Optional Science

Grade 10

Government of Nepal
Ministry of Education
Curriculum Development Centre
Sanothimi,Bhaktpur

Optional Science

Grade - 10

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PREFACE

The curriculum and curricular materials have been developed and revised on a regular basis with the aim of making the education objective-oriented, practical, relevant and job oriented. It is necessary to instill the feelings of nationalism, national integrity and democratic spirit in the students and equip them with morality, discipline and self reliance, creativity and thoughtfulness. It is essential to develop in them the linguistic and mathematical skills, knowledge of science, information and communication technology, environment, health and population and life skills. It is also necessary to bring in them the feeling of preserving and promoting arts and aesthetics, humanistic norms, values and ideals. It has become the need of the present time to make the students aware of respect for ethnicity, gender, disabilities, languages, religion, cultures, regional diversity, human rights and social values so as to make them capable of playing the role of responsible citizens. This textbook has been developed in line with the Secondary Level Optional Science Curriculum, 2072 by incorporating the recommendations of various education commissions and the feedback obtained from various schools, workshops, seminars and interaction programs attended by the teachers, students and parents.

In bringing out the textbook in this form, the contribution of the Executive Director of the Curriculum Development Centre (CDC) Krishna Prasad Kapri, Dr. Hridhaya Ratna Bajracharya, Umanath Lamsal, Baburam Gautam, Puspa Raj Dhakal, Devraj Gurung, Keshar Khulal, Manumaya Bhattrai. The content of the book was edited by Yubraj Adhikari. Language of the book was edited by Ramesh Prasad Ghimire. The layout and artworks of the book were done by Jayaram Kuikel. CDC extends sincere thanks to all those who have contributed in developing this textbook.

This textbook contains a variety of learning materials and exercises which will help learners to achieve the competency and learning outcomes set in the curriculum. Each unit contains various interesting activities and the content required for meaningful learner engagement and interaction. There is uniformity in the presentation of the activities which will certainly make it convenient for the students. The teachers, students and other stakeholders are expected to make constructive comments and suggestions to make it a more useful learning material.

Curriculum Development Centre

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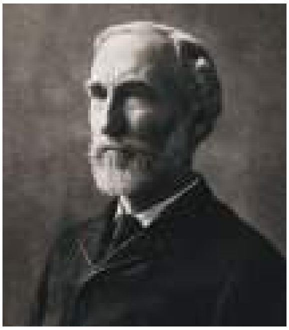
Unit 1

Force

Josiah Willard Gibbs (1839 – 1903) was an American scientist who made important theoretical contributions to physics, chemistry, and mathematics. His work on the applications of thermodynamics was instrumental in transforming physical chemistry into a rigorous inductive science. Together with James Clerk Maxwell and Ludwig Boltzmann, he created statistical mechanics. As a mathematician, he invented modern vector calculus.



Josiah Willard Gibbs (1839-1903)

Introduction

Force is an external agency that brings or tends to bring a body from rest to motion or motion to rest. Push and pull are the examples of force. We use force to perform various types of work. We experience various types of forces in our daily life. They are magnetic force, muscular force, frictional force, electrical force, inter nuclear force, gravitational force, centrifugal force, centripetal force, etc.

In this unit we will discuss in detail the vector and scalar along with their addition and subtraction, escape velocity, centripetal and centrifugal force, centre of gravity and gravitational field intensity.

1.1 Physical quantity

We observe some physical quantities that depend on direction and some other physical quantities that do not depend on direction. For example, the volume of an object i.e. the three-dimensional space that an object occupies does not depend on direction. If a person says Pashupati temple is 12 km from Bhaktapur, this information is not sufficient to reach the temple because the direction is necessary for it. This means, to specify the exact location of a place or an object, it is necessary to have direction.

The quantities that can be measured directly or indirectly are called physical quantities. In other words, measurable quantities are known as physical quantities. From the above discussion, it is clear that the physical quantities are of two types:

- First quantity having magnitude not direction (scalar)
- Second quantity having magnitude as well as direction (vector)

Scalar quantities

Those physical quantities which are entirely described only by magnitude are called scalar quantities. In other words, the physical quantities which are expressed in magnitude only are called scalar quantities. It does not require direction for its understanding. For example, Butwal is about 114 km from Narayangagh. In this example, only magnitude is given but direction is not given. Some of the examples of scalar quantities are length, mass, time, temperature, volume, density, electric charge, etc.

Addition of scalar quantities

The resultant of a scalar quantity is obtained by simple algebraic addition. For example, suppose a mass of a body A is 6 kg and mass of a body B is 10 kg. Here, both the mass of A and B are scalars and the total mass is given by simple addition i.e. 6 kg + 10 kg = 16 kg. Another example of scalar quantity is if we add 20 cm^3 of water with 50 cm^3 of water, then the sum is equal to $20cm^3 + 50cm^3 = 70cm^3$.

Vector quantities

Those physical quantities which are entirely described not only by magnitude but also by direction are called vector quantities. In other words, the physical quantities which are expressed in magnitude as well as direction are called vector quantities. For example, Butwal is about 114 km towards west from Narayangagh. In this example, both the magnitude and direction are given there. Some of the examples of vector quantities are displacement, velocity, acceleration, momentum, force, weight, electric and magnetic field, etc.

Differences between scalars and vectors quantities

Scalars	Vectors
1. They possess only magnitude.	1. They possess both magnitude and direction.
2. They are represented by a number with specified units.	2. They are represented by a letter with an arrow head over it.

Scalars	Vectors
3. They can be added or subtracted algebraically.	3. They can be added or subtracted vectorally.
4. The sum of scalars is always positive number.	4. The sum of vectors may be positive, zero or negative.
5. The product of scalars is a scalar.	5. Scalar product of vectors is a scalar and vector product of vectors is a vector.
6. They cannot easily be plotted on a graph paper.	6. They can easily be plotted on graph paper.

Addition of Vectors

If a single vector that produces the same effect as produced by a number of vectors acting at a time on an object, then the single vector is known as sum or the resultant of those numbers of vectors. For example, if more than one force is acting at a time on an object is displaced in a direction with a force which is different from the individual forces acting on it. This force is the resultant of those forces acting on the object. The direction of displacement may be different from the direction of any individual force which gives a resultant direction.

The process of obtaining the resultant of sum of the vectors is called addition or composition of vectors. The vector quantities are not added like scalars but they can be added geometrically using special rules of vector algebra. In other words, the resultant of any number of vectors acting at a point can be determined either by triangle law or parallelogram law of vectors.

Let us consider the following cases:

Case i: When vectors are acting in the same direction

In this case, the addition of vectors in one dimension is straight forward. We suppose th at a particle is displaced through 6 m in west and again 3 m in the same direction. If these two displacement vectors OP of magnitude 6 m and PQ of magnitude 3 m are represented by a and b, sum (OP + PQ) of the two displacement vector will be a displacement OQ of 9 m due east as shown in fig. 1.1. In vector form, the result is written by

$$\vec{OP} + \vec{PQ} = \vec{OQ}$$

or, $\vec{a} + \vec{b} = \vec{c}$

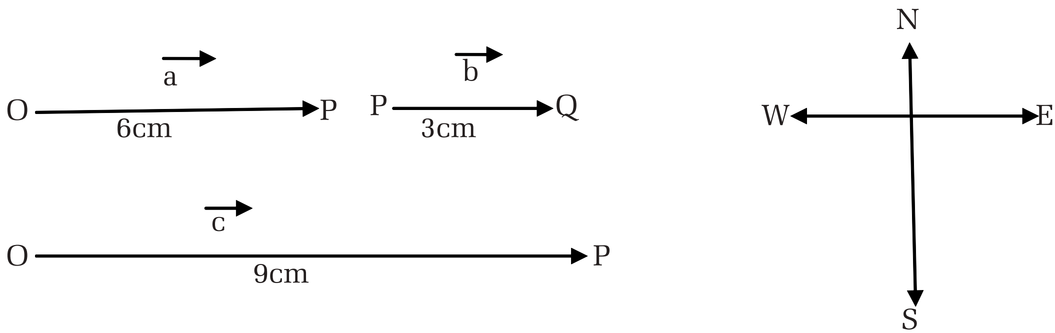


Fig 1.1 Addition of parallel vectors

Case ii: When vectors are acting in the opposite directions

When a particle is displaced 7 m in due west and 4 m in east and if these vectors are added then the result vector $\vec{c}$ = 3 m west i.e., the direction of resultant vector will be the direction of bigger of the two vectors a and b as shown in fig 1.2. This result can also be written as

$\vec{a} + (-\vec{b}) = \vec{c}$

Since, the vector c alone produces the same result as the vectors $\vec{a}$ and $\vec{b}$ together do, we say that $\vec{c}$ is the resultant (or sum) of vectors $\vec{a}$ and $\vec{b}$. Thus, it is clear that vectors are not added as the scalars.

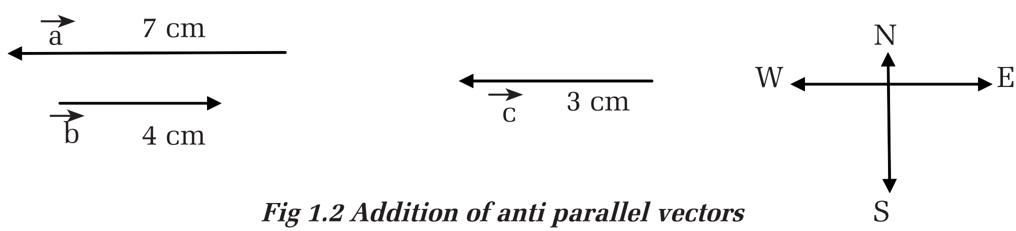


Fig 1.2 Addition of anti parallel vectors

Now, we are going to consider addition of vectors in two dimensions.

Case iii: When two vectors are acting at a right angle.

Let us suppose a man walks a distance of 4 m from $\vec{P}$ to $\vec{Q}$ in the east and then walks a distance of 3 m from $\vec{Q}$ to $\vec{R}$ in the north as shown in the fig. 1.3.

ΔRQP be a right angled triangle so resultant PR can be obtained by Pythagoras theorem $h^2 = p^2 + b^2$.

The two perpendicular displacement vectors $\overrightarrow{PQ}$ and $\overrightarrow{QR}$ are represented by $\vec{a}$ and $\vec{b}$ respectively.

If we join $\vec{P}$ and $\vec{R}$ and measure the length of vector $\vec{PR}$, we will get 5 m. If instead of walking a distance of 4 m from $\vec{P}$ to $\vec{Q}$ and then walking a distance of 3 m from $\vec{Q}$ to $\vec{R}$ (a total distance of 4 + 3 = 7 m), the man walks a distance of 5 m from $\vec{P}$ to $\vec{R}$, he would reach the same destination ($\vec{PR}$). Thus, the displacement of 5 m along PR produces the same effect as the displacement of 4 m along $\vec{PQ}$ and 3 m along $\vec{QR}$ together do. We say that the vector $\vec{PR}$ (or $\vec{c}$) is the resultant (or sum) of two vectors $\vec{PQ}$ (or a) and $\vec{QR}$ (or $\vec{b}$). In vector form, we can have,

$$\vec{PQ} + \vec{QR} = \vec{PR}$$

Or, $\vec{a} + \vec{b} = \vec{c}$

To find out magnitude,

$PR^2 = PQ^2 + QR^2$ $\quad (\therefore h^2 = p^2 + b^2)$

For direction $\tan\theta = \dfrac{QR}{PQ}$ $\quad (\therefore \tan\theta = \dfrac{p}{b})$

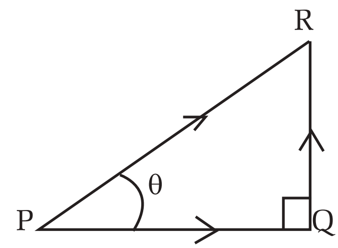


Fig 1.3 Addition of two vectors at right angled to each other

In this way we see that in every case, vectors are not added algebraically but are added geometrically.

Vector Subtraction

Subtraction of a vector from another vector is the same as the addition of a negative vector with another vector.

Let's consider, there are two vectors $\vec{p}$ and $\vec{q}$. $\vec{q}$ is to be subtracted from $\vec{p}$. For this, negative of $\vec{q}$ $(\vec{-q})$ is formed and it is added with p. The difference r is written as $\vec{r} = \vec{p} - \vec{q} = \vec{p} + (\vec{-q})$

The difference or resultant vector $\vec{r}$ is calculated using either triangle law or parallelogram law of vector addition.

1.2 Escape velocity

As we know that the earth pulls every object toward its centre. Due to the gravitational pull of the earth, all the slow moving objects thrown upward from the earth's surface come back to the earth. So, if an object like a rocket is to escape into the space from the earth gravitational grip, then it must have

> ***Do you know?***
>
> A body given less than the escape velocity will fall back towards the surface of the larger body; a body given a velocity equal to or greater than the escape velocity will still be attracted by the larger body, but this force will not be sufficient to cause it to return.

greater velocity so that it aquires sufficient kinetic energy to overcome the earth's gravity force. The minimum velocity which an object (like a rocket) should have in order to overcome the earth's gravity and enter into space is called escape velocity. The escape velocity for all the objects from the earth has been found to be 11.2 kilometers per second (11.2 km/s). This means that an object (like a rocket) should have a minimum velocity of 11.2 km/s to overcome the earth's pull and go into the outer space. It is clear that the escape velocity of an object does not depend upon its mass. It is the same for all the objects, irrespective of their masses. The escape velocity from the earth depends only on the mass of the earth and its radius.

Calculation of escape velocity

We have already studied about the escape velocity of an object from the earth is 11.2 kilometers per second.

Let's consider, if a rocket of mass m and escape velocity v, then its kinetic energy for escaping the earth's pull will be:

Kinetic energy of rocket = $1/2 \ mv^2$ ....(i)

The pulling force exerted by the earth with mass M and radius R on the rocket of mass m is given by:

Gravitational force (F) = $G \times \frac{M \times m}{R^2}$ ....(ii)

(Where G is the gravitational constant)

Now, the work done by the gravitational force of the earth against the rocket can be obtained by multiplying the gravitational force and the distance between the centre of earth and rocket, which is equal to radius of the earth, R. Thus,

Work done by earth on the rocket = Gravitational force x Radius of earth

$= F \times d$

$= G \times \frac{M \times m}{R^2} \times R$

$= gmR \ \left(\because g = \frac{GM}{R^2}\right)$ ....(iii)

Now, $1/2 \ mv^2 = gmR$

Or, $v^2 = 2gR$

$v = \sqrt{2gR}$ ....(iv)

If an object attains escape velocity, but is not directed straight away from the planet, then it will follow a curved path. Although this path does not form a closed shape, it is still considered an orbit. Assuming that gravity is the only significant force in the system, this object's speed at any point in the orbit will be equal to the escape velocity at that point. The shape of the orbit will be a parabola whose focus is located at the center of mass of the planet. An actual escape requires a course with an orbit that does not intersect with the planet, or its atmosphere, since this would cause the object to crash.

> ***Do you know?***
>
> The velocity of escape from the earth at its surface is about 7 miles (11.2 km) per sec, or 25,000 miles per hr; from the moon's surface it is 1.5 miles (2.4 km) per sec; and for a body at the earth's distance from the sun to escape from the sun's gravitation, the velocity must be 26 miles (41 km) per sec.

When there are many gravitating bodies, such as in the solar system, a rocket that travels at escape velocity from one body, say Earth, will not travel to an infinite distance because it needs an even higher speed to escape the Sun's gravity. Near the Earth, the rocket's orbit will appear parabolic, but it will become elliptical around the Sun.

1.3 Centrifugal force and Centripetal force

Centrifugal force

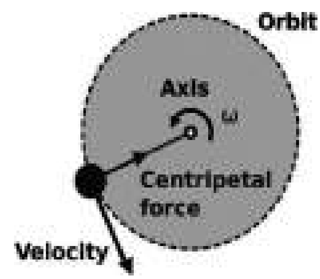
The force that produces the tendency on an object to fly away from the center, when it moves in a curved path is called centrifugal force. It is a fictitious force because it only comes to play when there is a centripetal force. The magnitude of centrifugal force is same as that of centripetal force, mv^2/r and direction is always away from the centre of the circular path. It is not a real force and this force is due to inertial property of the material body.

When a moving bus passes through a curved path, the passengers in the bus are deflected outward from their positions. This is due to the centrifugal force acting on the passengers.

Centripetal force

The force that keeps an object moving with a uniform speed along a circular path is called centripetal force. This force acts along the radius and is directed towards the centre of the circle.

When a body moves in a circle, its direction of motion at any instant is along the tangent to the circle at that instant. As shown in the figure, the direction of the velocity of the body goes in changing continuously and due to this change, there is an acceleration called the centripetal acceleration given as



$$a = \frac{v^2}{r} = rw^2$$

Where, v is linear velocity, ω is angular velocity of the body and r is radius of the circular path.

As F = ma, centripetal force is

F = mass x centripetal acceleration

$= m \times rw^2 = mrw^2$

Do you know?

The formula Huygens wrote for determining centripetal force multiplies an object's mass by the square of its tangential speed and divides the product by its radius of curvature. Newton described the laws of thermodynamics in the publication of his treatise "Principia Mathematica." Huygens described centripetal force as part of larger research into centrifugal force.

This force acts along the radius and is directed towards the centre of the circle.

When a stone is whirled in a circle by means of a string, the tension on it provides the centripetal force, electrostatic force provides the centripetal force for electrons to move around the nucleus and the gravitational force provides the centripetal force for planets to move round the sun.

Direction of centripetal force, centrifugal force and velocity

The centripetal force is directed inwards, from the object to the center of rotation. Technically, it is directed orthogonal to the velocity of the velocity, toward the fixed point of the instantaneous center of curvature of the path.

The centrifugal force is directed outwards; in the same direction as the velocity of the object. For circular motion the velocity at any given point in time is at a tangent to the arc of movement.

Both forces are calculated using the same formula

$$F = ma_c = \frac{mv^2}{r}$$

Where a_c is the centripetal acceleration, m is the mass of the body, moving at velocity v along a path with radius of curvature r.

Some examples of centrifugal and centripetal forces

Some common examples of centrifugal force at work are mud flying off a rotating tire and children feeling a force pushing them outwards while spinning on a round about. A major example of centripetal force is the rotation of satellites around a planet.

Uses of centrifugal and centripetal forces

Knowledge of centrifugal and centripetal forces can be applied to many everyday problems. For example, it is used when designing roads to prevent skidding and improve traction on curves and access ramps. It also allowed for the invention of the centrifuge which separates particles suspended in fluid by spinning test tubes at high speeds.

Comparison chart of centripetal force and centrifugal force

Criteria	Centrifugal force	Centripetal force
Meaning	The force that produces the tendency on an object to fly away from the center, when it moves in a curved path.	The force that keeps an object moving with a uniform speed along a circular path.
Direction	Along the radius of the circle, from the center towards the object.	Along the radius of the circle, from the object towards the center.
Example	Mud flying off a rotating tire; children pushed out on a round about.	Satellite orbiting a planet

1.4 Centre of Gravity

The center of gravity is a geometric property of any object. It is the average location of weight of an object. We can completely describe the motion of any object through space in terms of the translation of the center of gravity of the object from one place to another and the rotation of the object about its center of gravity if it is free to rotate. If an object is confined to rotate about some other point, like a hinge, we can still describe its motion. In flight, both airplanes and rockets rotate about their centers of gravity. A kite, on the other hand, rotates about the bridle point. But the trim of a kite still depends on the location of the center of gravity relative to the bridle point, because for every object the weight always acts through the center of gravity.

> Do you know?
>
> Something with a wide base and low height, such as a Formula 1 car, has a very low center of gravity in relation to the rest of the object. This means it is very stable and a large force must be applied to tip it over.

In general, determining the center of gravity (cg) is a complicated procedure because mass (and weight) may not be uniformly distributed throughout the object. The general case requires, use of calculus. If mass is uniformly distributed, the problem is greatly simplified. If an object has a line (or plane) of symmetry, the centre of gravity lies on the line of symmetry. For a solid block of uniform material, the center of gravity is simply at the average location of the physical dimensions.

For a general shaped object, there is a simple mechanical way to determine the center of gravity. i.e. if we just balance the object using a string or an edge, the point at which the object is balanced is the center of gravity.

> Do you know?
>
> A tall object with a narrow base, such as a bookcase, will have a high center of gravity and thus only a small force applied towards the top of the object is required to topple it over.

1.5 Gravitational field

The region around a mass/planet over which its gravitational force on other masses can be experienced is called the gravitational field of the mass. Theoretically, it extends upto infinity. In our practice, the gravitational field may become too weak to be measured beyond a particular distance.

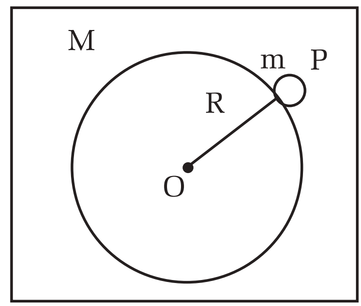
Gravitational field intensity

Gravitational field intensity is also called gravitational field strength. It is defined as the amount of force exerted on each unit mass at a point in space caused by the presence of another object. "Space" includes a region on or near the surface of a planet. The more massive object is usually considered as a source of the field. Gravitational field intensity has the units N/kg as force is in N and mass is in kg and is a vector that points to the source's centre of mass.

$g = GM/R^2$ where $G = 6.67 \times 10^{-11}\ Nm^2/kg^2$, M is mass in kg of the source R is the separation between the center of mass of the source and the point in space.

Let's consider, M is a mass producing gravitational field, m is another mass placed at point P at a distance R from centre of O.

Gravitational force of attraction (F) between them is,



$$F = \frac{GMm}{R^2}$$

$$\therefore I = \frac{F}{m} = \frac{GMm}{R^2 \times m} = \frac{GM}{R^2}$$

I is directed towards the centre of mass M which is producing the field.

Also, $I = \frac{GM}{R^2} = g; \left(\therefore g = \frac{GM}{R^2} \right)$

Hence, gravitational intensity I at a point in gravitational field is numerically equal to the acceleration due to gravity at that point. Unit of I is N kg^{-1} (or ms^{-2})

$$\text{Gravitational field intensity} = \frac{\text{Force experienced}}{\text{Mass}}$$

$$I = \frac{F}{m}$$

In SI system, force and mass are measured in N and kg. So its unit is N/kg.

Numerical Illustration

1. **The mass of the moon is 7.2×10^{22} kg and its radius is 1.7×10^6 m. Calculate the gravitational field intensity at a point on its surface.**

   Solution:

   Mass of the Moon (M) = 7.2×10^{22} kg

   Radius of Earth (R) = 1.7×10^6 m

   The gravitational field intensity (I) due to the Moon of mass M and distance R from its centre is given by:

$$I = \frac{GM}{R^2}$$

$$\text{or, } I = \frac{6.67 \times 10^{-11} \times 7.2 \times 10^{22}}{\left(1.7 \times 10^{6}\right)^2}$$

$$\text{or, } I = \frac{48.02 \times 10^{11}}{2.89 \times 10^{12}}$$

Or, I = 16.7×10^{-1}

$\therefore$ I = 1.67 N/kg

$\therefore$ The gravitational field intensity of the moon on its surface is 1.67 N/kg.

2. **The mass of the Earth is 6.0×10^{24} kg and its radius is 6.4×10^{6} m. Find the gravitational field intensity at a point on its surface. What is the weight of a body of mass 100 kg on its surface? (G = 6.67×10^{-11} Nm^2/kg^2)**

Solution:

Mass of the Earth (M) = 6.0×10^{24} kg

Radius of Earth (R) = 6.4×10^{6} m

The gravitational field intensity (I) due to the earth of mass M and distance R from its centre is given by

$$I = \frac{GM}{R^2}$$

$$\text{or, } I = \frac{6.67 \times 10^{-11} \times 7.2 \times 10^{24}}{\left(6.4 \times 10^{6}\right)^2}$$

$$\text{or, } I = \frac{48.02 \times 10^{13}}{40.96 \times 10^{12}}$$

Or, I = 0.977×10

$\therefore$ I = 9.77 N/kg

Now, the weight of a body of mass 100 kg at its surface

= Mass x Gravitational field intensity

= 100×9.77

= 977 N

$\therefore$ The weight of a body of mass 100 kg on its surface is 977 N.

Summary

1. Geometrically a vector quantity is represented by a straight line with an arrow head over it.
2. The process of obtaining the resultant of sum of the vectors is called addition or composition of vectors.
3. Escape velocity is defined as the minimum speed needed for an object to escape from the gravitational field of a massive body, without the aid of thrust, or suffering the resistance from friction. The escape velocity from Earth is about 11.2 km/s on the surface.
4. Centripetal force is the force required to move a body uniformly in a circular path.
5. The outward force experienced on a body when it changes its direction of motion is called centrifugal force.
6. The center of gravity is a geometric property of any object. The center of gravity is the average location of weight of an object.
7. Gravitational field intensity is the amount of force exerted on each unit of mass of an object at a point in space caused by the presence of another object.

Exercise

A. Tick (✓) the best alternative from the following.

1. A physical quantity having both magnitude and direction is called

   (i) Scalar　　(ii) Vector

   (iii) Both (i) and (ii)　　(iv) None of the above

2. Who defined centrifugal force?

   (i) Christian Huygens　　(ii) Isaac Newton

   (ii) Charles Darwin　　(iv) Michael Faraday

3. The escape velocity from the Earth is about ……………. on the surface.

   (i) 11.2 km/s　　(ii) 11.6 km/s

   (iii) 11.4 km/s　　(iv) 11.8 km/s

4. The formula used to calculate gravitational field intensity is

(i) $I = \frac{GM}{R^2}$ (ii) $I = \frac{GM}{(R+h)^2}$

(iii) $I = \frac{GMm}{R^2}$ (iv) $I = \frac{Gmm}{R^2}$

5. The gravitational field intensity of the moon on its surface is

(i) 9.8 N/kg (ii) 1.67 N/kg

(iii) 6.67 N/kg (iv) 11.6 N/kg

B. Answer the following in brief questions.

1. What is scalar quantity?
2. Mention any four examples of scalar quantities.
3. Distance is called scalar quantity. Why?
4. What is vector quantity?
5. How is a vector quantity represented symbolically?
6. Mention any four examples of vector quantities.
7. Displacement is called vector quantity, why?
8. What is a negative vector?
9. Define escape velocity.
10. Define centrifugal force with an example. Also write its formula.
11. Define centripetal force with an example. Also mention its formula.
12. Define centre of gravity.
13. What is meant by gravitational field intensity? Write its SI unit.

C. Give long answers to the following questions.

1. Differentiate between vector and scalar quantities.
2. Mention the mechanical way to determine the centre of gravity.
3. How is centrifugal force differ from centripetal force? Give some points.
4. How can we calculate gravitational field intensity?

D. Solve the following numerical problems.

1. The mass of the planet Jupiter is 1.9×10^{27} kg and its radius is 7.1×10^{7} m. Find the gravitational field intensity at a point on its surface and calculate the weight of a body of mass 90 kg on its surface. *(Ans: 24.14 N/kg and 2172.6 N)*

2. Find the gravitational field intensity at a height of 3600 km from the surface of the earth. The radius of the earth is 6400 km and its mass is 6×10^{24} kg. *(Ans: 4.002 N/kg)*

3. What will be the gravitational field intensity of the earth if its mass could be squeezed to the size of the moon? *(Ans: 138.4 N/kg)*

Project Work

Take a stick of 1 m length and hold it between the thumb and index finger. Hold both ends of the stick in position and move both hands closer to each other. Observe the points where both the hands meet. What is this point called and how to define this point? On the basis of above activity, prepare a report and present it.

Glossary

Frictional force: force applied to overcome friction

Squeeze: make small in size

Fork: A utensil with two or more prongs, used for eating or serving food

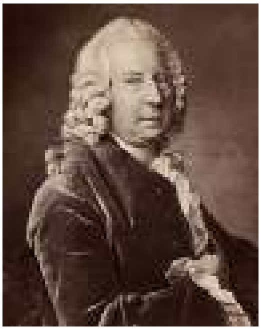
Plumb line: A line from which a weight is suspended to determine verticality or depth

Orthogonal: intersecting or lying at right angles

Unit 2

Pressure

Daniel Bernoulli (8 February, 1700 – 17, March 1782) was a Swiss mathematician and physicist. Applications of mathematics to mechanics, especially fluid mechanics, and for his pioneering work in probability and statistics are his main contributions. His name is commemorated in the Bernoulli's principle, a particular example of the conservation of energy.



Daniel Bernoulli
(1700 – 1782)

Introduction

As we have already studied about the pressure and its various aspects in previous grades. So we all are much familiar with the term 'pressure'. The word pressure is commonly generalized in various contexts such as atmospheric pressure, workload pressure, blood pressure, water pressure, etc.

In this unit, we will discuss liquid pressure and its calculation under various circumstances, surface tension, viscosity, elasticity and its limit in detail.

2.1 Fluid pressure

Fluids (liquid or gas) do not have fixed shape but they have weight and occupy space. So fluids can also exert pressure on the bottom and the walls of the container in which it is kept. The pressure exerted by a fluid per unit area is called fluid pressure. It depends upon height of fluid column, density of fluid and acceleration due to gravity.

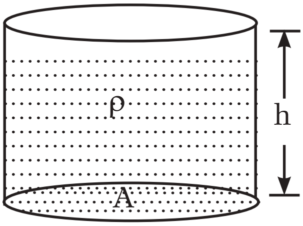


Fig. 2.1 liquid pressure

The force exerted by fluid per unit area on the base or wall of container is called fluid pressure.

Let us consider a container with a base area 'A', which is filled to a depth 'h' with a liquid of density 'ρ'.

The pressure acting on the base of the container is equal to the weight of liquid pressing down on the base.

We know,

$$P = \frac{F}{A}$$

Or, $P = \frac{m \times g}{A}$ ($\because$ force = weight of the liquid = m × g)

Or, $P = \frac{\rho \times V \times g}{A}$ ($\because$ mass = density × volume)

Or, $P = \frac{\rho \times A \times h \times g}{A}$ ($\because$ V = A × h)

$\therefore$ P= h ρ g

Therefore, pressure at a point vertically beneath the surface of a liquid P = h ρ g i.e. Pressure = depth × density of liquid × acceleration due to gravity.

