Jupiter's gravity prevented rocky objects in this part of the solar system from forming into a large planet. The giant planet Jupiter's gravity also helped throw objects out of the asteroid belt. The hundreds of thousands of asteroids now in the asteroid belt represent only a small fraction of the original population.

Figure 22.12: Asteroid Collision with Earth

Thousands of asteroids have orbits that lie outside the asteroid belt. Some of these asteroids have paths that cross the orbit of Earth. Many scientists think that an asteroid that hit Earth 65 million years ago caused the extinction of the dinosaurs. Because asteroids can pose a danger to people and other life on Earth, astronomers track asteroids that come near our planet. Space scientists are also studying ways to deflect or destroy an asteroid that might strike Earth in the future.

Many scientists believe that a large asteroid or comet struck Earth about 65 million years ago, changing the Earth's climate enough to kill off the dinosaurs.

Application activity 22.3

1. Which of the following is true about asteroids?

   a) They are made of rock and metal b) The orbit the Sun c) They are in outer space d) All of the above e) None of the Above.

2. According to the article, what are most asteroids shaped like?

   a) Ball b) Box c) Potato d) Triangle. e) Orange.

3. What two metals are metallic asteroids mostly made of?

   a) Iron and nickel b) Silver and gold c) Magnesium and copper d) Aluminum and zinc e) Lead and iron

4. What is the most common type of asteroid?

   a) Metallic b) Carbon c) Rocky d) Gas e) Round

5. The asteroid belt is located in orbit between what two planets?

   a) Earth and Mars b) Jupiter and Saturn c) Venus and Mercury d) Neptune and Saturn e) Mars and Jupiter.

6. What is the largest asteroid in the Solar System?

   a) Vesta b) Ceres c) pallas d) hygiea

7. What is the largest of the carbon asteroids?

   a) Vesta b) Ceres c) pallas d) hygiea

8. What is the brightest asteroid when viewed from Earth?

   a) Vesta b) Ceres c) pallas d) hygiea

9. What asteroid is the largest body in the Solar System that is not round?

   b) Vesta b) Ceres c) pallas d) hygiea

10. True or False: Most asteroids that hit the Earth are small enough to explode when they hit the Earth's atmosphere.

22.4. Kepler's Laws

Activity 22.4

1. It has been mentioned that planets move not on circle but follow an elliptical curve. State laws that govern this type of motion.
2. How do you relate the period and the mean radius of planet's orbit?

Johannes Kepler proposed three laws of planetary motion. Kepler was able to summarize the carefully collected data of his mentor with three statements that described the motion of planets in a sun-centered solar system.

Kepler's efforts to explain the underlying reasons for such motions are no longer accepted; nonetheless, the actual laws themselves are still considered an accurate description of the motion of any planet and any satellite.

Kepler's three laws of planetary motion can be described as follows:

22.4.1. First Kepler's law

Statement

"The path of the planets about the sun is elliptical in shape, with the center of the sun being located at one focus." (The Law of Ellipses)

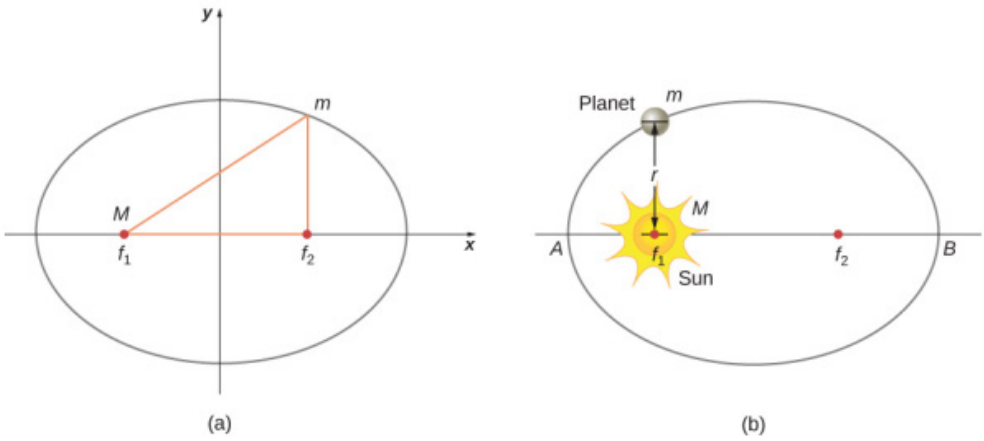


Figure 22.13: Elliptical motions

Figure 22.10 (a) An ellipse is a curve in which the sum of the distances from a point on the curve to two foci (f1 and f2) is a constant. From this definition, you can see that an ellipse can be created in the following way. Place a pin at each focus, and then place a loop of string around a pencil and the pins. Keeping the string taught, move the pencil around in a complete circuit. If the two foci occupy the same place, the result is a circle—a special case of an ellipse.