Thus, $P \propto h\rho$ (Keeping 'g' constant)

$P \propto g$ (Keeping h and ρ constant)

$P \propto \rho$ (Keeping g and h constant)

$P \propto h$ (Keeping g and ρ constant)

Factors affecting the pressure at a point in a liquid

The pressure at a point in a liquid depends on the following three factors:

i. Depth of the point below the free surface (height of the liquid column)

ii. Density of liquid, and

iii. Acceleration due to gravity

The pressure at a point at a certain depth in a liquid:

i. is same in all directions.

ii. applies force at 90^0 to any contact surface, and

iii. does not depend on the shape of the container.

Do you know?

Mercury is the only metal element that exists as a liquid at room temperature. Liquids are not easily compressible at standard conditions, but may be compressible at very high pressures.

1. **If depth of water in a rectangular tank is 8 m and the water pressure exerted at the bottom is 8.0×10^4 Pa, then calculate the density of water. ($g = 10\ m/s^2$)**

Solution:

Here,

Depth of water (h) = 8 m

Acceleration due to gravity (g) = $10\ m/s^2$

Water pressure (P) = 8.0×10^4 pa

Density of water (ρ) = ?

We know that,

$P = h \rho g$

$8.0 \times 10^4 \text{ pa} = 8 \times \rho \times 10$

$= 1000\ kg/m^3$

$\therefore$ The density of water (d) is $1000\ kg/m^3$

2. **The pressure due to mercury column is 101292.8 Pa. Calculate the height of mercury column of a Barometer at sea level. (Density of mercury = $13.6\ g/cm^3$ and acceleration due to gravity = $9.8\ m/s^2$)**

Solution:

Here,

Pressure due to mercury column = 101292.8 Pa

Density of mercury (ρ) = $13.6\ g/cm^3 = 13600\ kg/m^3$

Acceleration due to gravity (g) = $9.8\ m/s^2$

Height of mercury column (h) = ?

We know that,

$p = h \rho g$

or $101292.8 = h \times 13600 \times 9.8$

or $h = 76$ cm

$\therefore$ The height of the mercury column at sea level is 76 cm.

2.1.1 Some Consequences of Liquid Pressure

a. **Deep sea divers wear a stout steel suit:** As we know that with the increase of depth of water, the pressure increases. More pressure acting directly on diver creates problem in his body. In doing so, they cannot tolerate the pressure exerted by water. If they wear special suit, then their suit can withstand the pressure of water. In other words, to withstand the high pressure which acts on them at great depths, deep sea divers wear a stout steel suit and they can save their body.

> ***Do you know?***
>
> Liquids have various applications, including as lubricants, solvents and in hydraulic systems or devices. Paint thinners are solvents. A device such as a thermometer, which measures temperature, contains mercury.

b. **Walls of a dam are made thicker at the bottom:** The pressure at a point inside a liquid depends on the depth of point from the free surface. The pressure is very high at the bottom of the dam as it is in greater depth. To withstand this pressure exerted by water, the walls of dam are made thick at the bottom. But in upper parts of the dam, less pressure is experienced by the wall. So, the wall is constructed in such a way that its thickness goes on decreasing from top to bottom gradually.

c. **Water supply tank is placed at a height:** To supply water in a town, the water supply tank is made at a greater height. Greater the height of the tank, greater will be the pressure of water in the taps of a house. Thus for a good supply of water, the height of the supply tank must always be a few meter higher than the level at which supply of water is to be made.

2.2 Surface Tension

As we know that the free surface of a liquid acts like a stretched membrane. For example, a sewing needle placed carefully on a water surface makes a small depression on the surface and floats, even though its density is higher than the density of water. Similarly, some insects can walk on the surface of water, their feet making depression on the surface but not penetrating it. In each of the above case, the liquid surface behaves like a stretched membrane and this force acting parallel to the surface, arises from the attraction between the

> ***Do you know?***
>
> The surface tension of water is very strong, due to the intermolecular hydrogen bonding, and is responsible for the formation of drops, bubbles, and meniscuses as well as the rise of water in a capillary tube, the absorption of liquids by porous substances, and the ability of liquids to wet a surface.

molecules in the liquid surface. This effect is called surface tension. In other words, it is the elastic property or tendency of a fluid by which the fluid tends to acquire the least surface area. The cohesive forces among the molecules of a fluid are responsible for it. You will get idea about it from the figure.

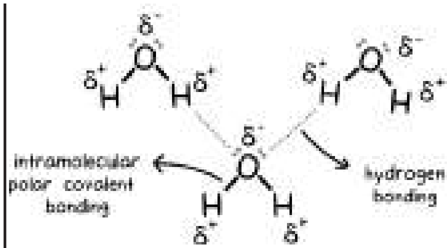


Fig. 2.2 Forces on molecules of water

Surface tension is defined as the attractive force exerted upon the surface molecules of a liquid by the molecules beneath that it tends to draw the surface molecules into the bulk of the liquid and makes the liquid assume the shape having the least surface area.

Some Examples of Surface Tension

Drops of water

While using a water dropper, the water does not flow in a continuous stream, but rather in a series of drops. The shape of the drops is caused by the surface tension of the water. The drop of water being completely spherical is because of the force of gravity acting on it. In the absence of gravity, the drop would minimize the surface area in order to minimize tension, which would result in a perfectly spherical shape.

Insects walking on water

Several insects are able to walk on water, such as the water spider. Their legs are formed to distribute their weight, causing the surface of the liquid to become depressed, minimizing the potential energy to create a balance of forces so that the spider can move across the surface of the water without breaking through the surface.

Formation of lead shots: In order to manufacture lead shots, melted lead is allowed to fall in water by spraying it from a sufficient height. During its fall, the melted lead forms small spherical drops due to surface tension forces and on entering water, they become solid.

Oil has less surface tension than water: When a drop of oil is dropped on the surface of water, due to higher surface tension of water, the oil is stretched in all directions as a thin film. Mosquitoes breed on the free surface of stagnant water. Due to the surface tension, the liquid layer supports the eggs laid by the mosquitoes. When oil is spread in water, there is little surface tension and the mosquitoes cannot breed.

Units of Surface Tension

Surface tension is measured in N/m (newton per meter), although the more common unit is the cgs unit dyn/cm (dyne per centimeter).

2.3 Viscosity

Viscosity is a scientific term that describes the resistance to flow of a fluid. The fluid can be a liquid or a gas, but the term is more commonly associated with liquids. As a simple example, syrup has a much higher viscosity than water: more force is required to move a spoon through a jar of syrup than in a jar of water because the syrup is more resistant to flowing around the spoon. This resistance is due to the friction produced by the fluid's molecules and affects both the extent to which a fluid will oppose the movement of an object through it and the pressure required to make a fluid move through a tube or pipe. Viscosity is affected by a number of factors, including the size and shape of the molecules, the interactions between them, and temperature.

Measurement of viscosity of liquid

The viscosity of a liquid can be measured by the devices called viscometers. These can either measure the time it takes for a fluid to move a particular distance through a tube or the time taken for an object with a given size and density to fall through the liquid of interest. The SI unit of measure for this is the Pascal-second, with the Pascal being the unit of pressure. This quality is therefore measured in terms of pressure and time, so that, under a given pressure, a viscous liquid will take more time to move a given distance than a less viscous one.

Factors affecting viscosity

A substance's viscosity (flow behavior) depends on the following factors:

i. The substances' inter molecular structure: the tighter the molecules are linked, the less it will be willing to flow.
ii. The outside or external forces acting upon the substance that deform it or make it flow. Both the intensity of the external force as well as the duration.
iii. The ambient conditions: the temperature and the pressure when the substance is stressed by external force. The higher the temperature is, the lower a substance's viscosity is. On the other hand, the viscosity of fluid increases with increasing pressure.

2.4 Elasticity and elastic limit

Elasticity is the ability of a body to resist a distorting influence or deforming force and to return to its original size and shape when that influence or force is removed. Solid objects will deform when adequate forces are applied on them. If the material is elastic, an object will regain to its initial shape and size when these forces are removed.

The physical reasons for elastic behavior vary for different materials. In metals, the atomic lattice changes size and shape when forces are applied (energy is added to the system). When forces are removed, the lattice goes back to the original lower energy state. For rubbers and other polymers, elasticity is caused by the stretching of polymer chains when forces are applied.

Perfect elasticity is an approximation of world. Most materials which possess elasticity in practice remain purely elastic only up to very small deformations. In engineering, the amount of elasticity of a material is determined by two types of material parameter. The first type of material parameter is called a modulus, which measures the amount of force per unit area needed to achieve a given amount of deformation. The SI unit of modulus is the pascal (Pa). A higher modulus typically indicates that the material is harder to deform. The second type of parameter measures the elastic limit, the maximum stress that can arise in a material before the onset of permanent deformation. Its SI unit is also pascal (Pa).

When describing the relative elasticity of two materials, both the modulus and the elastic limit have to be considered. Rubbers typically have a low modulus and tend to stretch a lot (that is, they have a high elastic limit) and so appear more elastic than metals (high modulus and low elastic limit) in everyday experience. Of two rubber materials with the same elastic limit, the one with a lower modulus will appear to be more elastic, which is however not correct.

2.4.1 Elasticity

The property of a body by virtue of which it tends to regain its original shape and size when the deforming force is removed is called elasticity.

Perfectly elastic body

If a body regains its original shape and size completely and instantaneously on the removal of the deforming force, then the body is called perfectly elastic. There is no body which is perfectly elastic so the concept of perfectly elastic body is only an ideal concept. Quartz is the nearest approach to the perfectly elastic body.

Perfectly plastic body

A body, which does not regain its original configuration at all on the removal of deforming force, however small deforming force may be is called perfectly plastic body. There is no such body which is perfectly plastic so the concept of perfectly plastic body is only an ideal concept. Paraffin wax, wet clay are the nearest approach to a perfectly plastic bodies.

The property of the material body by virtue of which it does not regain its original configuration when an external force acting on it is removed is called plasticity.

Stress

We know when a deforming force is applied on a body then the restoring force is developed inside the body. The restoring force per unit area of a body is called stress. The restoring force is equal and opposite to the deforming force. Therefore, stress may be defined as the deforming force per unit area of a body.

That is,

$$\text{Stress} = \frac{\text{external force or deforming force}}{\text{area of a body}}$$

$$\text{Stress} = \frac{F}{A}$$

The SI unit of stress is Nm^{-2} and in CGS system its unit is dyne cm^{-2}.

2.4.2 Elastic Limit

Elastic limit is the upper limit of deforming force up to which if deforming force is removed, the body regains its original form completely and beyond which if deforming force is increased, the body loses its property of elasticity and gets permanently deformed.

The elastic limit is the point beyond which the material you are stretching becomes permanently stretched so that the material does not return to its original length when the force is removed.

Do you know?

The elastic limit is denoted by the maximum force that can be applied per unit area before complete deformation. Applying a force greater than the elastic limit of the material would cause the material to bend permanently or to crack.

Hooke's law

The law of elasticity was discovered by the English scientist Robert Hooke in 1660. It states that, for relatively small deformations of an object, the displacement or size of the deformation is directly proportional to the deforming force or load. Under these conditions, the object returns to its original shape and size upon removal of the load.

Hooke's law, F = kx, where the applied force F is equal to constant.

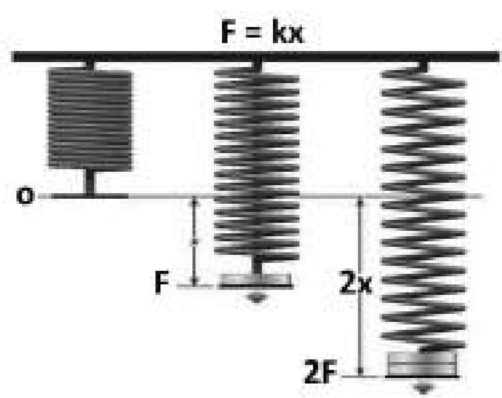


Fig. 2.3 Hooke's law

The deforming force may be applied to a solid by stretching, compressing, squeezing, bending, or twisting. Thus, a metal wire exhibits elastic behaviour according to Hooke's law because the small increase in its length when stretched by an applied force doubles each time the force is doubled.

Mathematically, Hooke's law states that the applied force F equals a constant k times the displacement or change in length x, i.e $F = kx$. The value of k depends not only on the kind of elastic material under consideration but also on its dimensions and shape.

At relatively large values of applied force, the deformation of the elastic material is often larger than expected on the basis of Hooke's law, even though the material remains elastic and returns to its original shape and size after the removal of force. Hooke's law describes the elastic properties of materials only in the range in which the force and displacement are proportional. Sometimes Hooke's law is formulated as $F = -kx$. In this expression F no longer means the applied force but rather means the equal and oppositely directed restoring force that causes elastic materials to return to their original dimensions.

Summary

1. The pressure exerted by a fluid per unit area is called fluid pressure.
2. Liquid pressure depends upon height of liquid column, density of liquid and acceleration due to gravity.
3. Factors affecting the pressure at a point in a liquid are

   i. Depth of the point below the free surface (h)

   ii. Density of liquid (ρ)

   iii. Acceleration due to gravity (g)
4. The property of a liquid at rest by virtue of which its surface behaves like a stretched membrane and tries to occupy minimum possible surface area is called surface tension.
5. The viscosity of a liquid can be measured by the devices called viscometers.
6. A fluid with a low viscosity is said to be "thin," while a high viscosity fluid is said to be "thick.
7. Elasticity is the ability of a body to resist a distorting influence or deforming force and to return to its original size and shape when that influence or force is removed.

8. Solid objects will deform when adequate forces are applied on them. If the material is elastic, an object will return to its initial shape and size when these forces are removed.
9. The property of a body by virtue of which it tends to regain its original shape and size when the deforming force is removed, then the body is called elasticity.
10. The restoring force per unit area of a body is called stress.
11. Hooke's law states that for relatively small deformations of an object, the displacement or size of the deformation is directly proportional to the deforming force or load.

Exercise

A. **Tick (√) the best alternative from the following.**

1. Liquid pressure depends upon
   i. acceleration due to gravity
   ii. height of liquid column and density of liquid
   iii. height of liquid column and acceleration due to gravity
   iv. height of liquid column, density of liquid and acceleration due to gravity
2. If depth of water in a rectangular tank is 2 m, then what will be the pressure exerted by water at the bottom?
   i. 20 pa ii. 200 pa
   iii. 2000 pa iv. 20000 pa
3. What is the name of device which is used to measure the viscosity of a liquid?
   i. viscometer ii. manometer
   iii. altimeter iv. barometer
4. What will happen, if the material is elastic?
   i. The object will return to its initial shape when the applied forces are removed.

ii. The object will not return to its initial shape and size when the applied forces are removed.

iii. The object will return to its initial shape but not in size when the applied forces are removed.

iv. The object will return to its initial shape and size when the applied forces are removed.

5. The restoring force per unit area of a body is called

i. Stress | ii. Elasticity

iii. Surface tension | iv. Viscosity

B. Answer the following questions in short.

1. What is meant by liquid pressure?
2. On what factors does liquid pressure depend?
3. Define surface tension with an example.
4. What is meant by viscosity?
5. Out of two fluids; water and honey, which one has more viscosity?
6. Name a device which is used to measure the viscosity of fluid?
7. What is meant by elasticity?
8. Define stress.
9. State Hooke's law.

C. Give long asnwer to the following questions.

1. Derive the relation $P = h \rho g$.
2. Mention some effects of liquid pressure.
3. Define surface tension with two examples.
4. Explain viscosity of a fluid.
5. List various factors that affect viscosity.
6. Explain different types of stress with examples.
7. Distinguish between perfectly elastic body and perfectly plastic body.

D. Numerical Problems

1. The depth of water in the Thada Lake of Arghakhanchi district is 8 m, find the pressure exerted by water at the bottom of it. *(Ans: 8.0×10^4 Pa)*

2. If the pressure exerted by water at the bottom of the Taudaha Lake is 9.8×10^4 pa, calculate the depth of this lake. *(Ans: 10 m)*

3. Calculate the pressure exerted by a mercury column of height 750 mm at its bottom. Given that the density of mercury is 13.6 g/cm^3 and $g = 9.8\ m/s^2$ *(Ans: 9.99×10^4 Pa)*

Project Work

Take a vessel having three holes A, B and C each with a stopcock. Water is poured into it. What happens when the stopcocks are opened simultaneously? From which taps the water comes out with greater force and why? On the basis of this information, prepare a report and present it to the class.

Glossary

Distorting : to change something from its natural or usual shape or condition

Restoring force : a force that gives rise to an equilibrium in a physical system

A fluid : a liquid or a gas

Polymers : a chemical compound or mixture of compounds formed by polymerization

Unit 3

Energy

Bernard Forest de Belidor, (1698 - 1761), is a military, civil engineer and author of a classic work on hydraulics. After serving in the French army at an early age, he developed an interest in science and worked on the measurement of an arc of the earth. He wrote several notable books on engineering, artillery, ballistics, and fortifications, but his fame rests primarily on Architecture hydraulique, in four volumes (1737–53), covering engineering mechanics, mills and waterwheels, pumps, harbours, and sea works.

Bernard Forest (1698 – 1761)

Introduction

We need energy to carry out different works in our daily life. Without energy, we cannot do anything. To operate various types of machines also, energy is needed. Some need more energy, whereas some others need less energy. The ability of a body to do various works is called energy.

We, human beings, obtain energy from food. Food enables us not only to breathe and think but also allows us to perform verious activities such as running, carrying things, playing a game, etc. Just as a vehicle requires fuel to run so do our body also requires fuel which is supplied in the form of food. Fuel supplies energy to the vehicle and makes it run. Food supplies energy to us which enables us to perform various activities.

Energy is contained in coal, wind, petrol, kerosene, the sun and other many more. There are two types of sources of energy. They are renewable and non renewable sources.

In this unit we will discuss about hydroelectricity and its production technology and fossil fuels, biogas and alternative sources of energy.

3.1 Hydroelectricity

As we know, water in rest position possesses potential energy, but the water in motion possesses kinetic energy. The water in motion is called flowing water. In fact, flowing water is the major source of energy. The electricity produced by using flowing water is known as hydroelectricity. A plant used to generate hydro electric power is known as hydro electric power plant.

Principle of generation of hydroelectricity

Potential energy of water stored in a dam is converted into kinetic energy when it runs through a pipe. The water falls on the turbine, so kinetic energy of the flowing water is used to rotate electromagnet and the armature of the generator connected to the turbine. Then kinetic energy is converted into the electrical energy known as hydroelectricity.

Technology of generation of hydroelectricity

A dam or reservoir tank is made over a river. The energy of stored water in the dam is potential energy. The water in a dam is allowed to fall on the water wheel or turbine. As a result of this, the turbine rotates whose axle is connected with an armature of the generator. The armature of generator remains fixed but the electromagnet rotates. The rotation of the electromagnet gives rise to electric current or electricity. This electricity is transmitted to the sub-stations for further distribution to the houses and factories.

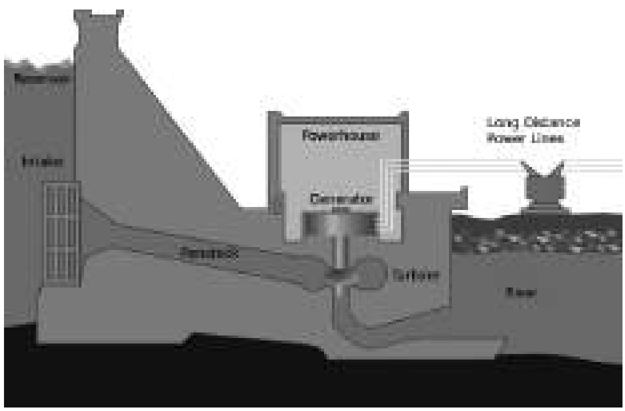


Fig. 3.1 Generation of hydroelectricity

The largest source of renewable energy in the world is hydroelectricity. It is claimed that Nepal is the second richest country in the world in hydroelectric power after Brazil. It is estimated that the total production capacity of hydroelectric power of Nepal is about 83 thousand megawatt (MW). In fact, hydropower is an indirect form of solar energy because it is driven by solar energy.

Advantages of generating hydroelectricity

a. It causes less environmental pollution, i.e. it is environment friendly.

b. It can be transmitted through wire to a long distance and be converted into different forms.

c. It is natural and renewable source of energy.

d. It is cheaper in a long run.

Disadvantages of generating hydroelectricity

a. A large number of people residing near the site of a dam are dislocated. So, many problems are to be faced by the concerned authority in their rehabilitating.

b. A large area of fertile land gets submerged at the site of the reservoir.

c. Hydroelectric power is generated only near the rivers having water throughout the year. This electric power has to be carried to the sub-stations for distribution and then to the consumers' houses and factories situated far off from the sites of hydro electric power stations. It is difficult and expensive.

d. It causes disturbance in the ecosystems when land is submerged under the water reservoir.

e. Energy level in hydropower plant varies from time to time i.e. it is not constant.

Application of hydroelectricity

a. It can be used as a multipurpose energy source such as cooling, heating, lighting, etc.

b. It is used in industries and factories to operate various electrical machines.

c. It is used in telecommunication to power electronic devices.

d. It is used in vehicles and automobiles.

3.2 Bio fuels

Biofuels are fuels produced directly or indirectly from organic materials like biomass (plant materials and animal wastes). Biofuels can be solid, gaseous or liquid. There are two types of bio fuels. They are primary and secondary biofuels. Primary biofuels, such as fuelwood, wood chips and pellets, organic materials are used in an unprocessed

Do you know?

Pharping Hydro Power, the first hydropower project, was established in 1911 as Chandrajyoti Hydro-electric power station by Prime Minister Chandra Shamsher Jang Bahadur Rana. Plant was Inaugurated by King Prithvi Bir Bikram Shah Dev on Monday, 22 May, 1911 by turning the lights on during a program in Tudhikhel, Kathmandu.

form, primarily for heating, cooking or electricity production. Secondary biofuels result from processing of biomass and include liquid biofuels such as ethanol and biodiesel that can be used in vehicles and industrial purposes.

Fossil fuels

Fossil fuels are fuels formed by natural processes such as anaerobic decomposition of buried dead organisms. Fossil fuels contain high percentages of carbon and include petroleum, coal, and natural gas. Other commonly used derivatives include kerosene and propane.

Briquette and its type

A briquette is a block of flammable matter, which is used as a fuel to start and maintain a fire. A briquette (or briquet) is a compressed block of coal dust or other combustible biomass material such as charcoal, sawdust, wood chips, peat, or paper used for fuel and kindling to start a fire. The term comes from the French language and is related to brick.

There are different types of briquettes. They are categorized as coal briquette, charcoal briquettes, Japanese briquettes, peat briquettes, biomass briquettes and paper briquettes. Among all, the biomass briquette is most popular in the context of Nepal. Here, we are going to describe about the biomass briquettes

Biomass briquettes

Biomass briquettes are made from agricultural waste and are a replacement for fossil fuels such as mineral oil or coal, and can be used to heat boilers and also have wide applications in developing countries. Biomass briquettes are a renewable source of energy and avoid adding fossil carbon to the atmosphere.

Fig. 3.2

Development of Biomass briquetting technology in Nepal

Biomass briquetting was introduced in Nepal in the year 1986 through a demonstration program organised by a Japanese private company with the support of Japanese Embassy. The technology used for the demonstration, was based on the extruder principle and manufactured by Fuji Conveyor. This program fostered a growth in the briquette manufacturing industry. In 1987/88, four extrusion type briquetting machines were imported from Sun Chain Company, Taiwan and established in Simara, Hetauda, Chitwan and Parwanipur. However, by early 1990s most of the briquetting industries, except one, closed down due to various reasons. Presently, only one plant, Mhaypi Briquette Industry Private Limited in Nawalparasi is in operation.

The extrusion type plant basically consists of the following components:

1. The electric motor, which drives the pulley for rotating the screw
2. The hopper, for feeding the raw material
3. The die heater and muff
4. The screw, which densifies the raw material to produce briquettes.

The electric motor drives the briquetting screw, which is housed inside the die, through a V- belt and pulley arrangement. Biomass raw material is fed to the screw through the hopper. The electric die-heater softens the lignin in the raw material as it passes through the die, which acts as a binding material. A smoke trapping system traps and removes the smoke from the vicinity during the briquetting process. The produced briquettes are collected at the opening provided at the bottom of the smoke collection box.

Advantages of Biomass briquettes

a. Biomass briquettes are renewable energy, which are made from industrial and agricultural wastes used for heating and cooking in developing countries.
b. Briquettes are easy to handle, transport and store because they have uniform size and quality.
c. Briquettes help to solve the residual disposal problem as they are made from industrial waste/ by product.
d. Briquettes minimize air pollution as well as cost because of high burning efficiency and complete burning.

Biogas Plant

Biogas typically refers to a mixture of different gases like methane (CH4), carbon dioxide (CO2) and may have small amounts of hydrogen sulphide (H2S), moisture, etc. produced by the breakdown of organic matter in the absence of oxygen. Biogas can be produced from raw materials such as agricultural waste, manure, municipal waste, plant material, sewage, green waste or food waste. Biogas is a renewable energy source and in many cases exerts a very small carbon footprint. **The arrangement of producing biogas from animals' dung, human excreta, industrial and domestic wastes is known as biogas plant**. In other words, the physical installation to produce biogas from biomass is called biogas plant.

Do you know?

Biogas was first introduced to Nepal on an experimental basis in 1955. The initial experiences showed the feasibility of this technology for meeting a significant portion of rural household energy needs.

Structure of Biogas Plant

A biogas plant consists of a well shaped, underground tank called digester, which is made of bricks, and has a dome shaped roof, also made of cement and bricks as shown in figure below. The digester is a kind of sealed tank in which there is no air. The dome of the digester tank acts as a gas-holder or storage tank for the biogas. There is a gas out-let at the top of the dome having a valve. On the left side of the digester tank is a sloping inlet chamber and on the right side is a rectangular outlet chamber, both are made up of bricks. The inlet chamber is for introducing fresh dung slurry into the main digester tank whereas the outlet chamber is for taking out the spent dung slurry after the extraction of biogas. The inlet chamber is connected to a mixing tank while the outlet chamber is connected to overflow tank.

Biogas Production

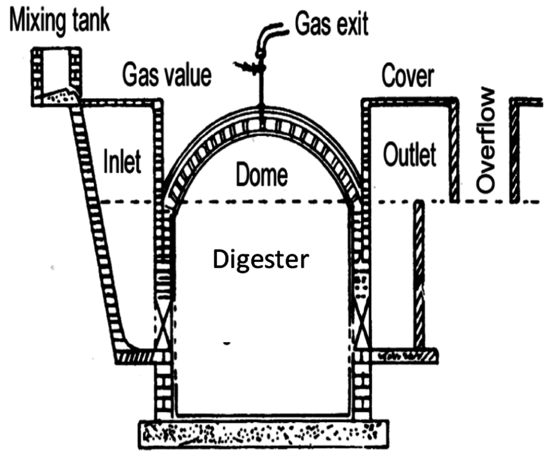


Fig. 3.3. Domestic biogas plant

Here, we will deal with the working of Biogas Plant in detail. First of all, cow-dung and water are mixed in equal proportions in the mixing tank to prepare a slurry. This slurry of dung and water is fed into the digester tank through the inlet chamber. The digester tank is filled with dung slurry up to the cylindrical level as shown in figure 3.2 and the dome being left free for the collection of biogas. It takes about 50 to 60 day for the new gas-plant to start functioning. During this period, the cow dung undergoes degradation by anaerobic bacteria in the presence of water (but in absence of oxygen) with the gradual evolution of biogas. This biogas starts collecting in the dome. As more and more biogas collects in the dome, it exerts pressure on the slurry in the digester tank, and forces the spent slurry to go into overflow tank through the outlet chamber. From the overflow tank, the spent slurry is removed gradually. The spent dung slurry, left after the extraction of biogas, is rich in nitrogen and phosphorus compounds and hence forms good manure.

Advantages of using biogas

a. Biogas is a good source of energy as it is used for cooking food and heating purposes.

b. Biogas is environment friendly i.e. it does not cause any environmental pollution.

c. It is a cheaper source of energy because the raw materials needed for the production of biogas are not costly and are locally available.

d. Nitrogen and phosphorous rich manure is produced after the decomposition of biomass, which is useful to yield crops.

3.3 Alternative sources of energy in Nepal

Many researches related to energy have concluded that the major portion of the world's energy demand is being fulfilled by fossil fuels such as petrol, diesel, natural gases, etc. The present rate of consumption of these fuels is so high. It is said that the present consumption rate of fossil fuel will bring energy crisis in near future. This is because of being these are non renewable sources of energy. To avoid such situation, we need to search alternative sources of energy. Then we can push back energy crisis. The sources of energy, which are available in nature, can be used by technology of recent time and used instead of nonrenewable sources of energy are called alternative sources of energy.

Do you know?

Alternative Energy Promotion Centre (AEPC) is a Government institution established on November 3, 1996 under the Ministry of Science and Technology with the objective of developing and promoting renewable/alternative energy technologies in Nepal. Currently, it is under Ministry of Population and Environment.

The renewable sources of energy, which are used instead of non renewable sources of energy, are called alternative sources of energy

Biomass Energy

The waste materials of plants and animals are called biomass. Dry wood, dry weeds, agricultural wastes and animals dung, etc. are some examples of biomass. They can be burnt directly to get heat energy. The energy obtained by the burning of biomass is called biomass energy. Biomass is a renewable and important source of energy in undeveloped or developing countries. Specially, people use wood, dung, husk, and remaining part of sugarcane after extraction of juice, straw, grass, etc. for domestic purposes. In the countryside, people dry cattle dung in the sun light called 'Guintho' and they use them as the source of energy. Such sources of energy are easily available and cheaper in use. But their huge amount can't be obtained at once and cause air pollution.

Wind Energy

The energy possessed by wind is called wind energy. Wind energy is the largest and the most viable form of renewable source of energy. The kinetic energy of moving air has been traditionally used to pump water and to run flour mills and to propel sailing boat. Improved versions of the wind mill are used to generate electricity from wind energy at present.

Do you know?

The wind-solar hybrid system was installed in Dhaubadi village of Nawalparasi district in December 2011 under ADB's regional technical assistance (RETA) for Effective Development of Distributed Small Wind Power Systems in Asian Rural Areas for which the Alternative Energy Promotion Centre (AEPC) was the implementing agency in Nepal.

Solar Energy

The energy emitted by the sun is termed as solar energy. In fact, solar energy sustains all forms of life on the earth. Solar energy has the greatest potential of all the sources of energy. It is understood that the energy given by the sun to the earth in one day is 50,000 times more than the total energy consumed by the world in one year. Almost all sources of energy on the earth originate from sun. Some of this energy gets stored in forms which we can use. Sun is an enormous source of energy. This energy is due to the nuclear fusion reaction taking place in the core of the sun. This energy is radiated by it in all directions in space. The earth and other planets receive only a small fraction of this energy. Solar energy is directly used to produce heat and also used to produce electricity. Those devices which convert solar energy into other forms of energy are called solar devices. e.g. solar cooker, solar heater, solar cells, solar power plants, etc. Solar energy is very useful to our country since in most parts of our country there are about 300 sunny days in a year.

Summary

1. The ability of a body to do various works is called energy.
2. The electricity produced by the flowing water is called hydroelectricity. A plant used to produce hydro electric power is called hydro electric power plant.
3. Fossil fuels are fuels formed by natural processes such as anaerobic decomposition of buried dead organisms, containing energy originated in ancient photosynthesis.
4. A briquette is a block of flammable matter, which is used as a fuel to start and maintain a fire.

5. Different types of briquettes are coal briquettes, charcoal briquettes, Japanese briquettes, peat briquettes, biomass briquettes and paper briquettes.
6. Biomass briquettes are made up of agricultural waste and are a replacement for fossil fuels such as oil or coal, and can be used to heat boilers in manufacturing plants.
7. Biogas refers to a mixture of different gases produced by the breakdown of organic matter in the absence of oxygen. Biogas can be produced from raw materials such as agricultural waste, manure, municipal waste, plant material, sewage, green waste or food waste.
8. The arrangement of producing biogas from animals' dung, human excreta, industrial and domestic wastes is called biogas plant.
9. The renewable sources of energy, which are used instead of non renewable sources of energy, are called alternative sources of energy.
10. The energy possessed by wind is called wind energy. Wind energy is the largest and the most viable form of renewable energy.
11. The energy emitted by the sun is called solar energy

Exercise

A. Tick (√) the best alternative from the following.

1. The existing hydroelectric power production in Nepal till 2074 B.S. is

   i. 932 MW　　ii. 1032 MW

   iii. 1132 MW　　iv. 1232 MW

2. How does energy transformation take place during the production of hydroelectricity?

   i. Potential → Kinetic → Electricity

   ii. Kinetic → Mechanical → Electricity

   iii. Potential → Kinetic → Mechanical → Electricity

   iv. Mechanical → Potential → Electricity

3. Which one of the following briquettes is most popular in Nepal?

   i. coal briquettes　　ii. peat briquettes

   iii. biomass briquettes　　iv. paper briquettes

4. Biogas primarily consists of

   i. methane (CH_4)

   ii. methane (CH_4) and carbon dioxide (CO_2)

iii. small amounts of hydrogen sulphide (H_2S)

iv. All of the above

B. Give short answers to the following questions:

1. What is hydroelectricity?
2. Mention any two advantages and disadvantages of hydroelectricity.
3. Write any three applications of hydroelectricity.
4. What is the technology used to produce hydroelectricity?
5. Define bio fuel.
6. What is meant by briquette?
7. What is biogas?
8. Mention any two advantages of bio gas plant.
9. What is meant by alternative source of energy?
10. Mention any two examples of alternative sources of energy.
11. Define nuclear energy.
12. What is meant by wind energy?
13. Define biomass energy.

C. Give long answers to the following questions:

1. How is hydroelectricity produced? Explain.
2. How is biogas produced? Explain.
3. Draw a well labeled diagram of domestic bio gas plant.
4. Explain the briquetting technology of biomass.
5. How is wind energy obtained? Explain.
6. What are the environmental implications of the following energy sources?

   i. fossil fuels ii. solar fuels
7. Write an essay on energy consumption scenario of our country.

Project Work

1. Visit a nearby biogas plant. Collect the information from the expert regarding the technology of biogas plant and prepare a report and present it to the class.
2. What are the sources of energy used at your home? Collect a data related to monthly expenditure electricity, kerosene, coal, wood and petrol. Make a bar graph with the help of this collected information and present a report to the class.

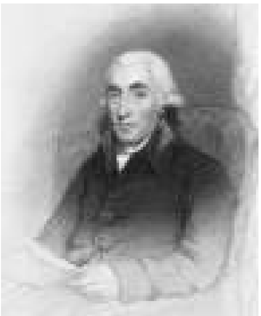
Key words

Kindling :	easily combustible material for starting a fire
Agro-residues :	all the organic materials produced from processing of agricultural crops
Extrusion :	process used to create objects of a fixed diameter
Concentric hole :	two or more objects of same center or axis
Yell :	to cry out loudly

Unit 4

Heat

Joseph Black (born April 6, 1728, Bordeaux, France - died Nov 10, 1799 Edinburgh, Scot.), British chemist and physist best known for the rediscovery of fixed air (carbondioxide), the concept of latent heat, and the discovery of the bicarbonates (such as bicarbonate of soda). Black lived and worked within the context of the Scottish Enlightment, a remarkable flourishing of intellectual life in Edinburgh, Glasgow, and Aberdeen during the latter half of the 18th century.



Joseph Black (1728– 1799)

Introduction

Matters exist in three physical states, i.e solid, liquid and gas. The physical states of a substance can be changed either by heating or cooling. To clarify this concept we can take an example of water. Water also exists in three physical states as mentioned above. Ice is the solid form of water, water itself is in liquid form and vapor is the gaseous form of water. When ice is heated, it melts and forms water. The process in which a solid changes into a liquid on heating is called melting or fusion. On further heating of water, changes into steam (vapour). The process in which a liquid changes into vapour is called vaporization. In the same way, on cooling vapour, changes into water and further cooling it, changes into ice. The process in which vapour changes into liquid is called condensation and water into ice is called solidification. So we can say that the physical states of a substance can be changed due to heat.

In this unit, we will discuss the latent heat, heat equation and its numerical problem and calorimetry and its calculation in detail.

Latent Heat

Latent heat is an energy absorbed or released by a substance during the change in its physical state (phase) that occurs without changing its temperature. The latent heat associated with melting a solid or freezing a liquid is called the heat of fusion; that associated with vaporizing a liquid or a solid or condensing a vapour is called the heat of vaporization. The latent heat is normally expressed as the amount of heat (in units of joules or calories) per mole or unit mass of the substance undergoing a change of state.

The latent heat of a substance is defined as the amount of heat absorbed by a unit mass of a substance to change its state without change of temperature.

For example, ice at 0 °C is melted into water at 0°C. Here, state is changed from solid (ice) to liquid (water) and vice versa without change in temperature 0°C. Hence, in latent heat, heat energy is utilized not for rise in temperature but it is used in the change of internal energy (molecular separation) i.e. for state changed. Similarly, other example is water at 100°C is converted into steam at 100°C. The heat of fusion for water at 0 °C is approximately 334 joules per gram, and the heat of vaporization at 100 °C is about 2,230 joules per gram. Because the heat of vaporization is so large, steam carries a great deal of thermal energy that is released when it condenses, making water an excellent working fluid for heat engines.

The S. I. unit of heat is joule and that of the mass is kilogram. So, the S.I. unit of latent heat is joules per kilogram (J/kg). It is generally represented by the letter L.

There are large number of molecules in a substance. There can be found the force of attraction among these molecules present in the substances. To increase the temperature of the substance, there should increase in the kinetic energy of the molecules. The latent heat is used up in overcoming the force of attraction among the molecules of the substance instead of rise in temperature. Latent heat can be categorized as Latent heat of fusion and Latent heat of vaporization. In this topic we will deal with the study of these types of latent heat in detail.

Do you know?

When water evaporates it takes up heat. As water vapor it carries that heat around as latent heat. Then when that vapor condenses it releases that latent heat, heating up the local environment, usually the air. This is what drives some types of storms, including thunderstorms, tornadoes, hurricanes and typhoons.

Latent heat of fusion

Latent heat of fusion is called latent heat of melting. It is also called solid to liquid change. To draw clear ideas about the latent heat, we can do the following activity.

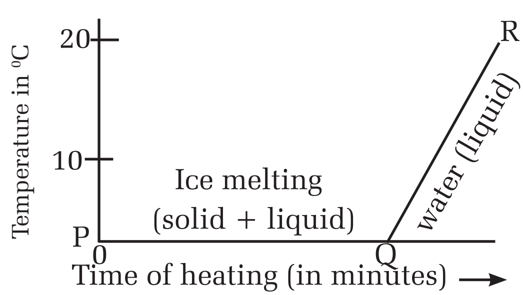


Fig. 4.1 Latent heat of fusion

Take some crushed ice in a beaker and note its temperature with the help of a laboratory thermometer, it is found to be 0^0C. Now heat these pieces of ice by using a small flame and again note the temperature after every minute. On heating of ice continuously, it starts melting to form water in the same pace but thermometer shows ^{0}C temperature. It means that there is no rise in temperature even if heat is being supplied continuously. The temperature in the thermometer remains ^{0}C until there remain some pieces of ice. When all the pieces of ice have melted to form water, on further heating, the temperature of water starts increasing. It can be further illustrated with the help of time - temperature graph as shown in the figure above.

In this graph, at point P, there are all crushed ices. When heat is supplied to this ice, it starts melting and changes into water. But the temperature of the ice and water remains zero. The line PQ in the graph as given shows a constant temperature of 0^0C during its melting. At point Q, all the pieces of ice have melted to form water. It means that there are no ice pieces at point Q. On further heating of the water beyond Q, the temperature of the water starts rising as shown in the figure alongside.

From this activity, we came to conclude that when ice melts and changes into water, the temperature remains constant at 0^0C though heat is being supplied continuously to ice. To convert the ice into water, some amount of heat energy is required. The essential energy is maintained by the heat given from outside. Since the heat absorbed during the change of state of a substance does not raise its temperature, which is called latent heat. It is experimentally found that 3.34×10^5 joules of heat energy is required to convert 1 kg of ice at its melting point to water at the same temperature. This is known as the latent heat of fusion of ice.

The latent heat of the fusion of a substance is the amount of heat required to convert a unit mass of the substance from the solid state to the liquid state without change of temperature.

If the S.I. unit of heat is joule and mass is kg, then the latent heat of fusion of a solid can be defined as the amount of heat in joules required to convert 1 kg of a solid to liquid, without any change in temperature. For example, ice melts at 0^0C means the latent heat of fusion of ice is the heat required to change 1 kg of ice at 0^0C to water at 0^0C. This is equal to 3.34×10^5 joules of heat energy. The latent heat of fusion of some of the substances is given in the table below:

Substance	Latent heat of fusion in S.I.units	Melting points of the substance
Ice	3.34 x 10^5 joules J/kg	0^0C
Lead	0.25 x 10^5 joules J/kg	327^0C
Silver	0.92 x 10^5 joules J/kg	961^0C
Zinc	1.13 x 10^5 joules J/kg	420^0C
Copper	1.80 x 10^5 joules J/kg	1083^0C
Aluminium	3.21 x 10^5 joules J/kg	658^0C

Latent heat of vaporization

Latent heat of vaporization is also called liquid to vapour change. As we have studied in earlier topic that heat is required to change a solid (ice) substance into a liquid (water). Similarly, heat is required to change a liquid (water) into a gas (vapour). The heat required to change a liquid into the vapour state is called latent heat of vaporization. Let us suppose there is some water in a beaker at 20^0C. Heat this water by using a burner and note its temperature after every minute with the help of a thermometer. In doing so, when heat is given, the temperature of water rises steadily from 20^0C to 100^0C. At 100°C, water boils and starts changing into water vapour or steam. On supplying more heat on it, more steam is formed but the temperature remains constant in it, i.e. 100^0C until all the water has changed into steam. When all the water has vaporized into steam, on further heating, the temperature of steam starts rising. It can be further illustrated with the help of a "time-temperature" graph as shown in the diagram above.

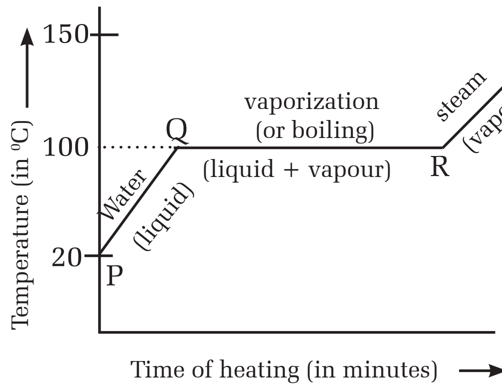


Fig. 4.2 Latent heat of vaporiation

In this graph, at point P, there is water at 20^0C. When heat is supplied to this water, it starts rising as shown by the sloping line PQ in the graph. Water starts boiling at 100^0C, which can be shown by point Q. The line PQ in the graph as given shows a constant temperature of 100^0C during its complete vaporization. On further heating of the water beyond Q, more water is being converted into steam. At point R

> ***Do you know?***
>
> A substance going directly from the solid to the gas state, or the reverse, the heat absorbed or given up is known as the latent heat of sublimation.

all the water has been converted into steam. On further heating of steam beyond R, the temperature of the steam starts rising as shown in the figure.

From this activity, we came to conclude that when water boils and changes into steam, the temperature remains constant at 100^0C though heat is being supplied continuously to water. To convert the water into steam, some amount of heat energy is required. The essential heat energy is maintained by the heat given from outside. Since the heat absorbed during the change of state of a substance does not rise its temperature, which is called latent heat. It is experimentally found that 22.5×10^5 joules of heat energy is required to convert 1 kg of water at its boiling point to steam at the same temperature. This is known as the latent heat of vaporization of water.

The latent heat of the vaporization of a substance can be defined as the amount of heat required to change a unit mass of the substance from the liquid state to vapour state without change of temperature.

If the S.I. unit of heat is joule and mass is kg, then the latent heat of vaporization of a liquid can be defined as the amount of heat in joules required to convert 1 kg of the liquid to steam, without any change in temperature. For example, water boils at 100^0C means the latent heat of vaporization of water is the heat required to change 1 kg of water at 100^0C to steam at 100^0C. This is equal to 22.5×10^5 joules of heat energy. The latent heat of vaporization of some of the substances is given in the table below:

Substance	Latent heat of vaporization in S.I.units	Boiling points of the substances
Water	22.5×10^5 joules J/kg	100^0C
Alcohol	8.5×10^5 joules J/kg	78.5^0C
Ether	3.9×10^5 joules J/kg	34.5^0C
Mercury	2.8×10^5 joules J/kg	357^0C
Sulphuric acid	5.1×10^5 joules J/kg	338^0C

Calculation of heat absorbed or given out during vaporization and heat given out during condensation

It is experimentally found that the quantity of heat absorbed or given out by a substance during its change of state is directly proportional to the mass of the substance.

i.e. $Q \propto m$

So, $Q = L \times m$

Where, L is a constant called "latent heat" of the substance and its value depends only on the nature of the substance. It can be written as:

$Q = m \times L$

Where, Q = Heat absorbed during the change of state

m = mass of the substance

L = Latent heat of the substance

∴ **Amount of heat absorbed or given out = Mass of the substance x Latent heat of substance during the change of state of a substance**

(Note: The above formula $Q = m \times L$ can only be applied when there is a change of state of a substance.)

Numerical Illustration

Calculate the amount of heat required to convert 200 g of ice into water without change of temperature. (Latent heat of ice = 3.34×10^5 J/kg)

Solution:

Mass of ice (m) = 200g

= 200/1000

= 0.2 kg

Latent heat of ice (L) = 3.34×10^5 J/kg

We know that,

$Q = m \times L$

Or, $Q = 0.2 \times 3.34 \times 10^5$

$= 6.68 \times 10^4$ joules

∴ The heat required is 6.68×10^4 joules.

Heat Equation

When heat is supplied to a cold body, it absorbs heat and its temperature rises; but when a hot body is cooled, it gives out heat and its temperature falls. Let the temperature of a body of mass 'm' rises its temperatures from t_1 to t_2, (change in temperature be Δt). when it is heated by supplying 'Q' amount of heat. So,

Q = quantity of the heat required

m = mass of a body

S = specific heat capacity

t_2, = final temperature

t_1 = initial temperture

Δt= change in temperature

It is experimentally found that the heat absorbed or given out by a body (Q) is directly proportional to the mass of the body (m) and change in temperature of the body (Δt). i.e.

$Q \propto m$ ....(i)

$Q \propto \Delta t$..............................(ii)

Combining equation (i) and (ii), we have,

$Q \propto m\Delta t$

$Q = s\, m\Delta t$

$Q = ms\Delta t$ ....(iii)

Where, s is a proportionality constant called specific heat capacity of the body and its value depends upon the nature of the material of the body. The mathematical relationship among these Q, m, s and Δt as derived in equation (iii) is called heat equation and it can be written in words in the following ways:

Heat equation can be defined as the product of mass, specific heat capacity and change in temperature is equal to the heat gained or heat given out by the body.

Mathematically,

Heat absorbed or given out = mass x specific heat capacity x change in temperature

Numerical Illustration

1. How much heat must be added to raise the temperature of 200 g of water from 15^0C to 85^0C?

   Solution:

   Mass of water (m) = 200g

   = 200/1000kg

   = 0.2 kg

   Specific heat of water (s) = 4.2×10^3 J/kg0C

   Initial temperature of water (t_1) = 15^0C

   Final temperature of water(t_2) = 85^0C

   So, Change (or rise) in temperature (Δt) = $(85 - 15)°C$

   = $70^0 C$

We know that,

$$Q = m \times s \times \Delta t$$

$$Q = 0.2 \times 4.2 \times 10^3 \times 70$$

$$= 58800 \text{ joules}$$

$\therefore$ The heat required is 58800 joules.

2. An electrical kettle contains 200gm of water at 20^0C. How much heat is to be supplied in order to boil the water? (Assume that the water boils at 100^0C and the kettle takes no heat)

Solution:

Mass of water, m = 200 g = 0.2 kg

Specific heat of water (s) = 4200 J/kg^0C

Initial temperature of water (t_1) = 20^0C

Final temperature of water (t_2) = 100^0C

Heat supplied to the water (Q) = ?

We have,

$$Q = m \times s \times \Delta t$$

Or, $Q = 0.2 \times 4200 \times (100-20)$

Or, $Q = 0.2 \times 4200 \times 80$

$\therefore$ $Q = 67200$ joules

$\therefore$ The amount of heat to be supplied is 67200 joules.

Calorimetry

Calorimetry is defined as the measurement of heat. So, the principle of calorimetry means the principle of measurement of heat. According to the principle of calorimetry, when two bodies of different temperatures are kept in thermal contact, the heat flows from a body at a higher temperature to a body at the lower temperature until each of them has equal temperature. According to the principle of conservation of energy, energy can neither be created nor be destroyed, but its forms only can be changed. The heat lost by a body with a higher temperature is equal to the heat gained by the body with the lower temperature, keeping other factors constant.

Heat lost by hot body = Heat gained by cold body

i.e. Heat lost = Heat gained

Heat lost = Heat gained is only valid where there is no change of states of material of a body.

Both of them is calculated by using the same formula i.e. $Q = ms\Delta t$,

Where,

Q = quantity of the heat required

m = mass of a body

S = specific heat capacity

Δt= Change in temperature

To be clearer about calorimetry, let's us discuss with the help of an example:

Let,

Mass of a hot body = m_1

Specific heat capacity of the body = s_1

Fall in temperature of the hot body = t_1

Mass of a cold body = m_2

Specific heat capacity of the cold body = s_2

Rise in temperature of the cold body = t_2

According to the principle of calorimetry,

Heat lost = Heat gained

Or, Q1 = Q2

Or, $m_1 \times s_1 \times t_1 = m_2 \times s_2 \times t_2$

$\therefore m_1 \times s_1 \times t_1 = m_2 \times s_2 \times t_2$

Numerical Illustration

A brass rod of 0.4 kg mass at 100⁰C is dropped into 1.0 kg of water at 20⁰C. The final temperature is 23⁰C. Calculate the specific heat capacity of brass.

Solution:

Calculation of heat lost by brass rod

Mass of brass (m_1) = 0.4 kg

Specific heat of brass (s_1) = ?

Initial temperature of brass rod (t_1) = 100^0C

Final temperature of brass rod (t_2) = 23^0C

So, change in temperature of brass rod (Δt) = $100^0 - 23^0$

$= 77^0C$

Thus, heat lost by brass rod (Q1) = $m_1 \times s_1 \times \Delta t$

$= 0.4 \times s_1 \times 77$

$= 30.8 \times s_1$ joules

Calculation of heat gained by water

Mass of water (m_2) = 1.0 kg

Specific heat capacity of water (s_2) = 4.2 x 103 J/kg^{0}C

Initial temperature of water (t_1) = 20^0C

Final temperature of water (t_2) = 23^0C

So, change in temperature of water, (Δt) = $23^0 - 20^0$

$= 3^0C$

Thus, heat gained by water (Q2) = $m_2 \times s_2 \times \Delta t$

$= 1.0 \times 4.2 \times 10^3 \times 3$

$= 12600$ joules

According to the principle of calorimetry,

Heat lost = Heat gained

So, $30.8 \times s_1 = 12600$

And, $s_1 = 12600/30.8$

$s_1 = 409.09$ J/kg^{0}C

$\therefore$ The specific heat of brass is 409.09 J/kg^{0}C

Summary

1. Matters exist in three physical states, i.e solid, liquid and gas. The physical states of a substance can be changed either by heating or cooling.
2. The amount of heat absorbed by a unit mass of the substance to change its state without the change of temperature is called the latent heat of a substance.
3. The S. I. unit of heat is joule and that of the mass is kilogram. So, the S.I. unit of latent heat is joules per kilogram (J/kg). It is generally represented by the letter L.

4. Latent heat of fusion is called latent heat of melting. It is also called solid to liquid change.
5. The amount of heat required to convert unit mass of a substance from the solid state to the liquid state without change of temperature is called latent heat of fusion of a substance.
6. Amount of heat absorbed or given out = Mass of the substance x Latent heat of substance during the change of state of a substance
7. The heat required to change a liquid into the vapour state is called latent heat of vaporization.
8. The amount of heat required to change a unit mass of the substance from the liquid state to vapour state without change of temperature is called the latent heat of the vaporization of a substance.
9. Heat equation can be defined as the product of mass, specific heat capacity and change in temperature is equal to the heat gained or heat given out by the body.
10. The measurement of heat is called calorimetry. So, the principle of calorimetry means the principle of measurement of heat.
11. According to the principle of calorimetry, when two bodies of different temperatures are kept in thermal contact, the heat flows from a body with a higher temperature to a body with the lower temperature until each of them has equal temperature.

Exercise

A. Tick (√) the best alternative from the following:

1. The physical state of matter can be changed by

   i. Heating ii. Cooling

   iii. Both (i) and (ii) iv. None of the above

2. The S.I. unit of latent heat is given by

   i. N/kg ii. J/kg

   iii. W/kg iv. Pa/kg

3. What is called the Latent heat of fusion?

   i. Latent heat of melting ii. Latent heat of boiling

   iii. Latent heat of vaporization iv. Both (i) and (iii)

4. Amount of heat absorbed or given out is given by the formula

   i. Mass of the substance x Latent heat of substance during the change of state of a substance

   ii. Weight of the substance x Latent heat of substance during the change of state of a substance

   iii. Proton of the substance x Latent heat of substance during the change of state of a substance

   iv. Neutron of the substance x Latent heat of substance during the change of state of a substance

5. Heat equation can be defined as

   i. The product of mass and specific heat capacity is equal to the heat gained or heat given out by the body.

   ii. The product of mass and change in temperature is equal to the heat gained or heat given out by the body.

   iii. The product of mass, specific heat capacity and change in temperature is equal to the heat gained or heat given out by the body.

   Both (i) and (ii)

B. Give short answers to the following questions:

1. Mention three states of matter.
2. How can we change the states of matter?
3. What is called the latent heat of a substance? Also write its SI unit.
4. Latent heat of fusion is called latent heat of melting. Why?
5. Define the latent heat of fusion of a substance.
6. What is called the latent heat of vaporization of a substance?
7. Define heat equation and give its equation.
8. Define calorimetry and principle of calorimetry.

C. Give long answers to the following questions:

1, How can we demonstrate the latent heat? Explain.

2. How can we calculate the heat absorbed or given out during change of state of a substance?

3. How can we calculate the heat absorbed or given out during vaporization and heat given out during condensation?

4. Derive the relation Q = $ms\Delta t$, where the letters used have their usual meaning.

5. Spread 200 grams of sand homogeneously is one saucer and put 200 grams of water in another saucer and place them under the sun. Measure the temperatures after 10 minutes. The temperature of sand is found to be more than water, why?

6. Explain in short about the principles of calorimetry.

D. Solve the following numerical problems:

1, Calculate the amount of heat required to convert 500g of ice into water without change of temperature. (Latent heat of ice = 3.34×10^5 J/kg)

(Ans:16.7 x 10^5 J)

2. How much heat must be added to raise the temperature of 40^0g of water from 30^0C to 90^0C? *(Ans: 1.008 x 10^5 J)*

3. An electrical kettle contains 500gm of water at 50^0C. How much heat is to be supplied to boil the water? (Assume that water boils at 100^0C and the kettle takes no heat)

(Ans: 1.05 x 10^5 J)

4. The temperature of 20 kg water in the radiator of engine of a car is 30^0C. If the temperature of water increases upto 100^0C after the engine is heated, what quantity of heat will the water absorb?

(Ans: 5.88 x 106 J)

5. What will be the quantity of heat required to raise the temperature of 2kg paraffin by 10^0C if 44000 joules of heat energy is required to raise the temperature of the paraffin by 200C?

(Ans: 4.4 x 10^4 J)

5. Hot water of mass 7 kg at 98^0C is cooled for taking bath by mixing 14 kg of water at 10^0C. Calculate the final temperature of water. Specific heat capacity of water is 4200J/kg^{0}C. Neglect the heat taken by the bucket.

(Ans: 39.33^{0}C)

Project Work

Take two beakers of the same size. Take 100 ml of edible oil in one beaker and 100 ml of water in the other. Measure the temperature of both of them. Heat them with two spirit lamps of identical size and burning capacity. Exchange the spirit lamps every two minutes. Measure separately the temperature of both of them every two minutes. Which of them gets fast rise in the temperature and why? Based on this observation, prepare a report and present it to the class.

Glossary

Latent	:	hidden
Melting	:	change of state from solid to liquid
Joule	:	SI unit of energy which is equal to one Newton metre
Calorimetry	:	measurement of quantity of heat
Saucer	:	a small, round, shallow dish to hold a cup

Unit 5

Light

Christian Huygens
(1773 – 1829)

Born 14 April 1629 in the Hague, Netherlands and died 8 July 1695 in the Hague, Netherlands. The Dutch mathematician, astronomer, and physicist Christian Huygens was the first to recognize the rings of Saturn made pioneering studies of the dynamics of moving bodies, and was the leading advocate of the wave, or pulse, the theory of light.

Introduction

We have already discussed about light and its phenomena in the junior class. A lens is a portion of transparent medium bounded by two refracting surfaces at least one of which is spherical. Lenses are used in many optical instruments such as camera, telescope, microscope, binocular and other several optical devices.

In this unit, we will discuss about lens, its power and magnification, optical instruments such as binocular, compound microscope and terrestrial telescope.

5.1 Lens

In our everyday practice, we see many people around us who are wearing spectacles to observe nearby objects or distant objects. Scientists and Pathologists use microscope to observe microorganism present in different samples. Astronauts use telescopes to view heavenly objects in the sky or objects found on the earth. With the help of magnify glasses, we see the tiny specimens in the laboratory. The surface of magnifying glass is not plane, but thicker in the middle and thinner at the edges. Thus, lens is a portion of transparent refracting medium bounded by two surfaces at least one of which is a spherical surface. Lens is usually made up of glass or plastics. There are two spherical surfaces. They may be convex or concave.

Types of Lens

Two types of lens are used in our everyday practice.

They are:

(a) Convex lens (Converging lens)

(b) Concave lens (Diverging lens)

Convex lens (Converging lens): The lens which is thicker in the middle and thinner at the edges is called convex lens. These lenses are of three types. They are: biconvex, plano convex and concavo-convex.

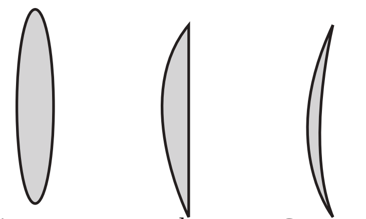


Fig. 5.1

Concave lens (Diverging lens): The lens which is thinner in the middle and thinner at the edges is called concave lens. These lenses are also of three types. They are: biconcave, planoconcave and convexo-concave.

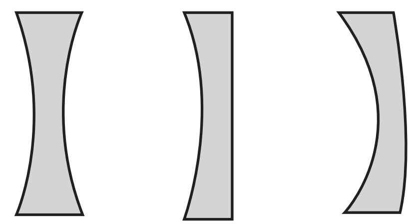


Fig. 5.2

Power of a Lens

Optical power is the degree to which a lens, mirror, or other optical system converges or diverges light. It is equal to the reciprocal of the focal length of the device: P = 1/f(m) . High optical power corresponds to short focal length. The SI unit for optical power is the inverse metre, which is commonly called the dioptre.

Converging lenses have positive optical power, while diverging lenses have negative power. When a lens is immersed in a refractive medium, its optical power and focal length change.