Figure 22.10. (b) For an elliptical orbit, if m << M , then m follows an elliptical path with M at one focus. More exactly, both m and M move in their own ellipse about the common center of mass.

22.4.2. Kepler's Second Law

Kepler's second law states that"A line segment joining a planet and the Sun sweeps out equal areas during equal intervals of time."

Therefore, an imaginary line drawn from the center of the sun to the center of the planet will sweep out equal areas in equal intervals of time. (The Law of Equal Areas).

Consider Figure 22.14 The time it takes a planet to move from position A to B, sweeping out area A1, is exactly the time taken to move from position C to D, sweeping area A2, and to move from E to F, sweeping out area A3. These areas are the same: A1 = A2 = A3.

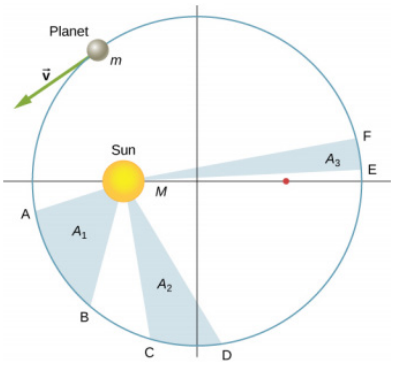


Figure 22.14: *The shaded regions shown have equal areas and represent the same time interval*

Comparing the areas in the figure and the distance traveled along the ellipse in each case, we can see that in order for the areas to be equal, the planet must speed up as it gets closer to the Sun and slow down as it moves away. But we will show that Kepler's second law is actually a consequence of the conservation of angular momentum, which holds for any system with only radial forces.

22.4.3. Kepler's Third Law (The Law of Harmonies)

Statement

"The Square of the orbital period of a planet is proportional to the cube of the semi-major axis of its orbit".

In Satellite Orbits and Energy, we derived Kepler's third law for the special case of a circular orbit.

Application activity 22.4

1. Write in symbol the law of harmonies
2. Kepler's second law is the consequence of which physical quantity. Write the symbolical definition of that physical quantity.
3. Explain the law of equal area.

22.5. Star Patterns: Constellations

Activity 22.5

1. Research on star patterns and do report about constellations.
2. Research on uses of constellation.

Constellations

A constellation is a group of visible stars that form a pattern when viewed from Earth. The pattern they form may take the shape of an animal, a mythological creature, a man, a woman, or an inanimate object such as a microscope, a compass, or a crown.

How many constellations are there? The sky was divided up into 88 different constellations in 1922. This included 48 ancient constellations listed by the Greek astronomer Ptolemy as well as 40 new constellations.

Star Maps

The 88 different constellations divide up the entire night sky as seen from all around the Earth. Star maps are made of the brightest stars and the patterns that they make which give rise to the names of the constellations. The maps of the stars represent the position of the stars as we see them from Earth. The stars in each constellation may not be close to each other at all. Some of them are bright because they are close to Earth while others are bright because they are very large stars.

Hemispheres and Seasons

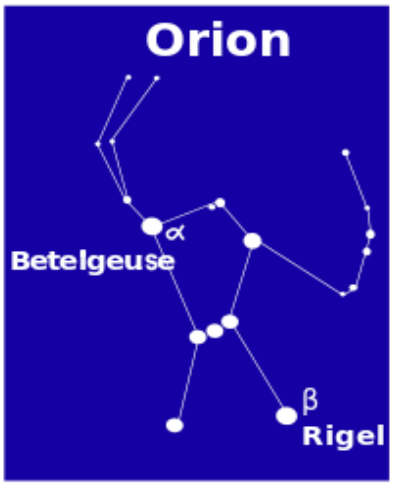
Not all of the constellations are visible from any one point on Earth. The star maps are typically divided into maps for the northern hemisphere and maps for the southern hemisphere. The season of the year can also affect what constellations are visible from where you are located on Earth.