For two or more thin lenses close together, the optical power of the combined lenses is approximately equal to the sum of the optical powers of each lens: $P = P_1 + P_2$. Similarly, the optical power of a single lens is roughly equal to the sum of the powers of each surface. These approximations are commonly used in optometry.

An eye that has too much or too little refractive power to focus light onto the retina has a refractive error. A myopic eye has too much power so light is focused in front of the retina. Conversely, a hyperopic eye has too little power so when the eye is relaxed, light is focused behind the retina. An eye with a refractive power in one meridian that is different from the refractive power of the other meridians has astigmatism. An isometropia is the condition in which one eye has a different refractive power than the other eye.

Dioptre is the unit of of the optical power of a lens or curved mirror, which is equal to the reciprocal of the focal length measured in metres. It is thus a unit of reciprocal length. For example, a 3-dioptre lens brings parallel rays of light to focus at 1/3 metre. A flat window has an optical power of zero dioptres, and does not converge or diverge light. Dioptres are also sometimes used for other reciprocals of distance, particularly radii of curvature and the vergence of optical beams.

Power of a lens is defined as the ability of the lens to converse or diverse the rays of light incident on it.

The power is said to be positive if the lens converges the rays and negative if it diverges. If the lens diverges the rays of light, shorter will be the focal length of a lens. Thus, the power of a lens is defined as the reciprocal to its focal length (f) measured in metre.

Symbolically, $P = 1/f(m)$

If the focal length is measured in centimetre

Symbolically, $p = 100/f(cm)$

Unit of Power of Lens

The SI unit of power of a lens is dioptre, which is denoted by D.

$P = 1/f(m)$ if $f = 1$ m

Then $P = 1$ D

A lens is said to be a power of 1 dioptre if its focal length is 1 meter long. In the case of convex lens, the power of a lens is positive. But in concave lens, the power is negative.

Uses of Lens

Convex:

a. It is used in optical instruments such as photographic camera, telescope, etc.

b. It is used in spectacles for the remedy of hypermetropic defect of vision.

Concave:

a. It is used in various optical instruments.

b. It is used for the remedy of myopic defect of vision.

Numerical Illustration

1. If power of a lens is + 2 D, then what is the focal length of the lens?

   Solution:

   Power of a lens (P) = +2 D

   Focal length (f) = ?

   We know that

   Power of lens (P) = 1/(f(m))

   Or, Focal length (f) = 1/P

   = 1/2

   = 0.5 m

   = +50 cm.

   ∴ The focal length of the lens is +50 cm.

2. If a lens is of focal length 50 cm, then calculate its power.

   Solution:

   Focal length (f) = 50 cm

   Power (P) = 100/f(cm)

   = 100/50 = 2.0 D

   ∴ The power of the lens is +2.0 D.

3. Calculate the focal length of lens whose power is - 2.25 D.

   Solution:

   Power of a lens (P) = -2.25 D

   Focal length (f) = ?

   We know that

   Power of lens (P) = 1/f(m)

   Or, Focal length (f) = 1/P

   = -1/2.25

   = - 0.44 m

   = - 44 cm.

   ∴ The focal length of the lens is 44 cm.

Magnification

Magnification is the process of enlarging something only in appearance, not in physical size. This enlargement is quantified by a calculated number also called "magnification". When this number is less than one, it refers to a reduction in size, sometimes called "minification" or "de-magnification". Thus, the magnification is the ratio of size of image to the size of object. In other words, the ratio of image distance to object distance is called magnification.

Examples of magnification

(a) A magnifying glass, which uses a positive (convex) lens to make things look bigger by allowing the user to hold them closer to their eye.

(b) A telescope, which uses its large objective lens to create an image of a distant object and then allows the user to examine the image closely with a smaller eyepiece lens thus making the object look larger.

(c) A microscope, which makes a small object appear as a much larger object at a comfortable distance for viewing. A microscope is similar in layout to a telescope except that the object being viewed is close to the object, which is usually much smaller than the eyepiece.

(e) A slide projector, which projects a large image of a small slide on a screen.

Magnification as a number

Optical magnification is the ratio between the apparent size of an object (or its size in an image) and its true size, and thus it is a dimensionless number. Optical magnification is sometimes referred to as "power" (for example "10 × power"), although this can lead to confusion with optical power.

Establishment of the relationship among I, O, v and u

Let an object PQ be placed on the principal axis of a convex lens and perpendicular to its principal axis between F_2 and F_1. A ray PL parallel to the principal axis passes through F_2 after refraction through it and another ray PO passes undeviated through its LP'. These refracted rays LP' and OP' meet at P'. Hence, P' is the real image of P and P'Q' is the real image of PQ.

Δ QPO is similar to ΔQ'P'O, we have

$\angle$PQO = $\angle$P'Q'O (∴ both being 90^0)

$\angle$POQ = $\angle$P'OQ' (∴ being vertically opposite angles)

$\angle$QPO = $\angle$Q'P'O (∴ being remaining angles of each triangle)

∴ ΔQPO and ΔQ'P'O are similar triangles. In this case, we write, So, $\frac{Q'P'}{QP} = \frac{OQ'}{OQ}$

$$\frac{\text{height of image}}{\text{height of object}} = \frac{\text{image distance}}{\text{object distance}}$$

$$\therefore \frac{I}{O} = \frac{v}{u}$$

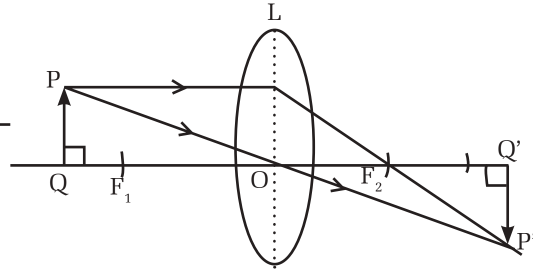


Fig. 5.3 Magnification

Hence, magnification is the ratio of image distance to the object distance. In another words, it is the ratio of height of image to the height of object. So it can be measured by using these two ratios. It provides the information about the size of image. It means that whether the size of image is same or magnified or diminished. Magnification of a lens less than one means the size of image is smaller than the object. Similarly, magnification of a lens more than one means image is larger the object. If magnification of a lens is equal to one, then the size of image is equal to the size of the object.

Numerical Illustration

1. An object is placed at a distance of 3 cm from a convex lens. If a real image forms at a distance of 12 cm from the lens, then how much does the lens magnify the size of object?

   Solution:

   Object distance (u) = 3 cm

   Image distance (v) = 12 cm

   Magnification (m) = ?

   We know that

   m = v/u

   or, m = 12/3

   or, m = 4

   $\therefore$ The lens magnifies the image four times the size of the object.

2. A lens magnifies an image 3 times the size of the object. If an object is placed at a distance of 4 cm from a convex lens, then how far the image is formed from the convex lens?

Solution:

Object distance (u) = 4 cm

Magnification (m) = 3

Image distance (v) = ?

We know that

m = v/u

or, v = m x u

or, v = 3 x 4

or, v = 12 cm

∴ The real image is formed at a distance of 12 cm from the convex lens.

Optical instruments

Eye is a kind of optical instrument and is used to make scientific observations. However, it is not possible to see everything using the eyes only. Therefore, several instruments have been in use to observe and interpret the images distinctly. For example, a planet is seen by naked eye but its details can be seen only after magnifying the size of its image. Similarly small bacteria cannot be seen by naked eyes unless magnified manifolds instruments are used which aid in obtaining the image of distant objects or minute objects. Telescope, microscope, cameras, etc. are some examples of optical instruments.

1.2 Terrestrial Telescope

As we have known that an astronomical telescope is a device, which is used to view the heavenly objects in the sky. But this telescope cannot be used to view the objects on the surface of the earth. To view the objects on the surface

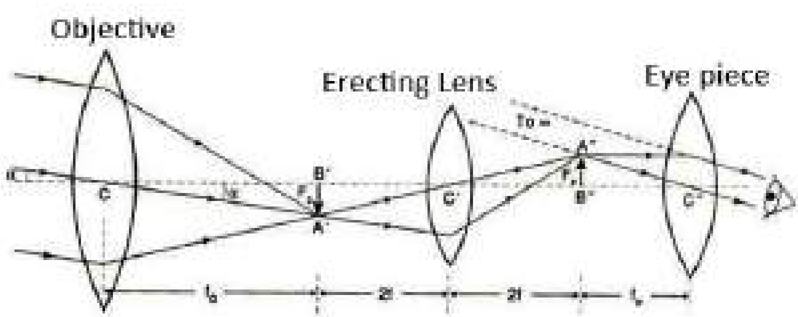


Fig. 5.4 Terrestrial telescope

of the earth, people use another type of telescope, called terrestrial telescope. Therefore, **terrestrial telescope is an optical viewing device, which is primarily designed to look at objects on the earth's surfaces, rather than in the night sky**. It is also known as spyglass.

Construction and Working

Terrestrial telescope consists of three lenses with an objective lens with a very short focal length and an eyepiece lens having slightly longer focal length. These three lenses are placed inside two metallic tubes blackened from inside such that they can easily slide into one another.

As astronomical telescope forms an inverted image of the object at infinity. But terrestrial telescope forms the final image erect with respect to the object. The third lens of short focal length f is placed at 2f which forms inverted image of the object. This image serves as the object for the eye-piece. The lens placed in the center of the telescope which actually erects the image is called as the erecting lens.

1.2 Compound Microscope

To view very tiny objects in laboratory or pathology, we use a device, called microscope. **The optical instruments which magnifies tiny objects and enables us to see them very clearly is called microscope**. There are two types of microscopes in our daily use. They are Simple microscope and Compound microscope.

Compound microscope: Compound microscope is a device, which is used for producing greater magnification than the simple microscope by combining two convex lenses. This microscope can magnify an object to about 1000 times.

Principle: Compound microscope is based on the principle that an image formed by the objective lens serves as an object for the eye lens.

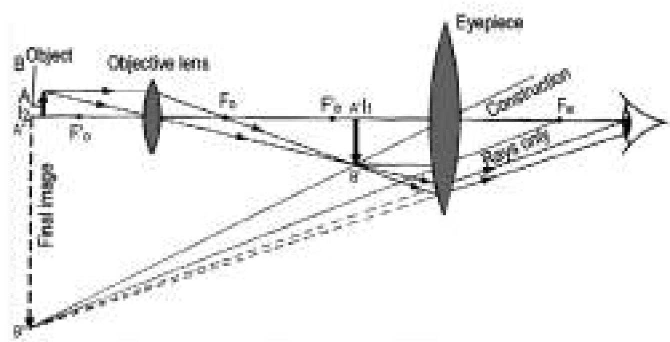


Fig. 5.5 Compound microscope

Construction and working: Compound microscope consists of an objective lens with a very short focal length and an eyepiece lens having slightly longer focal length. These two lenses are placed inside two metallic tubes blackened from inside such that they can easily slide into one another. The objective lens is fixed at one end of the smaller tube, whereas the eye piece is fixed at one end of the bigger tube.

The object AB is placed just beyond the focus F of the objective lens forms a real, inverted and magnified image A'B' which is close to the focus of the eyepiece. The eyepiece which serves as a simple magnifying glass produces an image of A'B' at A''B''. The final image is virtual and inverted.

Uses:

a. It is used by pathologist to examine stool, blood and urine of a patient.

b. It is also used in study and research.

1.3 Binocular

Binocular is an optical device consisting of two small telescopes filled together side by side, each telescope having two prisms between the eye piece and objective for erecting the image.

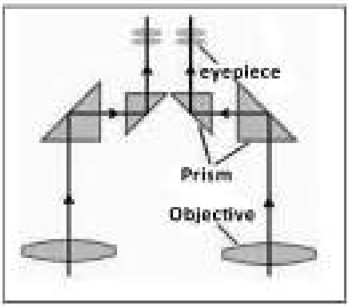
A hand held device consisting a series of lenses and prisms, used to magnify object so that they can be better seen from a distance and looked at through both eyes is called binocular.

Structure and working of binoculars

The objective lenses are situated at each end of the binoculars. The purpose of the objective lens is to collect light from the object that the user is looking at and bringing the collected light into focus in the eyepiece lens, which creates a visible and magnified image.

Binoculars have two prisms. Inside the prisms, the light to each eye is bent after being reflected by a pair of glass prisms. This happens so that the power of the binoculars is equal to the magnifying power of a longer telescope. The light from the object you are observing enters the objective lens. Then it enters two prisms set at right angles to each other. The prisms are arranged at right angles so that the final image can be placed right side up. Then the light exits the ocular (eyepiece) lens giving you a magnified image with the correct orientation. No light is lost because of total internal reflection. These prisms can lengthen the light path between the objective lens and the ocular lens, thereby, increasing the magnification.

Although, not only the binoculars can magnify an image, they can also make an image seen smaller by zooming out.



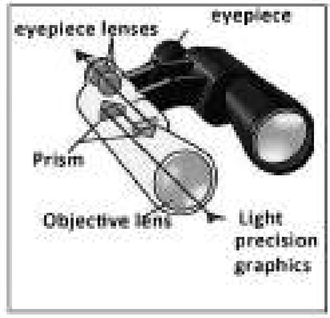


Fig. 5.6

Uses of binoculars

They are almost a necessity for the astronomer, hunter, saltwater fisherman, boater, traveler, birdwatcher, sports man, and experienced traveler. Binoculars are designed to give a correctly oriented, right side up view. This makes them ideal for terrestrial viewing, or for locating astronomical objects in the night sky.

Summary

1. A lens is a portion of transparent refracting medium bounded by two surfaces at least one of which is a spherical surface.
2. Power of lens is defined as the ability of the lens to converse or diverse the incident rays. The SI unit of power of a lens is dioptre, which is denoted by D.
3. A lens is said to be a power of 1 dioptre if its focal length is 1 meter long. In the case of convex lens, the power of a lens is positive. But the power of a concave lens is negative.
4. Magnification is the ratio of image distance to the object distance or it is the ratio of height of image to height of object. There are various types of magnification.
5. Terrestrial telescope is an optical viewing device, which is primarily designed to look at objects on the Earth surfaces, rather than in the night sky.
6. Compound microscope is a device, which is used for producing greater magnification than the simple microscope by combining two convex lenses.
7. Binocular is an optical device consisting of two small telescopes fitted together side by side, each telescope having two prisms between the eyepiece and objective for erecting the image.
8. Prisms let you to see a correctly oriented image when you look through a pair of binoculars.

Exercise

A. Tick (√) the best alternative from the following.

1. What is the SI unit of power of a lens?

   i. dioptre　　ii. metre

   iii. dioptre/meter　　iv. metre/ dioptre

2. In the case of convex lens, the power of a lens is:

   i. Positive　　ii. negative

   iii. Both (i) and (ii)　　iv. None of the above

3. What is magnification?

   i. the ratio of object distance to the image distance

   ii. the ratio of height of object to the height of image

   iii. the ratio of image distance to the object distance

   iv. Both (i) and (ii)

4. Terrestrial telescope is an optical viewing device, which is used to see:

   i. the objects on the earth surfaces

   ii. the objects in the sky

   iii. the minute objects

   iv. Both (i) and (ii)

B. Give short answers to the following questions:

1. What is a concave lens? Why is concave lens called a diverging lens?
2. Define convex lens. Why is convex lens called a converging lens?
3. What is meant by a microscope?
4. Draw a neat and clear ray diagram of a compound microscope. Also mention any two uses of it.
5. What is meant by a terrestrial telescope?
6. Draw a neat and clear ray diagram of a terrestrial telescope.
7. Write down any two uses of a terrestrial telescope.
8. Distinguish between the nature of image formed by the eye lens and objective lens of telescope.

9. Distinguish between compound microscope and terrestrial telescope.

C. Give long answers to the following questions:

1. What is meant by magnification? Give four examples of it.
2. Establish the relationship among I, O, v and u.
3. Explain the working process of a terrestrial telescope.
4. Explain the working process of a compound microscope.
5. How does a binocular work?
6. Mention the uses of binoculars.

D. Solve the following numerical problems:

1. A hand lens has the power of 20 dioptre. If the letters of a book are to be read with the help of this hand lens, calculate how far from the lens must the book be held. *(Ans: within 0.05 m)*
2. The power of a lens is + 5 D. What are the focal length and nature of the lens? *(Ans: 0.2 m)*
3. A lens has a power of –0.5 D. What are the focal length and the nature of the lens? *(Ans: 2 m)*
4. An object is placed at a distance of 4 m from a convex lens. If a real image forms at a distance of 16 cm from the lens, calculate its magnification. *(Ans: 4)*
5. What is the magnification produced by a lens if it forms a real image on a screen 20 cm behind the lens of an object 4 cm in front of the lens? *(Ans: 5)*

Project Work

1. Take about 30 cm long hollow bamboo piece or a paper tube. Take two convex lenses of focal lengths of 5 cm and 10 cm and attach them one on either end of the tube. To support the lenses, cut paper rings and paste them on either side of the lens. By holding the lens with longer focal length next to the eye. Observe some tiny objects keeping them at the distance of about 6 to 7 cm from the objective lens. Do the objects look bigger? Based on observation, prepare a report and present it to the class.

2. Take two plastic pipes; one foot long pipe of diameter 2 inch and next 1 foot long pipe of diameter 4 inch. Take two lenses each of focal length 5 cm and 20 cm. Attach the lens of focal length 5 cm at one end of the smaller tube and next lens at one end of the larger tube. Make a ring of cardboard so that the smaller pipe gets tightly inserted inside the larger tube. Now insert the smaller tube inside the ring of the card board then the ring inside the large tube. Observe the heavenly bodies with the objective lens toward them by sliding the smaller tube up and down. On the basis of the observation, prepare a report and present it to the class.

Key words

Optometry : the occupation of measuring egesight

Minification : a reduction in size

Undeviated : goes straight or does not bend

Diverting : scattering

Twilight conditions : when an eye is not yet fully dark-adapted

Astigmatism : a type of refractive error in which the eye does not focus light on retina

Isometropia: equality in refraction of the two eyes

Vergence: the turning motion of the eyeballs toward or away from each other

Unit 6

Current Electricity and Magnetism

Heinrich Rudolf Hertz (1857 - 1894) was a German physicist who first conclusively proved the existence of the electromagnetic waves theorized by James Clerk Maxwell's electromagnetic theory of light. Hertz proved the theory by engineering instruments to transmit and receive radio pulses using experimental procedures that ruled out all other known wireless phenomena.

Heinrich Rudolf Hertz (1857 – 1894)

Introduction

Electricity is the form of energy which can be transformed from a system to its surrounding. Electricity can be generated by the use of magnet. Generally a magnet attracts magnetic substances. The motion can also be generated in a current carrying conductor placed when kept to move freely in another magnetic field. To change the magnitude of the current, a transformer is used in our daily life.

In this unit we will discuss the properties of magnetic substances, electromagnetic induction, faraday's law of electromagnetic induction, diode, resistor, transistor, generator, motor effect and transformer.

Properties of magnetic substances

The substances which are attracted by a magnet are called magnetic substances. The magnetic substances are of mainly three types. They are ferromagnetic substances, paramagnetic substances and diamagnetic substances. Substances which are strongly attracted by a magnet are called ferromagnetic substances, whereas the substances which are less attracted by the magnet are called paramagnetic substances, and which are not attracted by the magnet are called diamagnetic substances.

The molecules in a magnetic substance are arranged in close chain like structure, where their polar strength is neutralized by each of them. When it becomes magnet, its molecular structure is changed. As a result the molecular structure is modified as open chain like structure and it possesses magnetic strength in its polar region.

Properties of magnetic substances

Magnetic substances possess different properties. Some of their properties can be illustrated in the following ways:

Properties of Dia magnetic substances

a. These substances are weakly repelled by magnets.

b. A diamagnetic substance moves from a stronger to a weaker parts of the field in non uniform magnetic field.

c. A diamagnetic rod freely suspended in a uniform magnetic field, slowly turns to set at right angle to the applied magnetic field.

Properties of Para magnetic substances

a. These substances are weakly attracted by a magnet.

b. A paramagnetic substance moves from the weaker to the stronger part of the field in non uniform magnetic field.

c. When a rod of paramagnetic substance is freely suspended, in a magnetic field, it aligns along the field.

Properties of Ferro magnetic substances

The properties of ferromagnetic substances are similar to that of paramagnetic substances but are exhibited in a large scale.

a. These substances are strongly attracted by a magnet.

b. A ferromagnetic substance moves from weaker part of the magnetic field to the stronger part in a non uniform magnetic field.

c. When a rod of a ferromagnetic substance is freely suspended in a uniform magnetic field, it rotates and aligns itself parallel to the field.

Magnetization

The process of making a magnet from magnetic substances either by rubbing or passing electricity using is called magnetization.

Permeability

The degree to which the lines of force can penetrate in a substance placed in magnetizing field is called the permeability of the substance.

Susceptibility

The property of the magnetic substance due to which magnetization of it takes place to different extents is called susceptibility.

Magnetic flux and its variation

Magnetic lines of force start from N pole and end at South Pole outside a magnet but they start from South Pole and end at North Pole inside the magnet. These magnetic lines of force are always in continuous curve and they never intersect each other. Let us consider a bar magnet is placed at rest facing its north pole towards a coil. Some magnetic lines of force will pass through the coil while travelling from its north pole to its south pole. Therefore, magnetic flux through a surface is defined as the numbers of force passing the surface held perpendicular to these lines of force. If a magnet or a coil moved closer to one another, then the magnetic flux through the surface of the coil increases. Conversely, they are moved away from each other, and then it decreases.

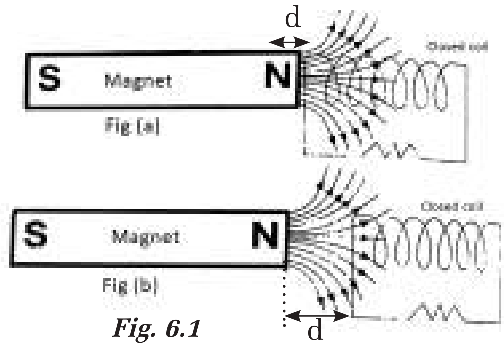


Fig. 6.1

Electromagnetic Induction

Current is produced in a conductor when it is moved through a magnetic field because the magnetic lines of force are applying a force on the free electrons in the conductor and causing them to move. This process of generating current in a conductor kept in a changing magnetic field is called electromagnetic induction. This is called induction because there is no physical connection between the conductor and the magnet. The current is said to be induced in the conductor by the magnetic field.

The requirement for this electromagnetic induction to take place is that the conductor (piece of wire/coil) must be held perpendicular to the magnetic lines of force (magnetic field) in order to produce the maximum force on the free electrons. The direction of induced current is determined by the direction of lines of force and the direction of movement of conductor in the field.

The process of generating electric current in a conductor kept in a changing magnetic field is called electromagnetic induction.

Activity to demonstrate electromagnetic Induction

The following activity can be helpful in understanding the concept of electromagnetic induction. To perform this activity, we need a long piece of metal wire (PQ), a permanent horse shoe type magnet and a galvanometer (G). First of all, a long piece of metal wire is held between the N and S poles of a permanent horse shoe type magnet as shown in the figure 6.2. Then, these two ends of the wire are connected to a very sensitive current detecting instrument called galvanometer. There is no deflection in the needle of the galvanometer where there is no motion in the wire. This is due to absence of current in the wire. In another words, there is not production of current in the wire when the wire is in rest position.

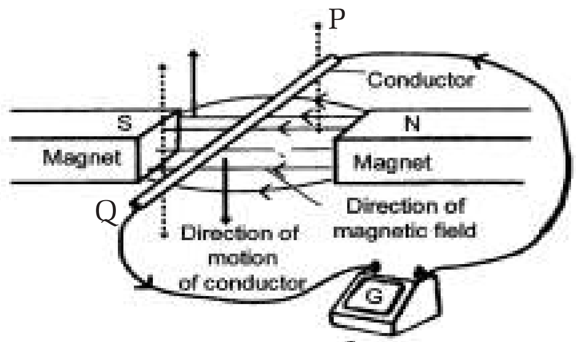


Fig. 6.2

On moving the wire PQ quickly downwards, the galvanometer pointer shows a momentary deflection. This is due to induced current produced in the wire. If the same wire is moved quickly upwards, then the galvanometer again shows a momentary deflection in the opposite direction as compared to previous one. This shows that when the direction of the motion of wire in the magnetic field is reversed, the direction of induced current is also reversed. Conversely, the direction of induced current in the wire also be reversed by reversing the positions of the poles of the bar magnet.

The galvanometer pointer shows deflection only when there is relative motion between the coil and the magnet. The relative motion between the coil and the magnet causes the change in magnitude flux.

Faraday's laws of electromagnetic induction

On the basis of a series of experiments, Michael Faraday concluded his ideas giving certain laws regarding electromagnetic induction, which are called laws of electromagnetic induction. He stated the laws in the following ways:

a. Whenever the magnetic flux (f) linked with a closed circuit (a coil) changes, an induced emf is produced in it.

b. The induced emf lasts until the change in magnetic flux continues.

c. The magnitude of the induced emf is directly proportional to the rate of change of magnetic flux within the closed coil.

   i.e. Induced emf $\propto$ rate of change of magnetic flux within the closed circuit.

Activity 6.1

Wind up about 50 turns of an insulated copper wire as shown in the figure. Connect two ends of the wire with a galvanometer. A powerful bar magnet should be moved in and out of a match box very quickly. Observe what effect is seen in the needle of the galvanometer as the bar magnet

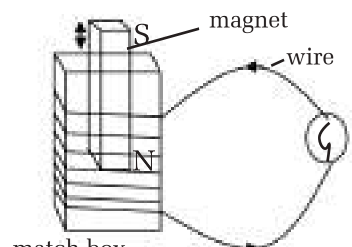


Fig. 6.3

is thus moved in and out of the match box very quickly.

On the basis of this observation, answer the following questions:

i. How does the needle of the galvanometer deflect?

ii. Does the needle deflect in the same direction as before even after the magnet is moved out?

iii. Why does the needle of the galvanometer deflect as the magnet is moved in and out?

Fleming's Right Hand Rule for the direction of induced current

The direction of the induced current produced in a straight conductor moving in a magnetic field can be predicted by using Fleming's Right Hand Rule. It states that " if the first three fingers of right hand are stretched mutually perpendicular to each other such that the thumb and index finger represent the direction of movement of conductor and the direction of magnetic field respectively, then the middle finger indicates the direction of induced current.

6.2 Activity

Stretch the first three fingers of your right hand as shown in the figure. In this case, 90^0 angle should be subtended between the middle finger and the index finger and between the index finger and the thumb finger as well. Consult your teacher whether the demonstrated hand matches the Fleming's Right Hand Rule.

Fleming's right-hand rule

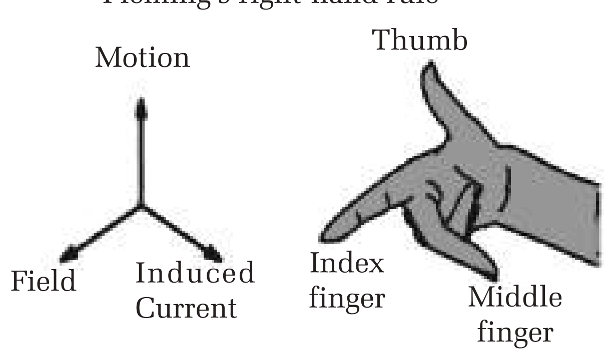


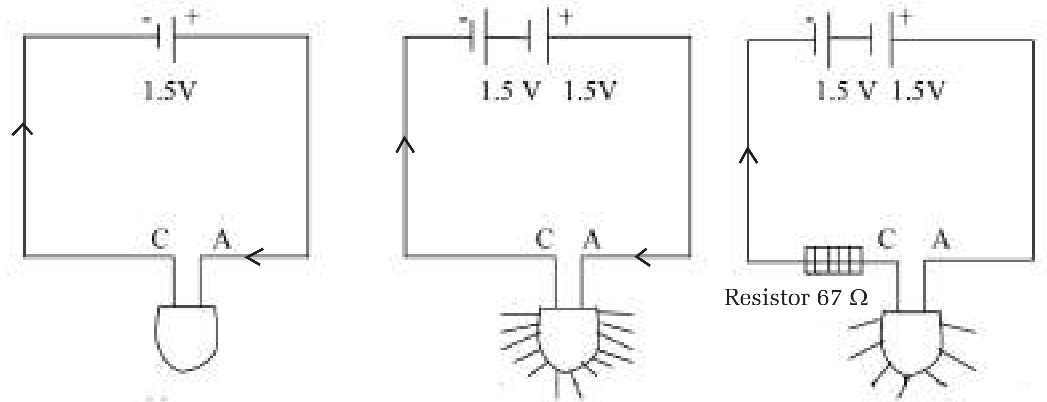
Fig. 6.4

Resistor, Transistor and Diode

Resistor

The electric device which is used to limit or regulate the proper flow of current in the electric circuit is called a resistor. It is also called electric load like electric bulb, electric fan, electric heater, etc. A resistor converts electric energy into other forms of energy. A number of resistors have to be used in an electric circuit to get desired value of current.

Suppose, it needs 2V across its two terminals to light up an LED. Study the following diagrams to understand the use of resistors and to select a suitable resistor for a circuit.



(i) No light (ii) Too much current flow (iii) Ideal operating condition

Fig. 6.5

In the figure (iii) the LED lights up with an appropriate brightness without placing too much load either on the LED or on the batteries.

Consider the fig. (iii), when the LED is working normally the voltage across its terminals will be 2V. Now the voltage V_R across the two terminals of the resistor is given by,

V_R = 3V (of battery) - 2V (normal operating voltage of LED)

= 1V

Let the current to be passed through the LED for normal operation Iled be 15 mA (1mA = 1/(1000) A)

∴ The capacity of resistor R, will be calculated by using Ohm's law.

R = V_R/ Iled = 1/0.015 = 67 ohm

This is the capacity of the resistor to be connected.

Resistors can be connected together in the circuit in the following three ways:

(a) Series connection (b) Parallel connection (c) Mixed connection

Here, we will deal with the study of series connection and parallel connection of resistors in the following discussion.

Series connection of resistors

Suppose R_1, R_2 and R_3 are three resistors connected in series with a source of potential difference 'V'. Let, current 'I' flows through the circuit. In this connection, same magnitude of current passes through each resistor.

By applying Ohm's law, the potential differences across the three resistances will be

$V_1 = IR_1$, $V_2 = IR_2$, $V_3 = IR_3$

If R be the equivalent resistance of the series connection, then on applying a voltage 'V' across it, the same current 'I' must flow through it.

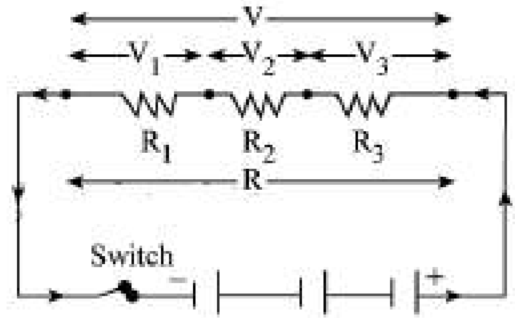


Fig. 6.6 Series connection of resistors

So,

$$V = IR$$

But, $V = V_1 + V_2 + V_3$

Or, $IR = IR_1 + IR_2 + IR_3$

$\therefore R = R_1 + R_2 + R_3$

When a number of resistors in a circuit are connected from end to end in order to get same current flowing through each of them in succession, the resistors are said to be connected in series connection.

Main characteristics of series connection of resistors

The following are the most important characteristics of series connection of resistors:

(a) Resistors are connected from end to end.

(b) The equivalent resistance R is equal to the sum of individual resistances connected in the circuit. i.e., $R = R_1 + R_2 + R_3$

(c) The potential difference remains constant, but the current decreases with the increase in number of resistors.

(d) The total voltage (V) of circuit is equal to the sum of the voltage of individual resistance. i.e., $V = V_1 + V_2 + V_3$.

(e) The same current will pass through each of the resistor. i.e., $I = I_1 = I_2 = I_3$

(h) If the circuit is broken anywhere between the loads, all the loads will stop their functioning.

Parallel connection of resistors

Suppose R_1, R_2 and R_3 are the three resistors connected between two common points A and B. Let a current 'I' flows through the circuit when a cell of voltage 'V' is connected across the connection. The current 'I' at point 'A' is divided into three parts I_1 along R_1, I_2 along R_2 and I_3 along R_3. These three parts recombine at point 'B' to give the same current 'I'.

$\therefore I = I_1 + I_2 + I_3$

By applying Ohm's law,

$$I_1 = \frac{V}{R_1}, \quad I_2 = \frac{V}{R_2}, \quad I_2 = \frac{V}{R_3} \quad (\because I = \frac{V}{R})$$

If R be the equivalent resistance of the parallel connection, then

$$\therefore \frac{V}{R} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3} \quad (\because I = I_1 + I_2 + I_3)$$

$$\therefore \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$$

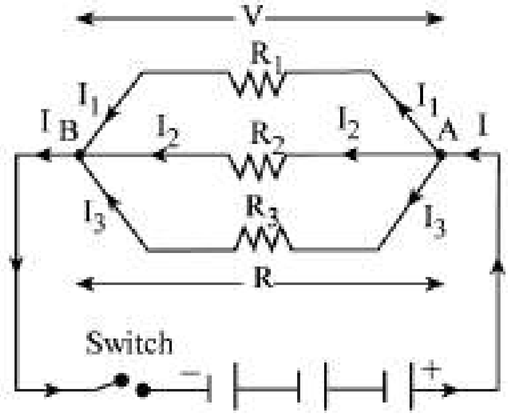


Fig. 6.7 Parallel connection of resistors

When a number of resistors in a circuit are connected between two common points in order to get different current, the resistors are said to be connected in parallel connection.

Main characteristics of parallel connection of resistors

(a) Resistors are connected between two common points.

(b) Individual use of resistor is possible.

(c) If any one of the resistors stops functioning, the rest resistors continue functioning.

(d) The reciprocal of the equivalent resistance is equal to the sum of the reciprocal of individual resistors connected in the circuit.
i. e. $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$

(e) The voltage across each resistance is same, which is equal to the applied voltage. i.e. $V = V_1 = V_2 = V_3$

(f) The current passing through the circuit is equal to the sum at the current through the individual resistance, i.e. $I = I_1 + I_2 + I_3$

Semiconductor

A semiconductor is a solid chemical element or compound that can conduct electricity under some conditions. It makes a good medium for the control of electrical current. The specific properties of a semiconductor depend on the impurities added to it. An N-type semiconductor carries current mainly in the form of negatively-charged electrons, in a manner similar to the conduction of current in a wire. A P- type semiconductor carries current predominantly as electron deficiencies called holes. A hole has a positive electric charge, equal and opposite to the charge on an electron. In a semiconductor material, the flow of holes occurs in a direction opposite to the flow of electrons. Germanium and Silicon are the examples of semiconductor.

P-N junction

A P-N junction is a boundary or interface between two types of semiconductor material, p-type and n-type, inside a single crystal of semiconductor. The "p" (positive) side contains an excess of holes, while the "n" (negative) side contains an excess of electrons. The p-n junction is created by doping, for example by ion implantation or diffusion of dopants. In semiconductor production, doping is the intentional introduction of impurities into an intrinsic semiconductor for the purpose of modulating its electrical properties. The doped material is referred to as extrinsic semiconductor.

Transistor

A transistor is a three terminal electronic device formed by the combination of two P-N junctions in specific manner. It is a main building block of all modern electronic systems. In communication system it is used as the primary component in the amplifier. In digital electronics it is used as a high speed electronic switch. Basically transistor is a crystal with three doped regions. They are called emitter, base and collector. Extreme regions are emitter and collector, the middle region is base.

Do you know?

The history of the transistor begins with the dramatic scientific discoveries of the 1800's scientists like Maxwell, Hertz, Faraday, and Edison made it possible to harness electricity for human uses. Inventors like Braun, Marconi, Fleming, and De Forest applied this knowledge in the development of useful electrical devices like radio.

All these three regions are provided with terminals which are labeled as E (for emitter), B (for base) and C (for collector) respectively. A brief description about each of them is given here:

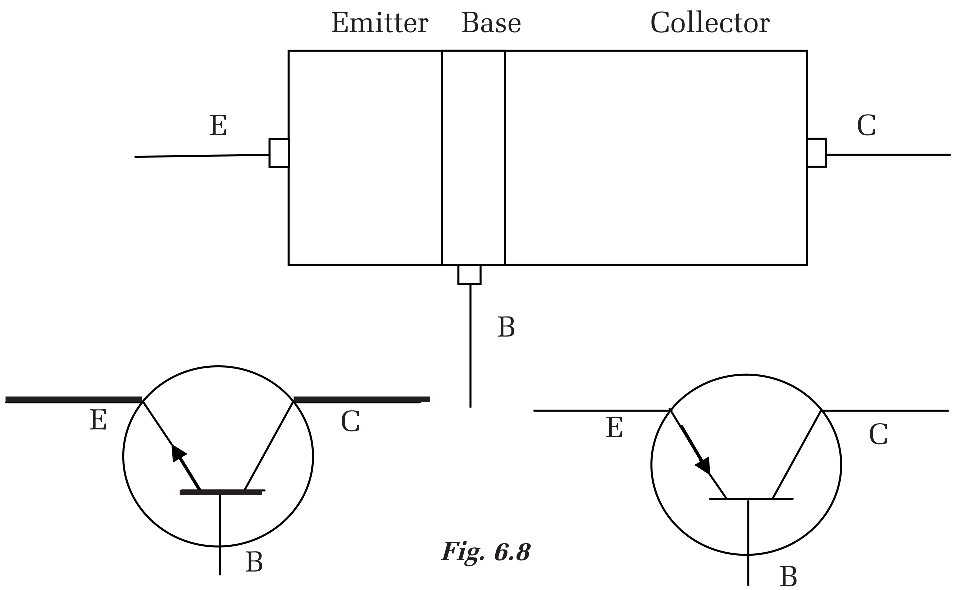


Fig. 6.8

Emitter: It is a region situated in one side of a transistor. It is heavily doped region. It is thinner than collector but much thicker than base. Its main function is to supply majority charge carriers to other regions. In circuit, it is denoted by E or e.

Base: It is middle region. It is thinnest region. It is lightly doped region. Its main function is to pass the majority charge carriers injected from the emitter on to the collector. The base is made thin and its dopping level is light to minimize the recombination of electrons and holes. In circuit, it is denoted by B or b.

Collector: It is situated on the other side of transistor. It is thickest region. It is moderately doped region. Its main function is to collect the majority charge carrier supplied by emitter and passed by the base. Collector is made thick to dissipate the heat produced in the collector circuit. In circuit, it is denoted by C or C_1.

Diode

A diode is a semiconductor device that essentially acts as a one-way switch for current. It allows current to flow easily in one direction, but severely restricts current from flowing in the opposite direction. It is also known as rectifier because it changes alternating current (ac) into pulsating direct current (dc). It is rated according to its type, voltage, and current capacity.

Fig. 6.9

Diodes have polarity, determined by an anode (positive lead) and cathode (negative lead). Most diodes allow current to flow only when positive voltage is applied to the anode. A variety of diode configurations are displayed in this graphic alongside:

Diodes are available in various configurations. From left: metal case, stud mount, plastic case with band, plastic case with chamfer and glass case. When a diode allows current to flow, it is forward-biased. When a diode is reverse-biased, it acts as an insulator and does not permit current to flow.

Do you know?

There are different kinds of diodes available. Some of the major types of diodes include light-emitting diodes (LEDs), Zener diodes, small signal diodes, power rectifiers, and photodiodes.

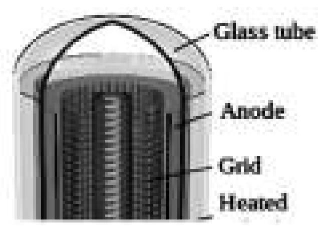


Fig. 6.10

The diode symbol's arrow points against the direction of electron flow. Reason: Engineers conceived the symbol, and their schematics show current flowing from the positive (+) side of the voltage source to the negative (-). It's the same convention used for semiconductor symbols that include arrows—the arrow points in the permitted direction of "conventional" flow, and against the permitted direction of electron flow.

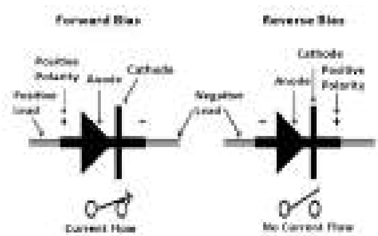


Fig. 6.11

A digital multimeter's Diode Test mode produces a small voltage between the test leads sufficient to forward-bias a diode junction. Normal voltage drop is 0.5 V to 0.8 V.

Generator (or Dynamo)

Generator can be defined as a device which converts mechanical energy (kinetic energy) into electrical energy. There are two types of generator. They are A.C. generator and D.C. generator. A.C. generator produces alternating current, which changes its polarity continuously and magnitude remains variable. A.C. dynamo is a small form of A.C. generator. D.C. generator produces direct current, which does not change its polarity and its magnitude remains constant.

Do you know?

Portable electric generators can be a good way to keep lights, refrigerators and other appliances running if a power outage occurs.

Principle of an A.C. generator

A.C. generator is based on the principle of electromagnetic induction i.e. whenever the magnetic flux (f) linked with a closed circuit (a coil) changes, an induced e.m.f. is produced in it. As a result electrical current flows in the coil.

Construction of an A.C. Generator

A.C. generator consists of a rectangular coil (a large number of turns of insulated copper wire) ABCD which can be rotated quickly between the poles N and S of a strong horse shoe type magnet. Two ends (A and D) of the rectangular coils are connected to slip rings R1 and R2 (two circular pieces of copper metal). When these two slip rings R1 and R2 are rotated with the coil, the carbon brushes, B1 and B2 (two pieces of carbon), keep contact with them. The current produced in the rotating coil can be tapped out through slip rings into the carbon brushes. From the carbon brushes, current is taken to various electrical appliances to run them through wires.

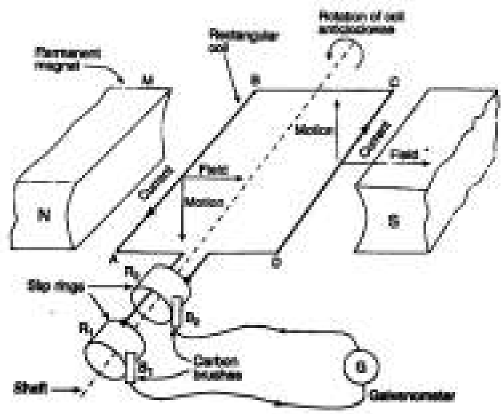


Fig. 6.12

Working principle of an A.C. generator

When coil ABCD is being rotated in the anticlockwise direction between the poles N and S of a horse shoe type magnet, the side AB of the coil moves down cutting the magnetic lines of force near the N-pole of the magnet, and side CD moves up, cutting the lines of force near the S-pole of the magnet. Due to this, induced current is produced in the sides AB and DC of the coil. On applying Fleming’s right-hand rule to the sides AB and DC of the coil, we find that the currents in the directions B to A and D to C. Thus, the induced current in the two sides of the coil are in the same direction, and we get an effective induced current in the direction BADC.

After half revolution, the reverse processes occur in their positions and functions as well. As a result, the direction of induced current in each side of the coil is reversed after half a revolution and the polarity of the two ends of the coil also changes after half revolution, i.e. positive end becomes negative and negative end becomes positive. Thus, in one revolution of the coil, the current changes its direction twice.

The alternating current (A.C.) supplied in Nepal has a frequency of 50 Hz, i.e. the coil is rotated at the rate of 50 revolutions per second. Since in one revolution of coil, the current changes its direction twice, so in 50 revolutions of coil, the current changes its direction one hundred times. Thus, the A.C. supply in Nepal changes its direction 100 times in 1 second.

Bicycle Dynamo

Bicycle dynamo is a device which converts mechanical energy into electrical energy.

Principle: Bicycle dynamo is based on the principle of electromagnetic induction i.e. whenever the magnetic flux (f) linked with a closed circuit (a coil) changes, an induced emf is produced in it. As a result, a current flows in the coil.

Construction and working principle : In a dynamo, a coil of insulated copper placed inside the magnetic field of a magnet as shown in the figure alongside. As the wheel of the bicycle rotates, it makes the wheel of the dynamo to spin causing the rotation of the magnet inside the dynamo. The magnetic lines of force turn as the magnet spins and the coil of the wire happens to cross the magnetic lines of the force. Because of this, current is induced in the coil. The coil of the wire is wound up in a soft magnetic material to increase the magnetic strength and current is induced.

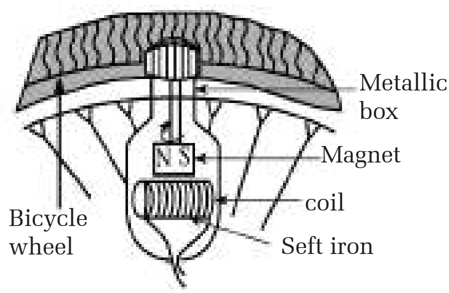


Fig. 6.13 Bicycle dynamo

Activity

Take a bicycle fitted with a dynamo. Get the upper part of the dynamo to touch the tyre. Support the bicycle in its stand and turns its wheel first slowly and then speed it up. Observe the brightness of light of the bicycle.

i. What changes occur when it is turned first slowly and then speeded up and why?

ii. What can be done to increase the magnitude of current?

D.C. generator

If the current always flows in the same direction, then it is called a direct current. Direct current generator (D.C. generator) produces direct current.

Construction of a D.C. generator

D.C. generator consists of a rectangular coil (a large number of turns of insulated copper wire) ABCD which can be rotated quickly between the poles N and S of a strong horse shoe type magnet M. Two ends (A and D) of the rectangular coils are connected to the two copper half rings or split rings R_1 and R_2 of a commutator (two circular pieces of copper metal). There are two carbon brushes B_1 and B_2 which press lightly against the two half rings. When the coil is rotated, the two half rings R_1 and R_2 touch the two carbon brushes B_1 and B_2 one by one. The current produced in the rotating coil can be tapped out through the commutator half rings into the carbon brushes. From the carbon brushes, current is taken through wire to various electrical appliances to run them.

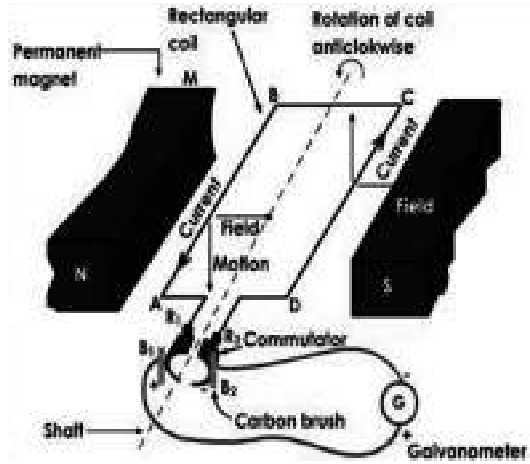


Fig. 6.14

Working principle of a D.C. generator

When a coil ABCD is being rotated in the anticlockwise direction between the poles N and S of a horse-shoe type magnet, the side AB of the coil moves down cutting the magnetic lines of force near the N-pole of the magnet, and side DC moves up, cutting the lines of force near the S-pole of the magnet. Due to this, induced current is produced in the sides AB and DC of the coil. On applying Fleming's right-hand rule to the sides AB and DC of the coil we find that the currents in them are in the directions B to A and D to C respectively. Thus, the induced currents in the two sides of the coil are in the same direction, and we get an effective induced current in the direction BADC. Due to this the brush B_1 becomes a positive (+) pole and B_2 becomes negative (-) pole of the generator.

After half revolution, the reverse processes occur in their positions and functions as well. When sides of coil interchange their positions, the two commutator half rings R_1 and R_2 automatically change their contacts from one carbon brush to another. Due to this change, the current keeps flowing in the same direction in the outer circuit. The brush B_1 always remains positive terminal and brush B_2 always remains negative terminal of the generator. Thus, a D.C. generator supplies a current in one direction by the use of a commutator consisting of two, half rings of copper.

The strength of the current generated by the generator or dynamo can be increased by the following methods:

a. By increasing the number of turns of wire in the coil

b. By increasing the speed of dynamo

c. By decreasing the distance between the coil and the magnet

d. By increasing the strength of magnetic field

Motor Effect

When an electric current is passed through a conductor, a magnetic field is developed around it. If the conductor is placed within a magnetic field, there is mutual attraction or repulsion due to the magnetic field developed between a current carrying conductor and magnetic field of the magnet. According to Newton's third law of motion, the magnet must also exert an equal and opposite force on the current carrying conductor. A French Scientist Andre Marie Ampere in 1820 experimentally demonstrated this force.

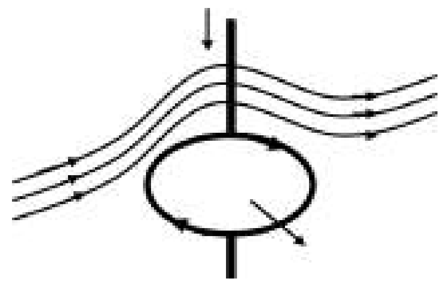


Fig. 6.15

The motion produced in a current carrying conductor kept to move freely in another magnetic field is called motor effect.

Electric motor: An electric motor is a rotating device which converts electrical energy into mechanical energy.

Principle of electric motor: It is based on the principle of motor effect, i.e., when a rectangular coil is placed in a magnetic field and current is passed through it, a torque acts on the coil which rotates it continuously.

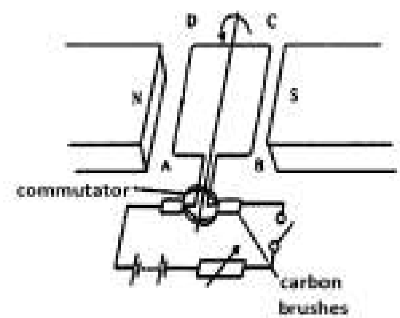


Fig. 6.16

Types of electric motor: Electric motor is of two types:

(a) **AC motor:** It is also called alternating current motor. It uses AC supply in the circuit. For example, motor of a fan, washing machine, etc.

(b) **DC motor:** It is also called direct current motor. It uses DC supply in the circuit. For example, motor of battery operated toys.

Construction and working principle

Electric motor consists of a shaft or spindle which rotates continuously when current is passed into it. When the shaft rotates, it drives various types of machines in homes and industry. A common electric motor works on direct current. That's why, it is called D.C.(direct current) motor.

Uses of electric motor

An electric motor is mainly used in domestic appliances like vacuum cleaners, electric fans, washing machines, refrigerators, mixer, grinder, computer, MP3 player, electric water pump, etc.

Comparison of D.C. generator with electric motor

It is concluded that the construction of a D.C. generator is similar to that of an electric motor. But its functioning is just opposite to that of a motor. In a D.C. generator, we rotate the coil and produce direct current whereas in an electric motor, we do apply direct current and rotate the coil. If a D.C. generator is connected to a battery, it will run as a motor. Conversely, if a motor coil is made to rotate, then it will behave as a D.C. generator, and produce direct current at the brushes.

The D.C. generator or D.C. dynamo produces a direct electric current by the use of two half rings of copper called commutator. If we connect slip rings instead of using two half rings of a commutator, then we will get an A.C. generator or A.C. dynamo.

Transformer

A transformer is a device used to convert low alternating voltage at high current into high alternating voltage at low current and vice versa. In other words, a transformer is an electrical device used to increase or decrease the magnitude of alternating voltage.

Construction of a transformer

The transformer consists of three important parts such as soft iron core, primary coil and secondary coil. The soft iron core is made by placing thin rectangular plates of soft iron one above the other. Each sheet is laminated by using varnish or shellac and held tightly together by clamps. Thus, it formed a rectangular frame called a laminated core. There are two coils which are not connected to one another in any way. These coils are wound on the iron core. One coil is supplied with alternating current and is called primary coil. The voltage supplied in primary coil is called primary voltage and another coil is also there, which is called secondary coil and output current comes out through this coil called secondary voltage.

> Do you know?
>
> An ideal transformer is a theoretical and linear transformer. It is lossless and perfectly coupled; i.e, there are no energy losses and flux is completely confined within the magnetic core.

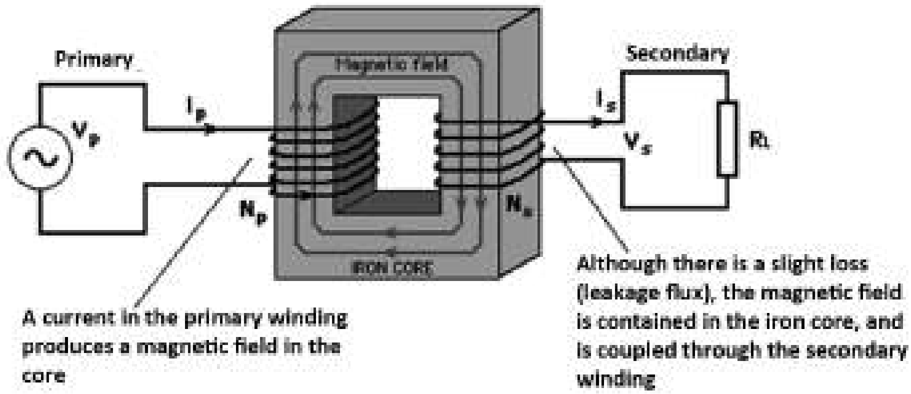


Fig. 6.17

Working Principle of Transformer

The working principle of a transformer is the phenomenon of mutual induction between two windings linked by common magnetic flux. The figure shows the simplest form of a transformer. Basically a transformer consists of two inductive coils; primary winding and secondary winding. The coils are electrically separated but magnetically linked to each other. When, primary winding is connected to a source of alternating voltage, alternating magnetic flux is produced around the winding. The core provides magnetic path for the flux, to get linked with the secondary winding. Most of the flux gets linked with the secondary winding which is called as 'useful flux' or main 'flux', and the flux which does not get linked with secondary winding is called as 'leakage flux'. As the flux produced, alternating emf gets induced in the secondary winding according to Faraday's law of electromagnetic induction. This emf is called 'mutually induced emf', and the frequency of mutually induced emf is same as that of supplied emf. If the secondary winding is closed circuit, then mutually induced current flows through it, and hence the electrical energy is transferred from one circuit (primary) to another circuit (secondary). The given formula shows the relation between the primary and secondary voltage:

$$\frac{\text{Number of turns in secondary coils } (N_2)}{\text{Number of turns in primary coils } (N_1)} = \frac{\text{Secondary voltages} (V_2)}{\text{Primary voltages } (V_1)}$$

i.e. $\frac{N_2}{N_1} = \frac{V_2}{V_1}$

Types of Transformer

Transformer can be classified on different basis such as types of construction of purpose, types of supply, types of cooling, etc.. Brief information about the various types of transformer is given below:

a. On the basis of construction 1. Core type transformer 2. Shell type transformer	b. On the basis of their purpose 1. Step up transformer 2. Step down transformer
c. On the basis of type of supply 1. Single phase transformer 2. Three phase transformer	d. On the basis of their use 1. Power transformer 2. Distribution transformer 3. Instrument transformer:
e. On the basis of cooling employed 1. Oil-filled self cooled type 2. Oil-filled water cooled type 3. Air blast type (air cooled)	

Among all, here, we will discuss the types of transformer on the basis of their purpose in detail.

a. Step up transformer: The transformer which converts low A.C. voltage at high current into high A.C. voltage at low current is called step up transformer. A step up transformer gives increased alternating output voltage. Step up transformer is shown in the figure.

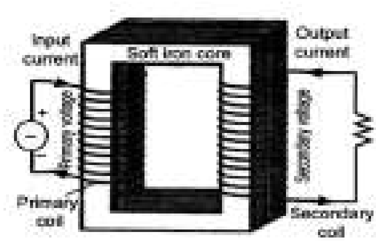


Fig. 6.18 Step up transformer

b. Step down transformer: The transformer which converts high A.C. voltage at low current into a low A.C. voltage at high current is called step down transformer. A step down transformer gives decreased alternative output voltage. Step down transformer is shown in the figure.

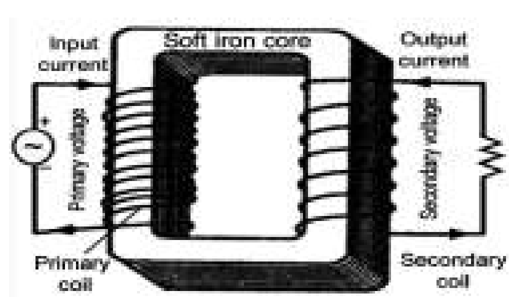


Fig. 6.19 Step down transformer

Differences between step up transformer and step down transformer

1. Step up transformer	1. Step down transformer
2. It consists of less number of primary coils in comparison to secondary coils.	2. It consists of more number of primary coils in comparison to secondary coils.
3. It converts low A.C. voltage at high current into high A.C. voltage at low current.	3. It converts high A.C. voltage at low current into a low A.C. voltage at high current.
4. It is mainly used in power plants.	4. It is mainly used in consumer's house.

Uses of a transformer

The transformers are used for various purposes in our daily life. They are used to regulate voltages while using different electronic gadgets such as computer, television, air conditioner, trolley buses, etc. They are used for doorbells, welding purpose and in electrical furnaces too. Depending upon the types, the step up transformer is used to transmit alternating current over long distances without much loss of current because it steps up the voltage. At sub-stations, the step down transformer is used for domestic and industrial purposes because it steps down the voltage.

Numerical Illustration

1. **The number of turns in primary and secondary coils of a transformer are 1600 and 400 respectively. If primary voltage is 220 V, then, calculate its secondary voltage.**

Solution:

Number of turns in primary coil (N_1) = 1200

Number of turns in secondary coil (N_2) = 400

Primary voltage (V_1) = 220 V

Secondary voltage (V_2) =?

We have,

$$\frac{N_2}{N_1}=\frac{V_2}{V_1} = \frac{400}{1600}=\frac{V_2}{220}$$

$$V_2=\frac{400\times 220}{1600}=55V$$

∴ The voltage across the secondary coil of the transformer is 55 V.

2. **A transformer of 1000 primary turns is connected to a 220 V A.C. supply. How many turns of secondary coil will be needed in order to generate 110 V from that transformer?**

Solution:

Primary voltage (V_1) = 220 V

Secondary voltage (V_2) = 110 V

Number of turns in primary coil (N_1) = 1000

Number of turns in secondary coil (N_2) =?

We have,

$$\frac{N_3}{N_1} = \frac{V_2}{V_2}$$

$$\frac{N_2}{1000} = \frac{110}{220}$$

$$N_2 = \frac{1000 \times 110}{220} = 500$$

∴ The number of turns in the secondary coil of a transformer is 500.

Summary

1. The substances which are attracted by a magnet are called magnetic substances.
2. Ferromagnetic substances, paramagnetic substances and diamagnetic substances are three magnetic substances.
3. The process of making a magnet using magnetic substances either by rubbing or passing electricity in them is called magnetization.
4. The numbers of lines force passing the surface held perpendicular to these lines of force is called magnetic flux through a surface.
5. The degree to which the lines of force can penetrate in a substance placed in magnetizing field is called the permeability of the substance.
6. The property of the magnetic substance due to which magnetization of it takes place to different extents is called susceptibility.

7. The process of production of electric current by moving a straight current conductor in a magnetic field is called electromagnetic induction.

8. The following are Faraday's three laws of electromagnetic induction:

   (a) Whenever the magnetic flux (f) linked with a closed circuit (a coil) changes, an induced emf is produced in it.

   (b) The induced emf lasts until the change in magnetic flux continues.

   (c) The magnitude of the induced emf is directly proportional to the rate of change of magnetic flux within the closed coil.

9. The direction of the induced current produced in a straight conductor moving in a magnetic field can be predicted by using Fleming's Right Hand Rule.

10. The device which regulates current in an electric circuit is called resistor. It is also called electric load. Resistors can be connected together in the circuit in following three ways:

    Series connection b. Parallel connection c. Mixed connection

11. When a number of resistors in a circuit are connected from end to end in order to get same current flowing through each of them in succession, the resistors are said to be connected in series connection.