Famous constellations

Here are a few of the more famous constellations:

Orion

Orion is one of the most visible constellations. Because of its location, it can be seen throughout the world. Orion is named after a hunter from Greek mythology. Its brightest stars are Betelgeuse and Rigel.



Constellation Orion

Figure 22.15: Constellation Orion

Ursa Major

Ursa Major is visible in the northern hemisphere. It means “Larger Bear” in Latin. The Big Dipper is part of the Ursa Major constellation. The Big Dipper is often used as a way to find the direction north.

Ursa Minor

Ursa Minor means “Smaller Bear” in Latin. It is located near Ursa Major and also has the pattern of a small ladle called the Little Dipper as part of its larger pattern.

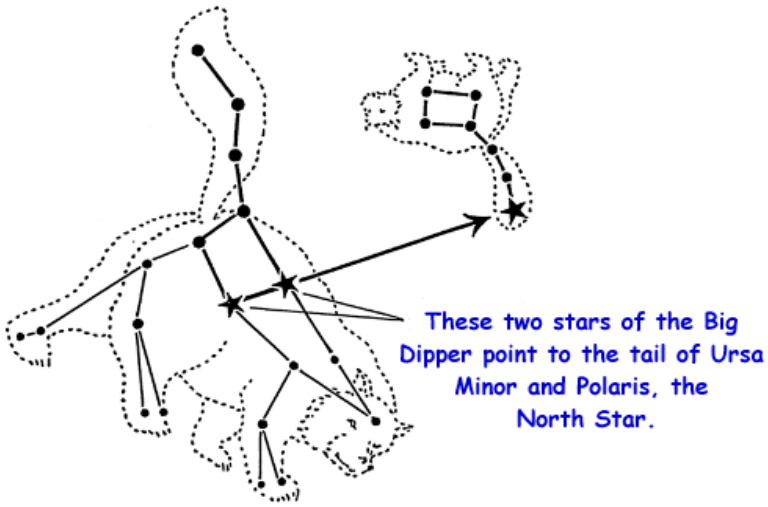


Fig 22.16: Ursa Minor

Draco constellation

The Draco constellation can be viewed in the northern hemisphere. It means "dragon" in Latin and was one of the 48 ancient constellations.

Pegasus

The Pegasus constellation is named after the flying horse by the same name from Greek mythology. It can be seen in northern sky.

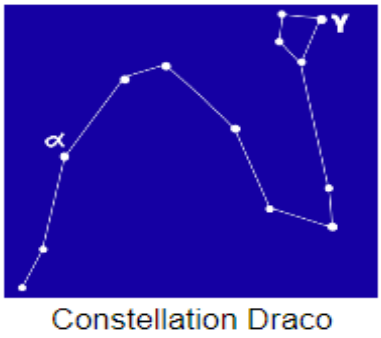


Figure 22.17: Draco Constellation

The zodiac

The Zodiac constellations are the constellations that are located within a band that is about 20^0 wide in the sky. This band is considered special because it s the band where the sun, the moon and the planets all move.

There are 13 Zodiac constellations. Twelve of these are also used as signs for the zodiac calendar and astrology. These are Capricornus, Aquarius, pisces,

Aries, Taurus, Gemini, Cancer, Leo, Virgo, Libra, Scorpius, Sagittarius and Ophiuchus.

Uses for Constellations

Constellations are useful because they can help people to recognize stars in the sky. By looking for patterns, the stars and locations can be much easier to spot. The constellations had uses in ancient times. They were used to help keep track of the calendar. This was very important so that people knew when to plant and harvest crops. Another important use for constellations was navigation. By finding Ursa Minor it is fairly easy to spot the North Star (Polaris). Using the height of the North Star in the sky, navigators could figure out their latitude helping ships to travel across the oceans.

Application activity 22.5

1. What is a constellation?
   a) A group of stars from the same galaxy.
   b) A group of stars that are physically close to each other.
   c) A group of visible stars that make a pattern when viewed from Earth.
   d) All of the above.
   e) None of the Above.