12. When a number of resistors in a circuit are connected between two common points in order to get different current following through each of them , the resistors are said to be connected in parallel connection.

13. A transistor is a device that controls the movement of electrons and consequently, electricity in an electrical circuit.

14. A diode is a semiconductor device that essentially acts as a one-way switch for current.

15. Generator can be defined as a device which converts mechanical energy (kinetic energy) into alternating form of electrical energy.

16. Bicycle dynamo is a device which converts mechanical energy into electrical energy.

17. The strength of the current generated by the generator or dynamo can be increased by the following methods:

    (a) By increasing the number of turns of wire in the coil

    (b) By increasing the speed of dynamo

    (c) By decreasing the distance between the coil and the magnet

(d) By increasing the strength of magnetic field

18. The motion produced in a current carrying conductor placed when kept to move freely in another magnetic field is called motor effect.
19. A transformer is a device used to convert low alternating voltage at high current into high alternating voltage at low current and vice versa.
20. The transformer which converts low A.C. voltage at high current into high A.C. voltage at low current is called step up transformer. A step up transformer gives increased alternating output voltage.
21. The transformer which converts high A.C. voltage at low current into a low A.C. voltage at high current is called step down transformer. A step down transformer gives decreased alternative output voltage.

Exercise

A. Tick (√) the best alternative from the following:

1. The magnetic substances which are highly attracted by the magnet are called:

   i. Paramagnetic substances　　ii. Ferromagnetic substances

   iii. Diamagnetic substances　　iv. Non magnetic substances

2. The process of production of electric current by moving a straight current carrying conductor in a magnetic field is called:

   i. Electromagnetic induction　　ii. Magnetization

   iii. Demagnetization　　iv. Motor effect

3. How are resistors connected together in the circuit?

   i. Series connection　　ii. Parallel connection

   iii. Mixed connection　　iv. All of the above

4. Bicycle dynamo is a device which converts:

   i. Electrical energy into mechanical energy

   ii. Mechanical energy into electrical energy

   iii. Magnetic energy into mechanical energy

   iv. Electrical energy into magnetic energy

5. The step up transformer is a device, which converts:

    i. low A.C. voltage at high current into high A.C. voltage at low current

    ii. high A.C. voltage at low current into a low A.C. voltage at high current

    iii. high A.C. voltage at high current into low A.C. voltage at low current

    iv. low A.C. voltage at low current into high A.C. voltage at high current

B. Give short answers to the following questions:

1. What are magnetic substances?
2. What is meant by magnetization?
3. Define magnetic flux through a surface.
4. Define a permeability and susceptibility of a substance.
5. What is relative permeability?
6. What is meant by electromagnetic induction?
7. What do three fingers in Fleming's Right Hand Rule indicate?
8. What is resistor?
9. By how many ways resistors can be connected together in the circuit?
10. Define series and parallel connection of resistors.
11. Write a short note on transistor.
12. What is a generator?
13. Define bicycle dynamo.
14. What is motor effect?
15. Define a transformer and mention its types.

C. Give long answers to the following questions:

1. Explain various types of magnetic substances with examples.
2. State Faraday's three laws of electromagnetic induction.
3. Explain an experiment to demonstrate electromagnetic induction.
4. How are resistors connected in the circuit? Explain.
5. Mention the characteristics of series and parallel connection of resistors.
6. Describe the construction and working principle of an A.C. generator.
7. How doe as bicycle dynamo work? Explain.
8. Describe the construction and working principle of a D.C. generator.

9. Explain the types and working principle of an electric motor.
10. Explain the construction and working principle of transformer.
11. What are various types of transformer? Mention a chart of it.
12. Differences between step up transformer and step down transformer.
13. Write the uses of a transformer.

D. Solve the following Numerical Problems

1. The number of turns in the primary and secondary coils of a transformer are 1200 and 300 respectively. If secondary voltage is 55 V, then, calculate its primary voltage. *(Ans: 220 V)*
2. A transformer of 2000 primary turns is connected to a 220 V A.C. supply. How many turns of secondary coil will be needed to generate 110 V ? *(Ans: 1000)*
3. A transformer has 240 input voltages, 30 A inputs current and 120 output voltages, then calculate out current. *(Ans: 60 A)*
4. The number of turns in the primary winding of a certain transformer is 150 times more than that in the secondary winding. Calculate the input emf in the primary winding if the emf generated in the secondary winding is 220 V AC. *(Ans: 33000V)*

Project Work

1. Open a dynamo of a bicycle and observe it carefully. Compare the parts contained inside it and the structure of the dynamo given in this unit. On the basis of this observation, prepare a report and present it to the class. (Note: if bicycle is available)
2. Make a visit to a nearby electric motar reparing centre and request a machine to explain about the parts of a motor. Prepare a report on it.

Glossary

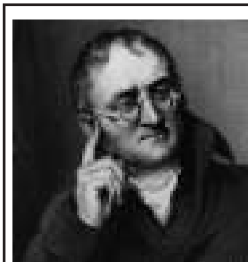
Missile : an object or weapon for throwing, hurling, or shooting, as a stone, bullet

Semiconductor device: a solid chemical element or compound that can conduct electricity under some conditions

Alternating : The direction of current that is continuously changing

Unit 7

Atomic Structure


John Dalton (1766-1844)

John Dalton born on September 6, 1766, in Eaglesfield, England died on July 26, 1844 in Manchester, England. He is well known for the modern atomic theory. He was also the first scientist to study color blindness. In 1803 he proposed the concept of Dalton's Law of Partial Pressures.

7.1 Introduction

Atoms are the smallest particles of matter. They can neither be created nor be destroyed. They take in chemical reaction without division. They are the building blocks of matter. They are same as bricks of a building. The word atom is derived from the word "*atomus*" meaning indivisible particle. An atom is made up of chiefly three types of small particles called sub-atomic particles. They are protons, neutrons and electrons. They are the main particles of the atom. So, they are also called fundamental particles of an atom. About 99% of an atom is made up of nothing but empty space. The remaining 1% is made up of these sub-atomic particles.

7.2 Atomic mass or Atomic weight

The atomic mass is the calculation of the total mass of an atom. To calculate the total mass of an atom, the mass of the protons, neutrons and electrons should be added. The proton and neutron are of roughly equal sized while the electron is about 1837 times smaller than a proton. Hence, the mass of an electron can be neglected while calculating the total mass of an atom.

The mass and weight are different things. Mass is the total amount of matter present in a body while weight is the amount of force applied by the earth to pull the body. For the bigger objects mass and weight have different values, but for the smaller particles like the sub-atomic particles of an atom, the mass and weight are nearly the same.

The total mass of protons and neutrons which are present in the nucleus of an atom is called atomic mass. The mass of a proton or a neutron is nearly equal to

1.67 x 10^{-27} kg. This mass is equal to $1/12^{th}$ mass of a carbon-12 isotope. This is also called atomic mass unit (amu). Hence, **amu is a unit which is used to measure the mass of protons, neutrons, electrons or atoms. It is equal to the mass of $1/12^{th}$ mass of a carbon-12 isotope.** The mass of one proton or neutron is equal to one amu. Therefore, we count the total number of protons and neutrons of an atom and calculate the total mass of the atom.

Let us consider a helium atom. In a helium atom, there are two protons and two neutrons. The total number of protons and neutrons is 4. Therefore, the mass of a helium atom is 4 amu.

Likewise, atomic mass of Hydrogen = p + n = 1+0 = 1 amu

Atomic mass of Sodium = p + n = 11 + 12 = 23 amu

7.3 Molecular mass or Molecular weight

An atom and a molecule are the same in case of mono atomicmolecules like helium, neon, argon, krypton, xenon, radon, etc. An atom is always single while molecules can be single as well as multiple of atoms. An atom may or may not exist independently but a molecule exists independently. For example:

a. The atoms of helium, neon, argon, krypton, xenon, radon, etc. can exist in dependently. So, atoms of these elements are also called molecule.
b. An atom of oxygen (O) cannot exist independently. It combines with another such atom and forms molecule (O_2).
c. Similarly, the smallest units of all compounds are called molecules. They can exist independently. For example, molecule of sodium chloride (NaCl), molecule of water (H_2O), etc.

The molecular mass of a compound is the mass of total number of atoms present in it. Remember that the mass of an atom can be calculated by adding the number of protons and neutrons. This concept can also be used to calculate the molecular mass. Let us consider an oxygen molecule (O_2). The oxygen molecule has two atoms in it. Therefore, its total mass is equal to the mass of two oxygen atoms, i.e. O_2 = 2 × mass of O = 2 × 16 amu = 32 amu. Likewise, mass of sulphuric acid (H_2SO_4) is equal to 2 × mass of hydrogen atom + 1 × mass of sulphur atom + 4 × mass of oxygen atom = 2×1 + 1×32 + 4×16 = 98 amu. The overall discussion concludes that, **molecular mass is the sum of atomic mass of all atoms present in a molecule**. It is equal to the mass of $1/12^{th}$ mass of a carbon-12 isotope. It means that molecular mass of oxygen is 32 times heavier than the $1/12^{th}$ mass of a carbon-12 isotope. Similarly, molecular mass of sulphuric acid is 98 times heavier than the $1/12^{th}$ mass of a carbon-12 isotope.

Example: 1 Calculate molecular mass of calcium carbonate ($CaCO_3$).

Molecular mass of $CaCO_3$ = 1×Ca + 1×C + 3×O

= 1×40 + 1×12 + 3×16 = 100 amu

Example: 2 Calculate molecular mass of Ammonium carbonate $(NH_4)_2CO_3$

Molecular mass of $(NH_4)_2CO_3$ $= 2\times N + 8\times H + 1\times C + 3\times O$

$= 1\times 14 + 8\times 1 + 1\times 12 + 3\times 16$

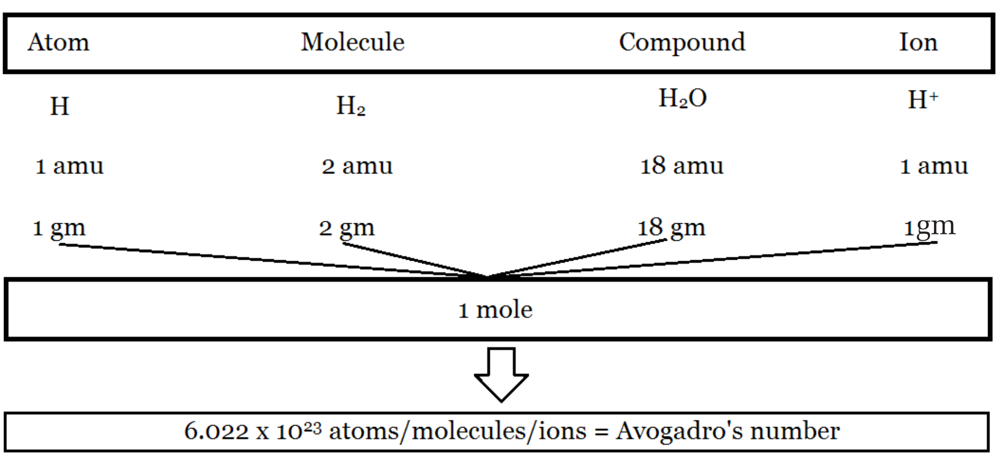
$= 28 + 8 + 12 + 48$

$= 96$ amu

7.4 Mole Concept

In any chemical reaction, fixed numbers of atoms or molecules or ions react to give the definite products. It is practically very difficult to perform a chemical reaction with definite number of atoms or molecules. This is because a small quantity of mater contains a large number of particles or atoms or ions. The mole is the standard quantity in chemistry. It is used to measure how much of a substance is present in a sample of substance. One mole of a substance contains as many atoms or molecules or ions as there are atoms in 0.012 kilogram or 12 grams of carbon-12 isotope. It is found that in 0.012 kilogram or 12 grams of carbon-12 isotope there are 6.022×10^{23} atoms. It is called one mole of carbon. Thus, mole is the collection of particles or units (atoms, ions, molecules or even sub-atomic particles).**The amount of substance which contains 6.022×10^{23} atoms, ions, molecules or even sub-atomic particles is called one mole.**

To understand mole concept clearly, let us see the following diagram:



Let us consider an atom of hydrogen, a molecule of hydrogen, a molecule of water and an ion of hydrogen. The weight of these things is 1 amu, 2 amu, 18 amu and 1 amu, respectively. Now, if we take one-gram molecular weight of these substances then we have 1 gm of hydrogen atom, 2 grams of hydrogen molecule, 18 gm of water and 1 gm of hydrogen ion. The amount of matter present in these substances is called 1 mol., i.e.

a. 1 gram of hydrogen atom is 1 mole of hydrogen atom

b. 2 grams of hydrogen molecule is 1 mole of hydrogen molecule

c. 18 grams of water molecule is 1 mole of water

d. 1 gram of hydrogen ion is 1 mole of hydrogen ion

The total amount of mass present in one mole of a pure substance is called molar mass. Its SI unit is gm/mol.

7.5 Avogadro's Number

The number which is equal to 6.022×10^{23} is called Avogadro's number. It is the number of atoms, molecules or ions present in one mol of a pure substance. The word Avogadro is named after an Italian chemist Amedeo Avogadro. He discovered this concept. The concept is 1 mol of a pure substance contains exactly equal number of atoms or molecules or ions which is numerically 6.022×10^{23}. This number 6.022×10^{23} is called Avogadro's number.

Number of atoms in 1 gm of hydrogen atom = Number of molecules in 2 gm of hydrogen molecule = Number of molecules in 18 gm of water = Number of ions in 1 gm of hydrogen ion

Example 1:

Calculate the molar mass of $NaNO_3$ and find the number of molecules present in 2 mole of $NaNO_3$.

Solution:

Molecular mass of $NaNO_3 = 23 + 14 + 3 \times 16 = 85$ amu

Number of molecules in 1 mole of $NaNO_3$ = Avogadro's number = 6.022×10^{23} molecules

Hence, number of molecules in 2 mole of $NaNO_3 = 2 \times 6.022 \text{ x } 10^{23} = 12.044 \text{ x } 10^{23}$ molecules

Example 2:

Calculate the number of molecules present in 1 kg of water.

Solution:

1 kg of water = 1000 gm

As we know, 1 mole of water = 18 gm

$= 6.022 \times 10^{23}$ molecules

$$1 \text{ gm of water} = \frac{6.022 \times 10^{23}}{18} \text{ molecules}$$

So, 1000 gm of water = $\frac{6.022\times10^{23}}{18} \times 1000$ molecules

$= 3.34\times 10^{25}$ molecules

Note: Number of molecules = Total mass (m)× Avogadrao's number/ Molar mass(M)

7.6 Quantum number

According to Bohr's model, an electron lies at a fixed distance from the nucleus. From this distance, electron revolves around the nucleus. But, this concept is proved to be false. The quantum mechanical model of an atom rejected the Bohr's model. This model of an atom is the successor of Bohr's model. This model states that **an electron can be anywhere within a certain region around an atom and its position cannot be accurately predicted**. The actual position and velocity of an electron cannot be calculated at the same time. The region around the nucleus where there is high probability of finding the electrons is called orbital. Orbitals are of four major types. They are s-orbital, p-orbital, d-orbital and f-orbital. But, only this information is not sufficient to study about the position of electrons. Electrons have certain energy. They are located at a certain position and they spin about their own axis in a particular direction. These properties of electrons are studied by a particular set of numbers called quantum numbers. Thus, **the characteristics of the electrons and their orbitals are determined by a specific number called quantum number**. Quantum numbers determine the position, angular momentum, orientation and spinning property of an electron. There are altogether four quantum numbers:

1. Principal quantum number (n)
2. Azimuthal or Angular momentum quantum number (l)
3. Magnetic quantum number (m)
4. Spin quantum number (s)

The principal quantum number (n)

The principal quantum number describes the size of the orbital and average distance from the nucleus. They have values in numbers like 1,2,3, etc. It is also called energy level or shell number. The principal quantum number n=2 is at a greater distance from the nucleus than the number n=1. Since, n=2 is at a greater distance, it forms a larger region of sphere than n=1. Hence, the size of shell having higher value of quantum number is larger than the shell having lower values.

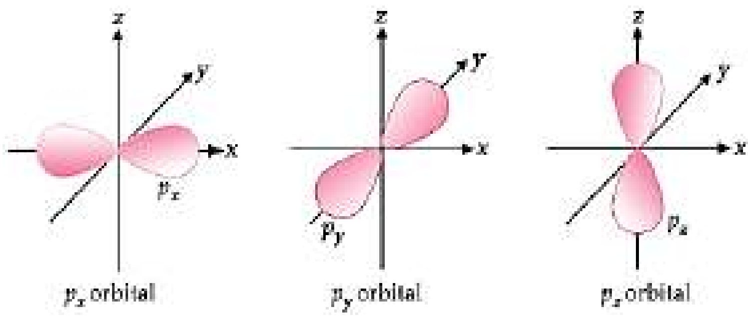
The azimuthal quantum number (l)

The azimuthal quantum number is also called angular momentum. It is because it describes the angular momentum of an electron. It also describes the shape of the orbital. It is symbolized by the letter l. The value for azimuthal quantum number can be 0,1,2,3……. (n-1). The azimuthal quantum number is upto one less than the principal quantum number. For example,

For n=1, l = 0. The value l= 0 represents s-orbital, which is spherical in shape.

For n=2, l=0,1. The value l=0 represents a spherical s-orbital and l=1 represents a dumb-bell shaped p-orbital. Similarly, for n=3, l=0,1,2 and so on.

Magnetic quantum number (m)



The azimuthal quantum describes the shape of an orbital. But, the same shape of the orbital can have different orientations. The orbital having one kind of orientation has different effects on the characteristics of an atom compared to when arranged in another. This effect of different orientations of orbitals was firstly observed in the presence of a magnetic field. Therefore these numbers are called magnetic quantum number. **Magnetic quantum number is the third set of quantum number which describes the sub-shell and their orientation**. It is denoted by the letter m or m_l.

The magnetic quantum numbers can have the values from –l to +l.

For example, when l=0, m=0

When l=1, m= -1,0,+1.

Spin quantum number (m_s)

One of the important properties of a pair of electrons in the orbital is that they not only revolve around the nucleus but also spin about their own axis. They spin in two ways. They are clockwise and anticlockwise. The clockwise spin is called up spin and is given the value of m_s= +1/2. Similarly, the anticlockwise spin is called a down spin and is given the value m_s=-1/2.

Allowed values of quantum numbers

a. The three quantum numbers (n, l, and m) that describe an orbital are integers: 0, 1, 2, 3, and so on.

b. The principal quantum number (n) cannot be zero. The allowed values of n are therefore 1, 2, 3, 4, and so on.

c. The angular quantum number (l) can be any integer between 0 and n - 1. For example, if n = 3, l can be either 0, 1, or 2.

d. The magnetic quantum number (m) can be any integer between -l and +l. If l = 2, m can be either -2, -1, 0, +1, or +2.

e. The spin quantum number (m_s or s) can have two values +1/2 or -1/2

Exercise 1 : Write down the value of quantum number for the last electron in the electronic configuration $2s^1$.

Ans : For $2s^1$

1. Principal quantum number (n) = 2
2. Azimuthal or Angular momentum quantum number (l) = 0 (Because it is s-sub shell)
3. Magnetic quantum number (m) = 0 $\begin{array}{c} 0 \\ \boxed{1} \end{array}$
4. Spin quantum number (s) = +1/2(Because it is up spin)

Exercise 2 : What are the quantum numbers for the last electron of the chlorine atom.

Ans : For $1s^2$, $2s^2$ $2p^6$,$3s^2$ $3p^5$

1. Principal quantum number (n) = 3
2. Azimuthal or Angular momentum quantum number (l) = 1 (Because it is p-sub shell)
3. Magnetic quantum number (m) = 0 $\begin{array}{|c|c|c|} \multicolumn{1}{c}{1} & \multicolumn{1}{c}{0} & \multicolumn{1}{c}{-1} \\ \hline 1\downharpoonright & 1\downharpoonright & 1 \\ \hline \end{array}$
4. Spin quantum number (s) = -1/2 (Because it is down spin)

7.7 Calculating Concentration of the Solution

1. **Molarity**

Molarity (M) is defined as the number of moles of a solute present in per litre of solution.

i.e. $\text{Molarity} = \dfrac{\text{Number of moles of solute}}{\text{Volume of solution in liter}}$

$\text{Molarity} = \dfrac{\text{Number of moles of solute}}{\text{Volume of solution in mL}} \times 1000$

Since, the number of moles of any substance can be obtained from its mass and molar mass. So,

$$\text{Molarity} = \frac{\text{Mass of solute}}{\text{Molar mass}} \times \frac{1000}{\text{Volume of solution in mL)}}$$

Molar solution: The solution containing one mole of solute in one litre solution is called molar solution. For example, if one mole of nitric acid (63 gram) is present in one litre of solution, it is called molar solution.

Decimolar solution: The solution containing 1/10 mole of a solute in one litre solution is called deci molar solution. For example, if 1/10 mole of nitric acid (6.3 gram) is present in one litre of solution, it is called decimolar solution.

2. **Molality**

Molality (m) is defined as the number of moles of solute present in one kilogram of solvent.

$$\text{i.e. Molality} = \frac{\text{Number of moles of solute}}{\text{Mass of of solvent in kilogram}}$$

$$\text{i.e. Molality} = \frac{\text{Number of moles of solute}}{\text{Mass of of solvent in gram}} \times 1000$$

$$\text{Molality} = \frac{\text{Mass of solute in gram}}{\text{Molar mass of solute gram}} \times \frac{1000}{\text{Mass of of solvent in gram}}$$

The molality of the solution does not change with temperature.

If 36.5 gram (1 mole) of hydrochloric acid is dissolved in 1000 gram of water, then it is called molal solution.

Molarity and Molality are completely different physical quantities. Molarity is measured in mol/litre of solution while molality is measured in mol/kg of solvent.

Note: For water as a solvent, molarity and molality have nearly same values. It is because 1 litre of water nearly weighs 1 kg. But, for other solvents like CCl_4, oil, etc., the values are different.

Example: 1 Calculate the molarity of 20g H_2SO_4 in 100 ml solution.

Solution:

Here, Given, mass of solute = 20g

Volume of solution = 100 ml

We know that molar mass of H_2SO_4 = 98 g

Again, number of moles of H_2SO_4 = $\frac{\text{Mass of solute}}{\text{Molecular weight of solute}} = \frac{20g}{98g} = \frac{20}{98}$

Also, volume of solution in litres = $\frac{100}{1000} = \frac{1}{10}$

Now, Molarity (M) = $\frac{\text{moles of solute}}{\text{litres of solution}} = \frac{20/98}{1/10} = 2.04M$

Example: 2 Calculate the molality if 45 g of NaCl is added into 900 gm of water.

Solution:

Here, Given, mass of solute = 45g

Mass of solution = 900 g

We know that molar mass of NaCl = 58 g (approx)

Number of moles of NaCl = $\frac{\text{Mass of solute}}{\text{Molecular weight of solute}} = \frac{45g}{58g}$

Mass of solvent in kg = $\frac{900}{1000} = \frac{9}{10}$

Now, Molality (m) = $\frac{\text{moles of solute}}{\text{mass of solvent in kg}} = \frac{45/58}{9/10} = 0.6M$

3. **Normality**

The number of equivalents of solute present in onelitre of solution is called normality (N).

i.e. normality is the number of mole equivalents in per litre of solution.

$$\text{Normality (N)} = \frac{\text{Number of equivalents of solute}}{\text{Volume of solution in litre}}$$

$$\text{Normality (N)} = \frac{\text{Number of equivalents of solute}}{\text{Volume of solution in mL}} \times 1000$$

$$\text{Normality (N)} = \frac{\text{Mass of solute in gram}}{\text{Equivalent mass}} \times \frac{1000}{\text{Volume of solution in mL}}$$

Normal solution (1 N): the solution containing one equivalent of solute in one litre of solution is called normal solution. For example, if 36.5 g of hydrochloric acid (one equivalent) is present in one litre of solution, then it is called normal solution of HCl.

Decinormal solution (N/10): the solution containing 1/10 equivalent of solute in one litre of solution is called decinormalsolution. For example, if 3.65 g of hydrochloric acid (N/10 equivalent) is present in one litre of solution, then it is called decinormal solution of HCl.

Normality is specifically used for acids and bases. As we can see above, to calculate normality, we need to find out the mole equivalent of a solution. The mole equivalent of an acid and a base can be calculated by using the given formula.

$$\text{Equivalent mass of acid} = \frac{\text{Molar mass of acid}}{\text{Basicity}}$$ (Here, basicity means number of replaceable hydrogens)

For example,

i. In HCl, there is one replaceable H^+ ion.Therefore, basicity of HCl is 1.

ii. In H_2SO_4, there are 2 replaceable H^+ ions.Therefore the basicity of H_2SO_4 is 2.

$$\text{Equivalent mass of base} = \frac{\text{Molar mass of base}}{\text{Acidity}}$$ (Here, acidity means number of replaceable hydroxyl ions)

For example,

i. In NaOH, there is one replaceable OH^- ion. Therefore acidity of NaOH is 1.

ii. In $Al(OH)_3$, there are 3 replaceable OH^- ions. Therefore acidity of $Al(OH)_3$ is 3.

To calculate equivalent weight of salt, the total charge in cation or anion is used in place of acidity and basicity. For example, $AlCl_3 = Al^{3+} + 3Cl^-$, in $AlCl_3$, the total charge in aluminium ion is 3. Hence, it is used to calculate equivalent weight.

If the molarity of the solution is given, the normality can be calculated by multiplying with acidity or basicity.

Normality = n × Molarity (where n is acidity or basicity)

For example:

a. For an acidic solution,

The 4 molar H_2SO_4 solution is the same as 8 normal H_2SO_4 solution.

b. For a basic solution,

The 2 molar $Ca(OH)_2$ solution is the same as 4 normal $Ca(OH)_2$ solution.

Note: The normality of a solution is equal to its molarity, if acidity or basicity is one.

Numerical Example:

1. Calculate the normality of the following:

a. 2 M of H_2SO_4

b. 0.0345 g of sodium carbonate in 300 mL of solution

Solution:

a. Given that, Molarity of the solution (M) = 2

n factor of H_2SO_4 = 2

As we know, Normality (N) = n × M = 2 x 2 =4

b. Given that, mass of solute = 0.0345 g

n factor for sodium carbonate = 2

No. of gram equivalent = moles × n factor = $\frac{0.0345}{106} \times 2 = 0.00065$

Now, Normality (N) = $\frac{\text{no. of mole equivalent}}{\text{1L of solution}} = \frac{0.00065}{\frac{300}{1000}\text{L}} = 0.0021\text{N}$

7.8 Grams per litre

Gram per litre is the mass of solute in a volume of solution expressed in litre. A gram per litre is often used to describe the concentration of a solid in a solution. It is represented by g/L.

Gram per liter (g/ L) = $\frac{\text{Mass of solute in gram}}{\text{Volume of solution in liter}}$

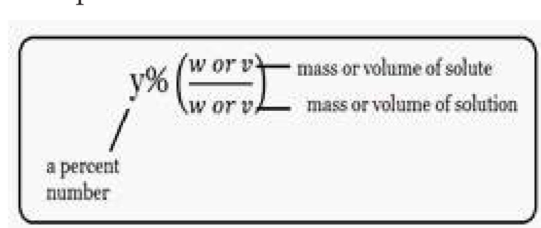
Gram per liter (g/ L) = $\frac{\text{Mass of solute in gram}}{\text{Volume of solution in mL}} \times 1000$

5 g/L sodium carbonate indicates that 5 gram sodium carbonate is present in 1000 mL of solution.

Percent composition

The percentage composition is the expression of mass of solution in 100 parts of volume of the solution. The percentage composition is written as:

The three forms of percent composition can be explained by the following examples.



a. $y\%\left(\frac{w}{w}\right)$ means yg of solute in 100 g of solution (w means weight, but often is mass).

b. $y\%\left(\frac{w}{v}\right)$ means yg of solute in 100 ml of solution (v means volume).

c. $y\%\left(\frac{v}{v}\right)$ means yml of solute in 100 ml of solution.

(i) **Percentage by mass (% w/w):** It represents the mass of the substance present in 100 g of the solution.

$$\boxed{\text{Percentage by mass (\% w/w)} = \frac{\text{Mass of solute in gram}}{\text{Mass of solution in g}} \times 100}$$

For example, 25% of glucose solution means that 25 g glucose is present in 100 g of the solution.

(ii) **Percentage by volume (% w/v):** It represents the mass of the substance present in 100 mL of the solution.

$$\boxed{\text{Percentage by volume (\% w/v)} = \frac{\text{Mass of solute in gram}}{\text{Volume of solution in mL}} \times 100}$$

For example, 15% of glucose solution means that 15 g glucose is present in 100 mL of the solution.

(iii) **Percentage by volume (% v/v) :** It represents the volume of the substance present in 100 mL of the solution.

$$\boxed{\text{Percentage by volume (\% v/v)} = \frac{\text{Volume of solute in mL}}{\text{Volume of solution in mL}} \times 100}$$

For example, 10% of HCl solution means that 10mL HCl is present in 100 mL of the solution.

Summary

1. Atoms are the smallest particles of matter. They can neither be created nor be destroyed. They take in chemical reaction without division.
2. The total mass of protons and neutrons which are present in the nucleus of an atom is called atomic mass.
3. amu is a unit which is used to measure the mass of protons, neutrons, electrons or atoms. It is equal to the mass of $1/12^{th}$ mass of a carbon-12 isotope.