2.How many constellations are there?
   a) 12 b) 22 c) 44 d) 88 e) 120

3. What constellation was named after a flying horse?
   a) Orion b) Ursa Minor c) Ursa Major d) Draco e) Pegasus

4. What constellation has the Little Dipper as part of its pattern?
   a) Orion b) Ursa Minor c) Ursa Major d) Draco e) Pegasus

5. What constellation was named after a hunter from Greek mythology?
   a) Orion b) Ursa Minor c) Ursa Major d) Draco e) Pegasus

6. What constellation has the Big Dipper as part of its pattern?
   a) Orion b) Ursa Minor c) Ursa Major d) Draco e) Pegasus

7. True or False: All the constellations are visible from anywhere on Earth throughout the year.

8. What group of constellations are used in astrology?

   a) Ursa constellations b) Predictive constellations c) Northern constellations d) Zodiac constellations e) Ancient constellations.

9. How are constellations useful?

   a) To help locate stars b) To keep track of the calendar c) To navigate d) To know when to plant crops. e) All of the above

Skills lab 22

1. With help of internet, visit youtube.com and watch “Seasons- what causes seasons”. watch again the video , summarize the causes and share the findings with your classmate and schoolmates.
2. Using local materials make a demonstration explaining Moon phases.

End unit assessment 22

1. Why does the moon seem to change its shape every night? Why can I see the moon in the daytime?
2. What are the Moon phases in order?
3. What is an approximation of largest meteorite?
4. What is the role of Constellations?
5. Which physical quantity is conserved in the case of law of areas?
6. With the help of diagrams explain types of eclipses.

REFERENCES

Kennedy P. and Sochacki F.(2008), OCR Biology, London,British, Heinemann.

Burton S. Guttman (1983), Understanding Biology, New York, Chicago San Fransisco, Harcout Brace Jovanovich.

Virgilio O.K, James K. and all (2013), Biology for Rwanda, Kampala Uganda.

Kenneth R., Miller and Joseph Levine (1993), Biology

Antony C., Dennis D. and all (2000), Chemistry fifth edition, Calfornia, Pearson Education.

Theodore L., Bruce E. and Julia R. (2003), Chemistry: the central science ninth Edition, New Jersy, Pearson Education.

Nelkson and Parker (2001), Advanced level Physics 7th edition, Heineman.

Duncan T. (2000), Advanced Physics 5th edition, Hodder Education an Hachette UK Compay.

Biology and Health Sciences For Rwanda Schools Senior Two Student's Book by Godfrey Wasswa Benson Mugarura Selah Ndiwa)

Ramsden, E. (2000). A-Level Chemistry 4th Edition. United Kingdom: Nelson Thornes.

Holman, G. Hill. (2000). Chemistry In Context 5th Edition. United Kingdom: Nelson Thornes.

Huhes, P. Cann. (2002). Chemistry for advanced level. London: JOHN MURRAY.

http://biology.kenyon.edu/courses/biol09/tetrahymena/serialdilution3.htm

https://simple.wikipedia.org › wiki › Copper(II)_sulfate

https://www.healthline.com/nutrition/simple-sugars

https://www.biologynotes.site/carbohydrates/

http://www.nutrientsreview.com/carbs/disaccharides.html

http://www.nutrientsreview.com/carbs/polysaccharides.html

http://www.nutrientsreview.com/carbs/polysaccharides.html

https://commons.wikimedia.org/wiki/File:Glycerol_structure.svg

https://www.chegg.com/homework-help/definitions/triglycerides-and-phospholipids-14.

https://www.quora.com/What-are-phospholipids-made-of

http://www.okstate.edu/jgelder/bondpage14.html

http://www.differencebetween.net/science/difference-between-ionic-and-covalent-compounds/#ixzz5yqvqXz16

https://study.com/academy/lesson/effect-of-intermolecular-forces-on-physical-properties.html

https://www.khanacademy.org/test-prep/mcat/chemical-processes/covalent-bonds/a/intramolecular-and-intermolecular-forces

https://www.thoughtco.com/covalent-or-molecular-compound-properties-608495

https://sciencing.com/difference-between-atoms-ions-5595821.html

https://www.adichemistry.com/general/chemicalbond/vsepr/vsepr-theory.html

https://www.askiitians.com/iit-jee-d-and-f-block-elements/position-in-the-periodic-table/