4. Molecular mass is the sum of atomic mass of all atoms present in a molecule.
5. The amount of substance which contains 6.022×10^{23} atoms, ions, molecules or even sub-atomic particles is called one mole.
6. The total amount of mass present in one mole of a pure substance is called molar mass. Its SI unit is gm/mol.
7. The number which is equal to 6.022×10^{23} is called Avogadro's number. It is the number of atoms, molecules or ions present in one mol of a pure substance.
8. An electron can be anywhere within a certain region around an atom and its position cannot be accurately predicted.
9. The characteristics of the electrons and their orbitals are determined by a specific number called quantum number.
10. The principal quantum number describes the size of the orbital and average distance from the nucleus.
11. The azimuthal quantum number is also called angular momentum. It is because it describes the angular momentum of an electron. It also describes the shape of the orbital.
12. Magnetic quantum number is the third set of quantum number which describes the sub-shell and their orientation.
13. Molarity (M) is defined as the number of moles of solute present in per litre of solution.
14. The solution containing one mole of solute in one litre solution is called molar solution.
15. The solution containing 1/10 mole of a solute in one litre solution is called decimolar solution.
16. Molality (m) is defined as the number of moles of solute present in one kilogram of solvent.
17. The number of equivalents of solute present in one litre of solution is called normality (N).
18. The solution containing one equivalent of solute in one litre of solution is called normal solution.
19. Gram per litre is the mass of solute in a volume of solution expressed in litre.
20. The percentage composition is the expression of mass of solution in 100 parts of volume of the solution.

Exercise

A. **Tick (√) the best alternative from the followings.**

1. The Avogadro's number is :

   i. 6.022×10^{23} atoms   ii. 6.022×10^{23} molecules

   iii. 6.022×10^{23} ions   iv. All of above a, b and c

2. The principal quantum number is indicated by

   i. n   ii. l   iii. m   iv. s

3. What is value of "l" for n=1 ?

   i. 1   ii. 2   iii. 0   iv. 3

4. What is the normality for one molar HCl?

   i. 1   ii. 2   iii. 3   iv. 4

5. What does it mean by N/10?

   i. Normal solution   ii. Decinormal solution

   iii. Molar solution   iv. Decimolar solution

B. **Answer the following short questions.**

1. What is an atom?
2. Write down any two characteristics of an atom.
3. Define atomic mass.
4. What is molecular mass?
5. How many molecules are there in one mole of water?
6. State Avogadro's law.
7. Write down Avogadro's number.
8. What are quantum numbers?
9. Write the name of four quantum numbers.
10. What information does principal quantum number give?
11. Define normality.
12. What is molarity?

C. **Answer the following long questions.**

1. Write down the relation of atoms, ions, molecules with Avogrado's number.
2. Define molarity and molality and write their formula.
3. Define normality and write its formula.

4. Describe the principal quantum number.
5. What is azimuthal quantum number? Write down the value of “l” for n =2.
6. Describe the information which are obtained from the magnetic quantum number.
7. Explain in short about the spin quantum number.

D. **Numerical Problems**

1. Calculate the molecular mass of sulphuric acid.
2. Calculate the number of molecules in 36 g water.
3. What is the molality of a solution prepared by dissolving 5g of toluene (C_7H_8) in 225 grams of Benzene (C_6H_6)?
4. Calculate the molarity of 40 g H_2SO_4 in 100 ml solution.
5. Calculate the molality if 90 g of NaCl is added to 450 g of water.
6. Calculate the normality of a 4.0 molar sulphuric acid solution.

Project work

To prepare 1 molar solution and 1 molal solution:

1. **To prepare 1 Molar solution of sodium chloride (NaCl):**

Steps:

a. Take a molar mass of NaCl, i.e. 58 g of NaCl.
b. Take one litre volumetric flask and add the NaCl crystals into the flask.
c. Add some water and stir it to make a solution.
d. Fill the volumetric flask to one litre.
e. You have now a one molar (1 M) salt solution.

2. **To prepare 1 molal solution of sodium chloride (NaCl):**

Steps:

a. Take a molar mass of NaCl, i.e. 58 g of NaCl.
b. Take a beaker and add exactly one kilogram of water to it.
c. Mix the NaCl crystals in the water, and stir it to make a solution.
d. Now, you have a 1 molal solution of NaCl in water.

Question to think: What is the main difference while making molar and molal solution?

Glossary

Quantum number	:	the number which are used to find out the location of electron in an atom
Molarity (M)	:	the number of moles of solute present in per litre of solution
Molar solution	:	the solution containing one mole of solute in one litre solution
Decimolar solution	:	the solution containing 1/10 mole of a solute in one litre solution
Molality (m)	:	the number of moles of solute present in one kilogram of solvent
Normality (N)	:	the number of equivalents of solute present in one litre of solution
Normal solution	:	the solution containing one equivalent of solute in one litre of solution
Decinormal solution	:	the solution containing 1/10 equivalent of solute in one litre of solution

Unit 8

Periodic Table and Periodic Laws

Wolfgang Pauli, born on April 25, 1900, Vienna, Austria and died on Dec. 15, 1958, in Switerland. He got Nobel Prize in 1945 in Physics for his discovery Pauli exclusion principle. He, along with Neils Bohr also formulated the Aufbau principle. Pauli made major contributions to the field of physics called quantum mechanics.

Wolfgang Pauli (1900-1958)

8.1 Mendeleev and the modern periodic table

During the early and mid-19th century, scientists had known dozens of elements. But, the study of the properties of these elements was very difficult because they had found no or little pattern among them. In search for a pattern in the elements, chemists tried to arrange the elements on several bases. But, no simple, systematic and scientific method was found till then. But, in 1869 A.D., a Russian chemist called Dmitri Mendeleev came up with an idea of arranging the elements on the basis of their atomic weights and keeping them systematically in a table. This table was called Mendeleev's periodic table.

Mendeleev arranged the elements according to their increasing atomic masses. This rule is called Mendeleev's periodic law. So, Mendeleev's periodic law states that, "**the physical and chemical properties of the elements are the periodic function of their atomic weights**." It means that if we arrange the elements in the order of their increasing atomic weights, we can find elements repeating their physical and chemical properties in a certain interval. A sample of Mendeleev's periodic table is shown below:

	Group I	Group II	Group III	Group IV	Group V	Group VI	Group VII	Group VIII
Period 1	H							
Period 2	Li	Be	B	C	N	O	F	
Period 3	Na	Mg	Al	Si	P	S	Cl	
Period 4	K Cu	Ca Zn	1* 2*	Ti 3*	V As	Cr Se	Mn Br	Fe Co Ni

	Group I	Group II	Group III	Group IV	Group V	Group VI	Group VII	Group VIII
Period 5	Rb Ag	Sr Cd	Y In	Zr Sn	Nb Sb	Mo Te	4* l	Ru Rh Pd
Period 6	Cs Au	Ba Hg	La Th	Hf Pb	Ta Bi	W Po	Re At	Os lr Pt

Name given by Mendeleev: 1 Eka - Aluminium,*

2* Eka - Boron, 3* Eka - Silicon, 4* Eka - Manganese

Characteristics of Mendeleev's periodic table

Mendeleev's periodic table was the first scientific and systematic table. It could arrange the 63 elements which were discovered at his time. The characteristics of the Mendeleev's periodic table are listed below.

1. In Mendeleev's periodic table, the elements are arranged on the basis of increasing atomic masses.
2. The elements are grouped under seven horizontal rows called periods and eight vertical columns called groups.
3. All the groups are sub-divided into two sub-groups except for the eighth group.
4. Gaps were left for the undiscovered elements. Mendeleev predicted that new elements would be discovered in the future to fill these gaps.
5. Inert gases (He, Ne, Ar, Kr, Xe, Rn) are absent in Mendeleev's table because they were not discovered.

Advantages of Mendeleev's periodic table

Mendeleev's periodic table was the first systematic table to predict the properties of the elements. Since, Mendeleev's table was the first scientific and systematic table which Mendeleev formulated, he is also known as the father of periodic table. There are many advantages of his table. Some of the advantages are given below.

1. It was the first organized, systematic and scientific table to show the pattern of the properties of the discovered elements.
2. In this table, metals and non-metals were roughly separated.
3. Gaps were left for the undiscovered elements so that they could fit later on when they are discovered. Some of these elements were called Eka-Aluminium, Eka-Boron, Eka-Silicon and Eka-Manganese (Eka in Sanskrit

means one). Mendeleev predicted the properties of these elements. When these elements were discovered, they matched their properties.

4. With the help of this table, we can calculate atomic weights of the several unknown elements.
5. This periodic table helps to discover other elements by predicting their properties.

> Names of undiscovered elements and their new names after discovery:
>
> a. Eka-aluminium – gallium
> b. Eka-boron – scandium
> c. Eka-silicon – germanium
> d. Eka-manganese – technetium

Drawbacks of Mendeleev's periodic table

Although Mendeleev's periodic table was able to clarify many concepts, it still had many defects. Some of the demerits of the Mendeleev's periodic table are as follows:

1. Mendeleev's periodic table could not locate the position for isotopes of the elements.
2. Mendeleev's periodic law was not universal for all elements. Some element pairs like cobalt-nickel and potassium-argon were placed wrongly in the table. Cobalt (atomic mass =58.9) was placed before Nickel (atomic mass=58.6). Similarly, argon (at. Mass=39.9) was placed before potassium (atomic mass =39.1). This was a violation of the Mendeleev's periodic law.
3. Some highly reactive elements were placed with the low reactive elements in the same group. For example: the highly reactive alkali metals like Li, Na, K, etc. were placed with less reactive metals like Cu, Ag and Au.
4. There is no proper place for lanthanides and actinides in the Mendeleev's table.
5. There were many elements in a single cell which did not match properties with each other.
6. All groups were divided into sub-groups except the eighth group.
7. The position of hydrogen is controversial. It is because hydrogen can be kept both in group IA and VIIA. The Mendeleev's periodic table does not provide the basis on which it should be kept among the elements of group IA.

8.2 Modern periodic table

Mendeleev's periodic table had many drawbacks. To overcome these defects, in 1913 A.D., an English Physicist Henry Moseley and his team realized that arranging the elements on the basis of their atomic weights was not the correct way. Instead, the most fundamental properties of the atoms were atomic number instead of atomic mass. So, they decided to arrange the elements on the basis of increasing atomic number. It is called modern periodic law. The Modern periodic law states that, "**the physical and chemical properties of the elements are the periodic function of their atomic number.**"

Till today, 118 elements are discovered in the modern periodic table. The first element is hydrogen and the 118th element is Oganesson (previously named as Ununoctium).

Characteristics of the Modern periodic table

The modern periodic table is much more detailed and systematic as compared to the Mendeleev's periodic table. Some of its characteristics are listed below:

1. In modern periodic table, elements are arranged according to their increasing atomic numbers.

2. The modern periodic table consists of seven horizontal rows called periods and eighteen vertical columns called groups. A group suggests the total number of electrons in the valence shell while a period suggest the total number of shells present in the atom.

3. Reactive metals, less reactive metals, transitional metals, reactive non-metals and less-reactive non-metals are kept separately. But, metalloids are not perfectly grouped. They are roughly scattered in the right side of the periodic table.

4. Hydrogen is kept in the first period and the first group.

5. Lanthanides and Actinides are kept separately below the main table.

6. Inert gases or noble gases are kept separately in zero '0' group.

7. The whole periodic table is based on atomic subshells and divided into four blocks: s-block, p-block, d-block and f-block. The s-block elements are on the extreme left, p-block elements are on the extreme right, d-block elements are placed at the centre while the f-block elements are kept separately below the main table.

8. Most of the synthetic elements that are prepared in the laboratory are generally present in the last periods (6th and 7th).

Why hydrogen can be placed in group IA	Why hydrogen can be placed in group VIIA
1. Both hydrogen and IA group elements have one valence electron.	1. Both hydrogen and group VIIA have valency one.
2. Both can form electropositive radicals.	2. Both can form electronegative radicals.
3. Both can form halides, oxides, sulphides, etc.	3. Both need 1 electron to complete their stable electronic configuration.
4. Both can react with halogens, oxygen and sulphur.	4. Both can react with metals.

Advantages of the Modern periodic table

The modern periodic table overcomes the drawbacks of the Mendeleev's periodic table and hence is the most scientific tabulation of elements so far. Some of its advantages over the Mendeleev's table are given below :

1. **The position of isotopes is fixed**. Isotopes had no proper position in the Mendeleev's periodic table but in modern periodic table, all isotopes of an element lie in a single cell. It is because they all have same atomic number. For example: the isotopes of hydrogen, viz.protium ($1H^1$), deuterium ($1H^2$) and tritium ($1H^3$) all lie in the first period and in the first group.

2. **The position of some wrongly placed element pairs like the cobalt-nickel and potassium-argon is corrected**. Cobalt (atomic number 27) can be placed before nickel (atomic number 28) because its atomic number is less. Similarly, argon also can be placed before potassium for the same reason.

3. **Hydrogen having atomic number one is placed in the first period and first group**. So, its position is not much controversial in the modern periodic table (though hydrogen has both the properties of group IA and group VIIA).

4. **Lanthanides and Actinides have a fixed position below the main table**. They are kept below main table separately because their properties do not match with the groups below if they are kept linearly with other elements. Also, these two series represent the f-blocks which should be kept separately without mixing with other blocks.

5. **The reactive metals, less reactive metals, reactive non-metals, less reactive non-metals and the noble gases are kept separately**. Similarly, the rare earth elements and synthetic elements also are kept separately as far as possible.

6. **The elements are arranged according to the electronic configuration of their orbitals, viz. s, p, d and f**. This divides the periodic table into four main blocks, i.e. s, p, d and f block.

Drawbacks of the modern periodic table

Even though the modern periodic table corrects many mistakes of the Mendeleev's periodic table and is itself very systematic, it is not perfect. It is because the elements in the nature cannot be grouped into perfect patterns. Some of the drawbacks of the modern periodic table are as follows:

1. The position of hydrogen is still controversial since it can be placed both in group IA and group VIIA.
2. The lanthanides and actinides do not have a clear group and their position is still separate from the rest of the elements.
3. The element Helium is kept in p-block but its last electron falls into the s-block.

The s,p,d,f block concept in the modern periodic table

The modern periodic table puts its elements on the basis of increasing atomic numbers. But, it also arranges its elements on the basis of their electronic configuration. On the basis of where the last electron falls, the periodic table of 118 elements is divided into four blocks:

1. **The s-block elements:** The group of elements in which the last electron falls upon the s-orbital are called s-block elements. There are two groups in the s-block, viz. group IA and group IIA. The s-block also includes the hydrogen and helium (although helium is placed in the p-block). Some examples of s-block elements are:
   a. Sodium (11): $1s^2\ 2s^2\ 2p^6\ 3s^1$ (last electron falls in the s-orbital)
   b. Calcium (20): $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6\ 4s^2$ (last electron falls in the s-orbital)
2. **The p-block elements:** The group of elements in which the last electron falls upon the p-orbital are called p-block elements. Group IIIA to zero group lie in the p-block.
   a. Oxygen (8): $1s^2\ 2s^2\ 2p^4$ (last electron falls in the p-orbital)
   b. Argon (18): $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6$ (last electron falls in the p-orbital)
3. **The d-block elements:** The group of elements in which the last electron falls upon the d-orbital are called d-block elements. These elements are placed at the centre of the periodic table. They are also called transitional elements because their last electron is in transition between the last and second last orbital. They also have variable valency due to this reason.
   a. Iron (26): $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6\ 4s^2\ 3d^6$ (last electron falls in the d-orbital)
   b. Zinc (30): $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6\ 4s^2\ 3d^{10}$ (last electron falls in the d-orbital)

4. **The f-block elements:** The group of elements in which the last electron falls upon the f-orbital are called f-block elements. Lanthanides and actinides are the f-block elements. They are kept separately below the main table. Lanthanides are also called rare earth elements. It is because they are found naturally in trace amounts on the earth. The f-block elements are also called inner-transitional elements. It is because they are derived from the inner 6^{th} and the 7^{th} periods of the transitional elements.

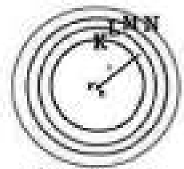
S.N.	**Block**	**Position in the periodic table**	**Group**	**Periods**
1	s-block	Left	IA and IIA	1^{st} to 7^{th}
2	p-block	Right	IIIA to 'o' group	1^{st} to 7^{th}
3	d-block	Centre	IB to VIII group	4^{th} to 7^{th}
4	f-block	Below	Ungrouped	6^{th} and 7^{th}

8.3 The Importance of a Periodic Table

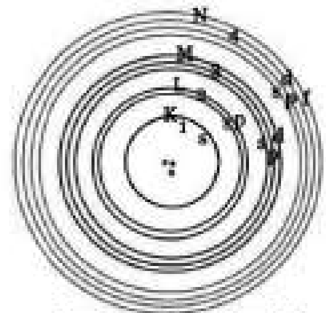
The modern periodic table is a useful tool for students, chemists, scientists and even common people for studying elements, their properties, and relationship between different groups, periods and the change of their properties according to the trend. The detailed form of the modern periodic table not only mentions the name, atomic number and position of elements but also their atomic weights, electronic configuration, nature, isotopes and other characteristics. Therefore, it is extremely useful. The importance of periodic table is listed below:

1. Detailed information of the elements like the number of electrons, number of protons, electronic configuration, atomic number, atomic weight, valency, nature, block, etc. can be obtained from the periodic table.
2. The properties of elements can be known by studying the properties of a group.
3. The trend of change in different properties of the elements like the atomic number, atomic radius, electronegativity, electropositivity etc. is uniform from top to bottom or left to right. This makes understanding chemical reactions easier.
4. The characteristics of the elements can be known by knowing their exact position in the periodic table.
5. The properties of the elements that are not yet discovered also can be predicted.
6. It makes the arrangement of elements systematic and scientific.

8.4 The Aufbau Principle



Bohr's model



Quantum mechanical model

The protons and the neutrons of an atom are located at the central region of an atom called nucleus. The electrons revolve around the nucleus with varying speeds and at varying distance from the nucleus. The distribution of electrons around the nucleus of an atom is not random. The Bohr's model of an atom predicted (wrongly) that electrons remain at a fixed distance from the nucleus and revolve around an atom. But, the more accurate model, i.e.quantum mechanical model suggests that electrons can revolve around the nucleus in certain allowed regions from the nucleus. These regions are distributed around the nucleus and are called shells. The shells have small regions of varying energies called as sub-shells. Likewise, inside a sub-shell, there are tiny regions of similar energy called orbitals. The electrons in an atom are filled from the lower energy levels to the higher energy levels. This concept was given by a principle called Aufbau principle.

The Aufbau principle is the rule according to which the electrons are arranged in the energy levels of an atom. The word Aufbau is German and means "building up or construction". This concept was formulated by the scientists Niels Bohr and Wolfgang Pauli. It states that, "**the filling of electrons occurs from the lower energy levels to the higher energy levels**."

An explanation of the Aufbau principle is given below:

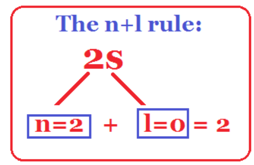
From the diagram shown above, the quantum mechanical model of an atom suggests that electrons revolve around the nucleus in certain three dimensional regions called orbitals (which are the part of the sub shells and sub shells are the part of a shell). The figure shows four shells in an atom. These four shells K, L, M and N represent by the four principal quantum numbers, i.e. n=1,2,3 and 4 respectively. Each shell has several specific regions where electrons could be found and are called sub-shells (azimuthal quantum numbers). The shells represented by the principal quantum number and their corresponding sub-shells or azimuthal quantum numbers is given in the table below:

S.N.	Principal quantum numbers	Corresponding azimuthal quantum numbers
1	For K-shell, n=1	l = 0
2	For L-shell, n=2	l = 0, 1
3	For M-shell, n=3	l = 0, 1, 2
4	For N-shell, n=4	l = 0, 1, 2, 3

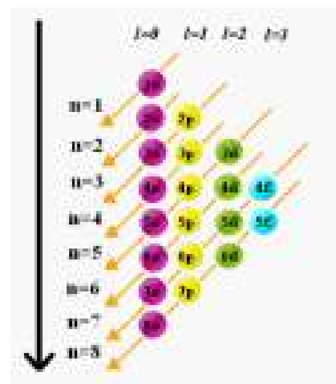
To understand the electron distribution, we should know the energy of each shell. The energy of the sub-shells can be known by adding their principal and azimuthal quantum numbers i.e.

$$\text{Energy of a sub-shell} = n + l$$

The sub shell having less value of n+l has less energy level and so it fills earlier than the shell having higher value of n+l. If two sub-shells have the same n+l value, the sub-shell having lower principal quantum number has the lower energy level and hence it fills first.



For example: the first sub-shell, i.e. 1s has principal quantum number 1 and azimuthal quantum number 0. Their sum gives the value 1+0=1. Likewise, the sum of all the shells and sub-shells is given below.



Thus, from the given table, the (n+l) value of 1s orbital is less than that of "2s". Therefore, the energy of "1s" orbital is also less than that of "2s". Also, notice that the energy level of 4s orbital is less than that of 3d.

Thus, the energy of the orbitals can be arranged in the ascending order as follows:

1s < 2s < 2p < 3s < 3p < 4s < 3d .... The filling of electrons also occurs in this sequence. "1s" orbital is filled first, then "2s" and so on.

Alternatively, there is an easy method to find out the energy of the orbitals using this arrow sequencing, as shown in the figure alongside. The arrows indicate that the orbitals are filled in first. Thus, the filling of electrons occurs in the following sequence of order from the lower to the higher energy level:

Orbitals	Principal quantum numbers (n)	Azimuthal quantum numbers (l)	Sum: n + l
1s	1	0	1+0=1
2s	2	0	2+0=2
2p	2	1	2+1=3
3s	3	0	3+0=3
3p	3	1	3+1=4
3d	3	2	3+2=5
4s	4	0	4+0=4

1s < 2s < 2p < 3s < 3p < 4s < 3d < 4p < 5s < 4d < 5p < 6s < 4f < 5d < 6p < 7s and so on.

The maximum number of electrons that can be accommodated in each of the s,p,d and f orbitals are 2, 6, 10 and 14 respectively.

S.N.	Orbitals	Maximum number of electrons that can be accommodated
1	s	2
2	p	6
3	d	10
4	f	14

Electronic configuration of the elements

The traditional electronic configuration of elements is written as the number of electrons that are filled in each shell. For example, the electronic configuration of sodium is 2,8,1 and calcium is 2,8,8,2. This method is easy and clear for the first 20 elements. But, beyond atomic number 20, it does not express the true electronic configuration. For example: suppose the electronic configuration of the element iron (atomic number 26). According to this rule, it seems the electronic configuration is 2,8,8,8. Isn't it? But, it is not the case. The actual electronic configuration is 2,8,14,2. But, how?

This can be explained by using the electronic configuration using the sub-shells s, p, d and f by Aufbau principle. It predicts the actual number of electrons that can accommodate in each shell. The electronic configuration of Iron is,

Iron (26) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^2$, $3d^6$

Arranging in the ascending order,

Iron (26) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $3d^6$, $4s^2$

Iron (26) =	**$1s^2$**	**$2s^22p^6$**	**$3s^23p^63d^6$**	**$4s^2$**
	2	**2+6 = 8**	**2+6+6 = 14**	**2**
	K	**L**	**M**	**N**

Similarly, the electronic configuration of the first 20 elements is given below:

S.N.	Element	Atomic number	Sub-shell electronic configuration
1	Hydrogen	1	$1s^1$
2	Helium	2	$1s^2$
3	Lithium	3	$1s^2$, $2s^1$
4	Beryllium	4	$1s^2$, $2s^2$
5	Boron	5	$1s^2$, $2s^2$ $2p^1$
6	Carbon	6	$1s^2$, $2s^2$ $2p^2$
7	Nitrogen	7	$1s^2$, $2s^2$ $2p^3$
8	Oxygen	8	$1s^2$, $2s^2$ $2p^4$

S.N.	Element	Atomic number	Sub-shell electronic configuration
9	Fluorine	9	$1s^2$, $2s^2$ $2p^5$
10	Neon	10	$1s^2$, $2s^2$ $2p^6$
11	Sodium	11	$1s^2$, $2s^2$ $2p^6$, $3s^1$
12	Magnesium	12	$1s^2$, $2s^2$ $2p^6$, $3s^2$
13	Aluminium	13	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^1$
14	Silicon	14	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^2$
15	Phosphorus	15	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^3$
16	Sulphur	16	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^4$
17	Chlorine	17	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^5$
18	Argon	18	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$
19	Potassium	19	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^1$
20	Calcium	20	$1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^2$

Limitations to Aufbau principle

The Aufbau principle predicts the electronic configuration of the elements to near accuracy. However, some elements do not follow this rule. For example, the wrong electronic configuration of copper and chromium as predicted by the Aufbau principle is:

Copper (29) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^2$, $3d^9$

Chromium (24) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^2$, $3d^4$

But, the actual electronic configuration is:

Copper (29) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^1$, $3d^{10}$

Chromium (24) = $1s^2$, $2s^2$ $2p^6$, $3s^2$ $3p^6$, $4s^1$, $3d^5$

This is because the d-orbital has the tendency to become either half filled (with 5 electrons) or fully filled (with 10 electrons) if it is nearly in half or full filled state. This is because a half filled or full filled d-orbital is more stable.

8.6 Valency and Variable Valency:

Except noble gases, all elements in the periodic table have the tendency to form compounds with other elements. When an element or a radical react with other elements or radicals, they give, take or share their outermost or valence electrons. After they give, take or share electrons, they attain a stable state of electronic configuration known as octet or duplet. Thus, valency is the number of electrons gained, lost or shared by an element or the number of charges present in a radical to form octet or duplet. Hence, it is also the combining capacity of an element. Generally, the valency of elements can be found by counting the outermost electrons. For example: the valency of sodium is 1 as its electronic configuration is 2,8,1 and it can give one electron to the other atom to form octet in the second shell. Likewise, the valency of oxygen is 2 as its electronic configuration is 2,6 and hence it can take 2 electrons form other atom to form octet in its second shell. In some ionic and covalent compounds, the valency of the element is the number of atoms of other elements that react with a single atom of that element. For example, in $MgCl_2$, the valency of magnesium is 2 which is the number of chlorine atoms that react with one atom of magnesium. Likewise, the valency of chlorine is 1. But, it does not apply to all compounds. For example: in CO_2, the valency of carbon is not 2.

Some elements, their electronic configuration and valencyis given below:

S.N.	**Element**	**Electronic configuration**	**Valency**
1	Hydrogen	1	1 (can give or take 1)
2	Helium	2	0 (already duplet)
3	Boron	2,3	3 (can give 3)
4	Magnesium	2,8,2	2 (can give 2)
5	Argon	2,8,8	0 (already octet)
6	Chlorine	2,8,7	1 (can take 1)
7	Sulphur	2,8,6	2 (can take 2)
8	Potassium	2,8,8,1	1 (can give 1)

The valency of radicals is the amount of charge they have on them. The charges are formed as a result of deficient or excess electrons they have. Radicals are of two types, i.e. electropositive or basic radicals and electronegative or acid radicals. Some radicals and their valencies are given below:

S.N.	Electropositive radicals	Valency	Electronegative radicals	Valency
1	Ammonium (NH_4^+)	1	Sulphate (SO_4^{--})	2
2	Sodium (Na^+)	1	Nitrate (NO_3^-)	1
3	Calcium (Ca^{++})	2	Carbonate (CO_3^{--})	2
4	Iron (Fe^{++} or Fe^{+++})	2 or 3	Hydroxide (OH^-)	1
5	Copper (Cu^+ or Cu^{++})	1 or 2	Phosphate (PO_4^{---})	3
6	Magnesium (Mg^{++})	2	Chloride (Cl^-)	1
7	Hydrogen (H^+)	1	Cyanide (CN^-)	1

Variable valency

Observe iron and copper in above table. They show more than one valency. This means iron has either 2 or 3 combining capacity depending upon the type of element it reacts and the nature of reaction.This shows that some elements have more than one combining capacity. This property of an element is called variable valency. The existence of variable valency of elements means that they can give, take or share different number of electrons with different atoms or in different conditions. The elements of d-block in the modern periodic table especially exhibit variable valency.

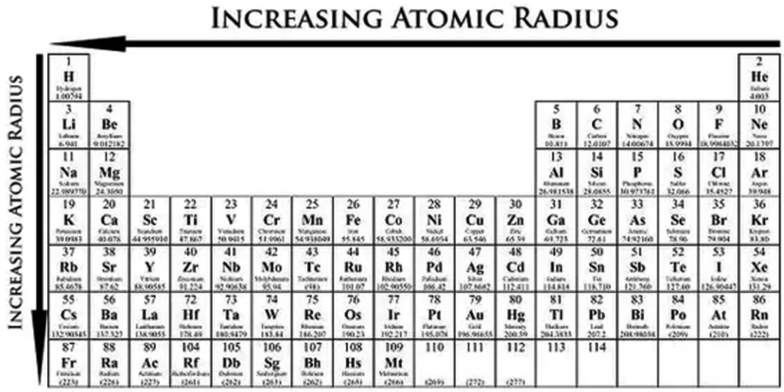
Some elements, their variable valency and the names of radicals that they form are given below:

S.N.	**Elements**	**Radicals formed**	**Valency**
1	Iron	Ferrous (Fe^{++} or Fe^{2+}) Ferric (Fe^{+++} or Fe^{3+})	2 3
2	Copper	Cuprous (Cu^{+}) Cupric (Cu^{++} or Cu^{2+})	1 2
3	Tin	Stannous (Sn^{++} or Sn^{2+}) Stannic (Sn^{++++} or Sn^{4+})	2 4
4	Lead	Plumbous (Pb^{++} or Pb^{2+}) Plumbic (Pb^{++++} or Pb^{4+})	2 4
5	Mercury	Mercurous (Hg^{+}) Mercuric (Hg^{++} or Hg^{2+})	1 2
6	Gold	Aurous (Au^{+}) Auric (Au^{+++} or Au^{3+})	1 3
7	Antimony	Antimonous (Sb^{+++} or Sb^{3+}) Antimonic (Sb^{+++++} or Sb^{5+})	3 5

8.7 Periodic Variation

The modern periodic table has horizontal rows of elements called periods and vertical columns of elements called groups. The elements in the groups are similar to each other in some ways and differ in other ways. Similarly, the elements across a period also have some similarity and differences. The variation of the properties of elements in periods and groups are periodic. It means that the same properties occur at regular intervals. **The variation of properties of elements across a period or down the group of the modern periodic table in a periodic manner is called periodic variation.**

1. **Atomic size**

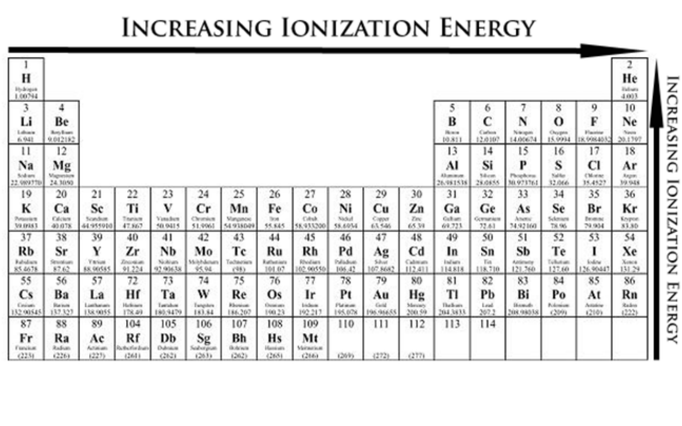


An atom is a three dimensional structure. It is similar to a sphere. Therefore, it has a certain size. The size of atoms is generally measured in terms of atomic radius. Atomic radius is the distance of the valence electron in the outermost orbital from the nucleus of an atom. It is roughly measured as its accurate value cannot be known. The periodic trend in the atomic size is discussed below:

a. Across a period, the atomic size decreases. It is because the number of shells remain constant but in each element, the number of protons and electrons increase. The increase in the number of protons is more effective which causes shrinking of an atom. So, atomic size decreases in the period left to right.

b. Down the group, the atomic size increases. The reason is due to the addition of a shell in each step as we go down a group. Due to the addition of the subsequent shells, the distance from the nucleus to the valence shell increases.

2. **Ionization potential**

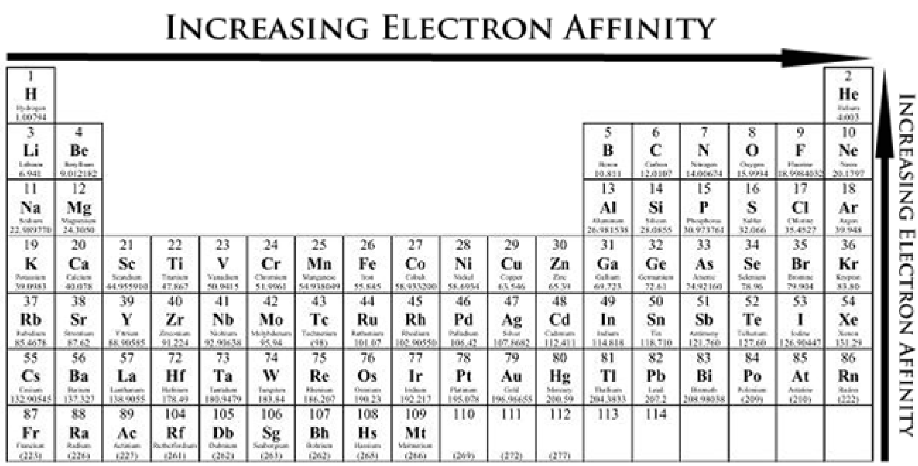


An ion is an atom which has given or taken electron/s and acquired positive or negative charge. Ionization potential or ionization energy is the amount of energy required to remove an electron from the valence shell of an isolated gaseous atom. The unit of ionization energy is kJ/mol and electron volt (eV). The less the value of ionization potential, the easier it is to remove the electron from the valence shell. Likewise, if the value of ionization potential is more, it requires more energy to remove the electron and hence is more difficult. The variation of IP in period and group is as follows:

a. Across a period, the ionization potential increases. It is because the atomic size decreases and the electrons are nearer to the nucleus due to which the attractive force is higher.

b. Down the group, the ionization potential decreases. It is because the atomic size of elements increases and thus electrons are far away from the nucleus. Since, it is easier to take out the electrons that are far away from the nucleus, the ionization potential down the group decreases.

3. **Electron affinity**

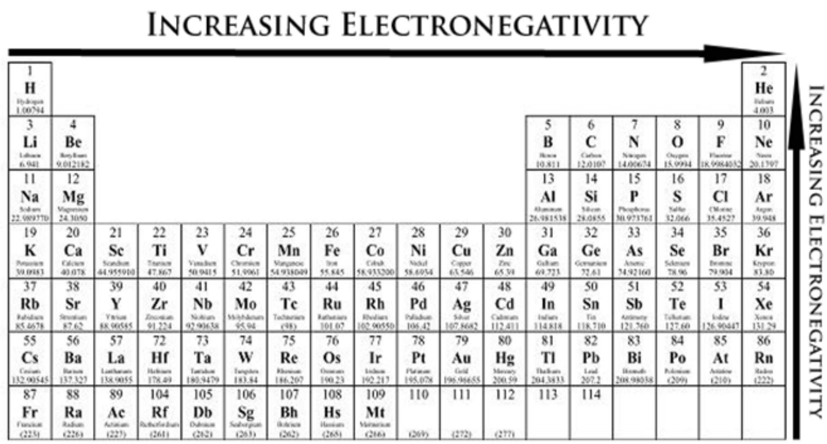


Electron affinity is conceptually opposite to that of ionization potential. When an electron is added to the valence shell of a neutral atom, the atom releases energy. The amount of energy released by an atom or the amount of energy change that occurs in an atom when an electron is added to its valence shell is called electron affinity. Thus, the electron affinity is the measurement of the electron pulling ability of an atom. The higher its value, the higher is the ability to pull an electron and the lower its value, the lower is the ability to pull the electron. Its unit is also kJ/mol and eV. The variation of electron affinity in periods and groups are as follows:

a. Across a period, the electron affinity increases. This is because the atomic size decreases across a period and as the size of the atom is lesser, the protons are nearer to the valence shell. Due to this decreased distance between the protons and electrons, there is more electrostatic attraction to pull the electrons.

b. Down the group, the electron affinity decreases. It is because the atomic size increases as we go down the group and when atomic size increases the protons become farther from the valence shell. This decreases the ability of the nucleus to pull electrons in the valence shell. Therefore, the electron affinity decreases.

4. **Electronegativity**



Electron affinity is the quantitative measure of electron pulling capacity of a neutral atom whereas electronegativity is the qualitative measure. This means electron affinity is the measure of energy but electronegativity is the estimation of ability of a neutral atom to pull an electron. **Electronegativity is defined as the ability of an atom to pull or attract an electron in its outermost shell.** It is measured in a scale called Pauling scale named after the chemist Linus Pauling. The variation of electronegativity in periods and groups is as follows:

a. Across a period, the electronegativity increases. It is because the atomic size of the elements across a period decreases. As the atomic size decreases, the protons which have the ability of pulling the electrons are nearer to the valence shell. This increases the ability of an atom to pull electrons.

b. Down a group, the electronegativity decreases. This is because the atomic size of elements down a group increases. The increase in the atomic size makes the valence shell farther from the protons which ultimately decreases the ability of the atom to pull electrons.

Summary

1. There are four orbitals in atoms where electrons might be found. They are s, p, d and f orbitals.
2. The elements in the modern periodic table are arranged according to increasing atomic number.
3. The Aufbau principle suggests that the filling of electrons should occur from the lower to the higher energy levels.
4. The arrangement of electrons in different shells of an atom is called electronic configuration.
5. Valency is the combining capacity of an atom with other atoms. It generally represents the number of electrons given, taken or shared.
6. Elements like iron, copper, mercury, lead, gold etc. exhibit variable valency.
7. The capacity of an atom to pull an electron in its outermost shell is called electronegativity.
8. The energy required to remove the outermost electron of an atom in its isolated gaseous state is called ionization energy.
9. Electron affinity is the amount of energy released when an electron is added to a neutral atom or molecule in the gaseous state to form a negative ion.

Exercise

A. **Tick (√) the best alternatives.**

1. Electrons are arranged from the.............

   i. Lower to higher energy levels

   ii. Higher to lower energy levels

   iii. Both lower to higher energy levels

   iv. Mostly lower to higher but sometimes higher to lower

2. The maximum number of electrons accommodated in the s and p orbitals are

   i. 2 and 5   ii. 2 and 9

   iii. 3 and 6   iv. 2 and 6

3. The valency of H in H_2SO_4 is

i. 1 ii. 2

iii. 3 iv. 4

4. Ferrous and ferric have the following symbols:

i. Fe^{2+} and Fe ii. Fe^{3+} and Fe^{2+}

iii. Fe^{2+} and Fe^{3+} iv. Fe and Fe^{+++}

5. The ionization energy ………………… ongoing form left to right of the periodic table.

i. Increases ii. Decreases

iii. First increases and then decreases

iv. First decreases and then increases

B. **Write very short answers to these questions.**

1. What is periodic table?
2. State Mendeleev's periodic law.
3. State Modren periodic law.
4. Write the meaning of valency.
5. State Aufbau principle.
6. Define electronic configuration.
7. Write two elements which exhibit variable valency.
8. What is periodic variation?
9. Define Ionization potential.
10. What is electronegativity?

C. **Write short answers to the following questions.**

1. Write the electronic configuration of sodium and calcium based on sub-shells.
2. Explain the demerits of the Mendeleev's periodic table in the points.
3. Enlist the advantages of the Mendeleev's periodic table.
4. Mention any three importance of periodic table.
5. Draw an electron distribution chart according to Aufbau principle.
6. How does atomic size change from top to bottom in a group?

7. Write the trend of variation of ionization potential and electronegativity down a group.
8. Why do some elements have variable valency?
9. Elements of IA group of the modern periodic table are called alkali metals. Why?
10. Elements of IIA group of the modern periodic table are called alkaline earth metals. Why?
11. Fluorine is kept in p-block of the modern periodic table. Why?
12. Potassium is kept in the s- block of the periodic table. Why?
13. Electron affinity increases in the period left to right and decreases top to bottom in the group. Why?

Activity

1. Make a sequence order of Aufbau principle using arrow head diagram in the chart paper and display on the wall.
2. Make a modern periodic table in the chart paper and show the variation of atomic size, valency, electronegativity and electron affinity across the period and in the group.

Glossary

Isotopes : group of atoms which have the same atomic number but different atomic mass.

Ionization energy : the amount of energy required to remove outermost electron from an atom

Electropositivity : the tendency of an element by which it loses electron to become cation.

Electronegativity : the tendency of an element to attract foreign electron.

Electronic configuration : the systematic distribution of electrons in the shells and sub-shells

Periodic variation : the change in the characteristics in the periods and groups.

Unit 9 Chemical Bonding

Avogadro (1776-1856)

Avogadro was born in 1776 and died in 1856. He is best known for his law that equal volumes of different gases contain an equal number of molecules, provided they are at the same temperature and pressure. He worked as a Professor of mathematical physics in Turin University

9.1 Chemical bonding

A bond is something that attaches two or more things together. For example, a glue attaches two pieces of woods together. The glue forms a bond between these pieces of woods. But, the bond we are talking about here is a physical bond. Chemical bond is different from a physical bond.

There are altogether 118 known elements in the periodic table. Most of these elements can form compounds by chemically combining with each other (except for some noble gases). When two or more elements combine with each other, an attractive force is produced between them. This attractive force binds them together. This attractive force is called chemical bond. Therefore, **the force of attraction which binds two or more elements together to make chemical compound is called bonding**. In order to make chemical bonding, it is generally necessary for an atom to attain stable electronic configuration called duplet and octet.

1. Octet

Octet is the state in which there are 8 electrons in the valence (outermost) shell of an atom. If an atom has 8 electrons in its outermost shell, it becomes stable and hence it can neither give or take nor share electrons in normal conditions. Therefore, octet is called stable electronic configuration. Some elements are already in octet state. For example, neon (Ne), Argon (Ar), Krypton (Kr), Xenon (Xe) and Radon (Rn). Likewise, some elements can give, take or share electrons to become octet. Most of the elements (except noble gases) like, sodium, potassium, calcium, chlorine, fluorine, etc. can attain octet in normal conditions after losing, gaining or sharing of electrons. It is call octet rule.

2. Duplet

Duplet is the state of stable electronic configuration in which an atom possesses only one shell (K-shell) with 2 electrons in it. Helium is an example of duplet atom. Hydrogen, lithium, beryllium and boron can attain duplet by giving, taking or sharing electrons. It is called duplet rule.

When two or more elements combined to make compounds, a chemical bond is formed between them. So due to chemical bond it is possible to make compounds. According to the nature of elements, different kinds of bonds are formed between the constituent atoms. Now, we will discuss the different types of chemical bonds present in the chemical compounds. The three major types of chemical bonds are given below.

1. Ionic or Electrovalent Bond
2. Covalent Bond
3. Coordinate Bond

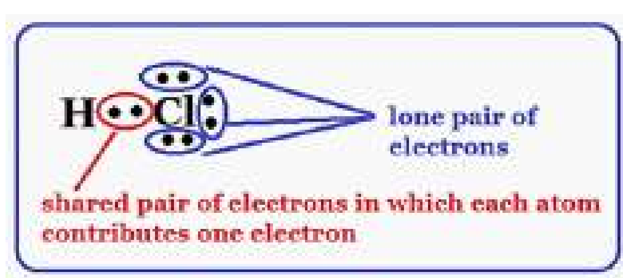
1. Ionic or electrovalent bond

The bond which is formed by giving and taking of electrons between two atoms or groups of atoms is called ionic bond. In an ionic bond, the atoms which donate electrons is called an electropositive ion or basic radical or cation. Similarly, the atom which receives electrons is called electronegative ion or acid radical or anion. The compounds containing electrovalent bond are called electrovalent compounds or ionic compounds. For example, sodium chloride (NaCl), magnesium chloride ($MgCl_2$), calcium oxide (CaO), potassium oxide (K_2O), etc.

2. Covalent bond

The bond in which one or more pairs of electrons are shared between two atoms or group of atoms is called covalent bond. Covalent bonding occurs by sharing of electrons when giving and taking electrons is generally not possible. When electrons are shared, both atoms contribute equal number of electrons. These electrons remain in between the sharing elements. The compounds which have covalent bonds in them are called covalent compounds. For example, H_2, N_2, O_2, Cl_2, H_2O, NH_3, CH_4, CCl_4, etc.

If a covalent bond exists between two atoms, the total pair of electrons they share come from both the atoms in equal number, i.e. each sharing atom contributes equal number of electrons to the bond. Let us take an example of the compound HCl. Hydrogen has one electron in its valence shell while chlorine has seven. In order to form a stable compound between H and Cl atoms, hydrogen has to form duplet and chlorine has to form octet. The number of electrons that hydrogen needs to complete its duplet is 1 and similarly chlorine also needs 1 electron to



complete its octet. Therefore, both elements provide one electron each to the bonding that they share between them, as shown in the figure. In this case, giving and taking electrons is not possible. The pair of electrons that both hydrogen and chlorine share is called shared pair of electrons. Likewise, the pair of unshared electrons that do not form the bond are called lone pairs of electrons. In the compound HCl, there are:

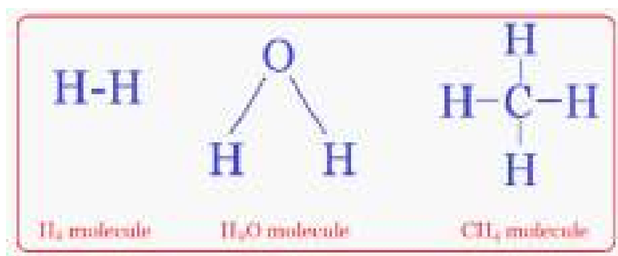
a. No lone pairs of electrons in hydrogen

b. 3 lone pairs of electrons in chlorine

c. One pair of shared electrons

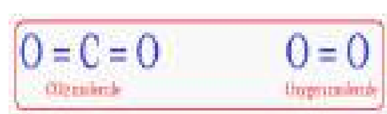
On the basis of number of electrons shared, there are three types of covalent bonds. They are:

a. **Single covalent bond**



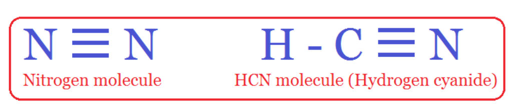
The type of covalent bond in which one pair of electrons is shared between two atoms in the bond is called single covalent bond. For example, the bond between Hydrogen atoms in H_2, the bond between hydrogen and oxygen atoms in H_2O, the bond between carbon and hydrogen atoms in CH_4, etc. The single covalent bond is represented by a single line (—) between any two elements.

b. **Double covalent bond**



The type of covalent bond in which two pairs of electrons are shared between two atoms in the bond is called double covalent bond. For example, the bond between oxygen atoms in O_2, the bond between carbon and oxygen atoms in CO_2, etc. The double covalent bond is represented by a double lines (═) between two atoms.

c. **Triple covalent bond**



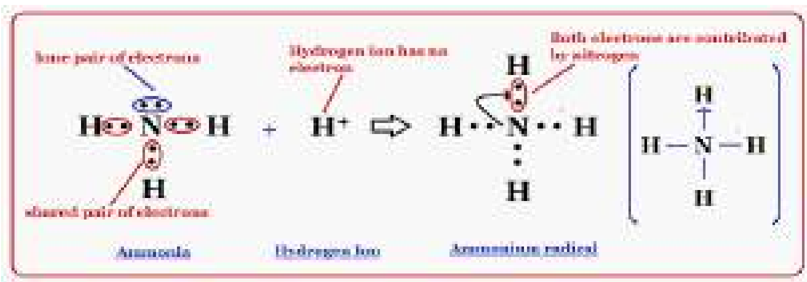
The type of covalent bond in which three pairs of electrons are shared between two atoms in the bond is called triple covalent bond. For example, the bond between nitrogen atoms in N_2, the bond between carbon and nitrogen atoms in HCN, etc. The triple covalent bond is represented by a triple lines (≡) between atoms.

3. Coordinate Bond

A coordinate bond is similar to a covalent bond. The only difference is that in a covalent bond, both atoms contribute equal number of electrons to the bond but in a coordinate bond, only one atom contributes for the bonding. It is because the other atom does not have sufficient number of electrons to form the bond. Consider the formation of ammonium radical (NH_4^+). The ammonium radical is formed by the combination of ammonia (NH_3) and hydrogen ion (H^+). i.e.

$NH_3 + H^+ \rightarrow NH_4^+$

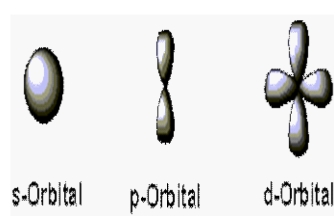
The special types of covalent bond which is formed by sharing lone pair of electrons between two atoms is called coordinate bond.



In ammonia, the nitrogen has a lone pair of electrons which is not bonded with any hydrogen atoms. Similarly, the hydrogen ion has lost one electron and so it does not have any electron to share with ammonia. When they combine, the nitrogen alone contributes a lone pair of electrons because hydrogen has zero valence electrons to share. Thus, nitrogen bonds with the hydrogen atom and forms a coordinate bond. We represent a coordinate bond by an arrow (→) instead of a dash. In this ammonium radical, nitrogen singly donates a pair of electrons to the bond, so it is called donor atom. Likewise, hydrogen receives the electrons pair and therefore it is called recipient atom.

Sigma and Pi bonds:

When atoms combine with each other to form a covalent bond, their valence orbitals overlap with each other. Different orbitals have different shapes. The s-orbital is spherical, p-orbital is dumb bell shaped, d-orbital is double dumb bell shaped and the shape of f-orbital is complex. When atoms combine with each other, these orbitals overlap in different ways.



The type of bond that forms between orbitals depends on how their orbitals overlap. The s orbital is spherical in shape and p orbital is dumb bell in shape as shown in the figure. The overlapping of orbitals can occur between s and s orbitals, p and p orbitals, s and p orbitals and so on. The bond formed by the overlapping of atomic orbitals is called sigma and pi bonds. These bonds are discussed below:

1. **Sigma Bond (σ–bond)**

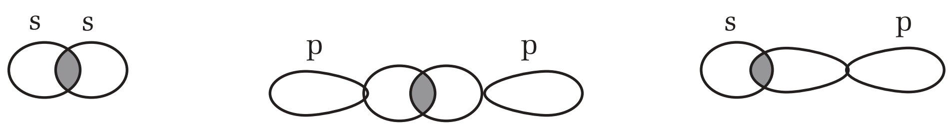
The type of bond formed between two orbitals by the head to head overlapping of the atomic orbitals is called sigma bond. It is represented by a Greek letter sigma (σ). A sigma bond is formed between:

a. Two s-orbitals (by head to head overlapping)

b. Two p-orbitals (by head to head overlapping)

c. s and p orbital (by head to head overlapping) and so on.

A sigma bond is stronger than pi bond.

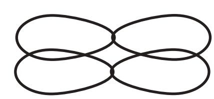


(a) s-s overlapping (b) p-p overlapping (c) s-p overlapping

*Figure of Bond (σ–**bond**)*

2. **Pi Bond (π- bond)**

The type of bond formed between two orbitals by the side to side overlapping of the atomic orbitals is called pi bond. It is represented by a Greek letter Pi (π). Pi bond is weaker type of covalent bond compared to the sigma bond. A pi bond formed by the overlapping of two p orbitals is given alongside.



p-p overlapping for π-bond

In carbonic compounds, the bonds between any two carbon atoms are of three types. They are C-C (single bond), C = C (double bond) and C ≡ C (triple bond). The total number of sigma and pi bonds between these bonds is given in the table below:

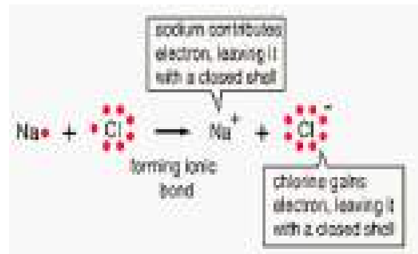
S.N.	Type of bond	Total number of sigma bond	Total number of pi bond
1	C — C	1	0
2	C = C	1	1
3	C ≡ C	1	2

It is evident that in a compound having single bond between two carbon atoms or C-H bond, there is a single sigma bond. Similarly, in a double bond, there is one sigma bond and one pi bond and in a triple bond, there is one sigma bond and two pi bonds.

Examples of ionic bonds

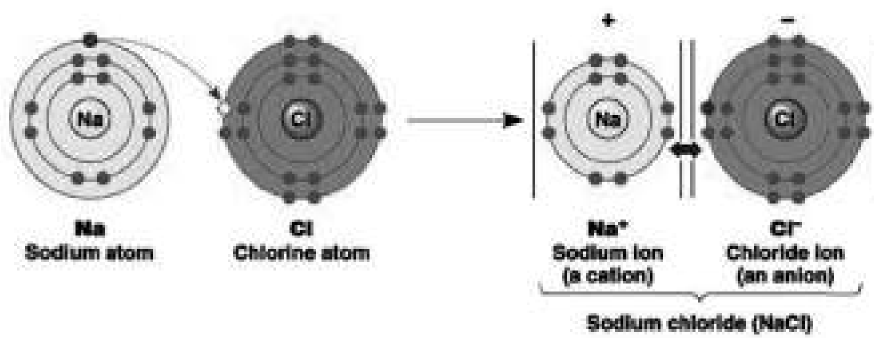
Those chemical compounds which are formed as a result of electrovalent bonding are called electrovalent compounds or ionic compounds. The structure of different ionic compounds like NaCl, $MgCl_2$,CaO, etc.are described below:

a. Structure of Sodium chloride (NaCl)



Sodium is a metal and chlorine is a non-metal. The electronic configuration of sodium is 2,8,1 and chlorine is 2,8,7. When sodium simply reacts with chlorine, they undergo chemical reaction to form sodium chloride (NaCl). In this process, the sodium atom loses one electron from its valence shell to form octet in its second shell. This donated electron is taken by the chlorine atom to become octet in its third shell. This makes sodium a cation and chlorine an anion. Cation and anion are opposite charged. Therefore, they attract each other strongly and stay close to each other forming the ionic bond. Hence, Na^+ and Cl^- attract each other strongly and form a bond called ionic bond. Note that it is called ionic bond because it is formed between ions.

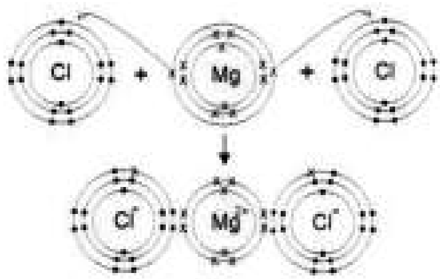
Thus, Na (2,8,1) + Cl (2,8,7) $\rightarrow$ Na^+ (2,8) Cl^- (2,8,8)

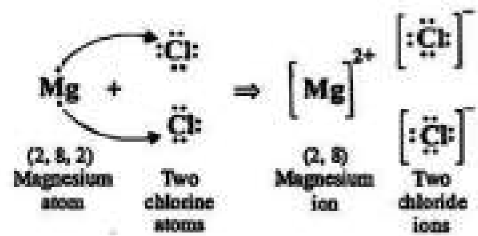


b. Structure of Magnesium chloride ($MgCl_2$)

Magnesium chloride contains one magnesium atom and two chlorine atoms. The electronic configuration of magnesium is 2,8,2 and that of chlorine is 2,8,7. When magnesium reacts with chlorine, magnesium chloride is formed. During this process, magnesium donates its two valence electrons to the two chlorine atoms so that all magnesium and chlorine atoms become stable by attaining octet in their valence shells. After giving and taking electrons, magnesium becomes positively charged, i.e. Mg^{2+} and each chlorine atom becomes negatively charged, i.e. Cl^-. Due to the formation of opposite charges, a strong force of attraction exists between them known as ionic bond. Thus, the formation of $MgCl_2$ takes place in this way:

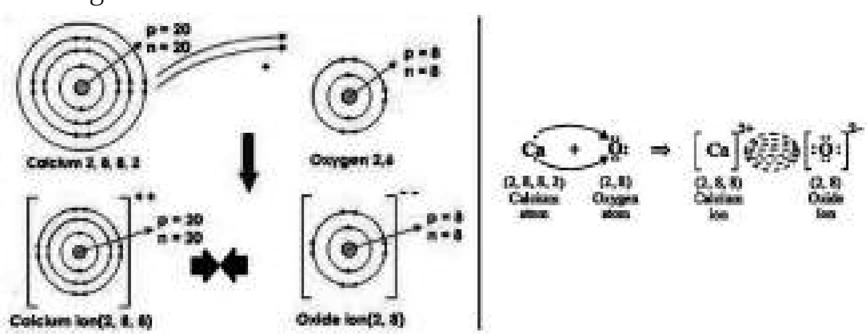
Thus, Mg (2,8,2) + 2Cl (2,8,7) → Mg^{++} (2,8) Cl^{-} (2,8,8) Cl^{-} (2,8,8)





c. **Structure of Calcium oxide (CaO)**

Calcium is a metal and oxygen is a non-metal. The electronic configuration of calcium is 2,8,8, 2 and that of oxygen is 2,6. When calcium reacts with oxygen, calcium oxide is formed. In this process, the calcium atom loses its two valence electrons and oxygen atom gains these two electrons. After Ca loses two electrons and oxygen gains two electrons, both attain octet and become stable. Calcium now gains two positive charges (Mg^{++}) by losing two electrons and oxygen gains two negative charges (O^{--}) by receiving two electrons. After they become opposite ions, a strong force of attraction exists between them i.e. ionic bond.

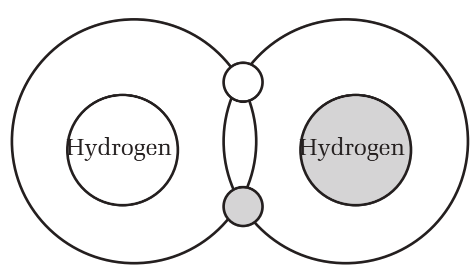


Examples of covalent molecules

The structure of different covalent molecules like O_2, N_2, H_2, H_2O, CH_4, NH_3, etc. is described below.

a. **Structure of hydrogen molecule (H_2)**

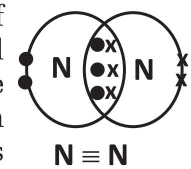
H_2 is the molecule of hydrogen. Hydrogen has one electron in its valence shell. A single atom of hydrogen combines with another hydrogen atom to form a molecule (H_2). These two hydrogen atoms share one electron each with each



other. Hence, each hydrogen atom shares one electron with the other atom to form a single covalent bond. They have a single shared pair of electrons.

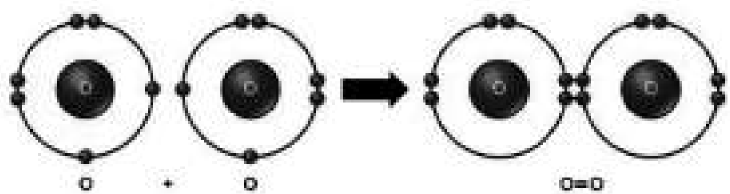
b. **Structure of nitrogen molecule (N_2)**

N_2 is a molecule of nitrogen. It is formed by the combination of two nitrogen atoms. Nitrogen has five electrons in its valence shell (from its electronic configuration 2,5). Thus, it needs three more electrons to form octet and maintain its stability. Each nitrogen atom, therefore, has to share three electrons so that these electrons can be shared by both the atoms. After sharing of electrons, nitrogen molecule forms a triple covalent bond at the region of the shared electrons. The structure of N_2 is shown in the figure



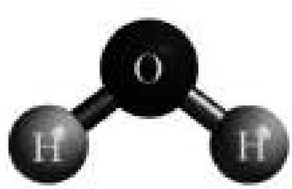
c. Structure of Oxygen molecule (O_2)

The oxygen molecule is formed by the combination of two oxygen atoms. The electronic configuration of oxygen is 2,6. Therefore, each oxygen atom needs two more electrons to complete their octet and become stable. Since, giving and taking electrons is not possible here, both the oxygen atoms share these two required electrons to form the bond. After sharing, a double covalent bond is formed between the oxygen atoms. The molecular structure of O_2 is given below.

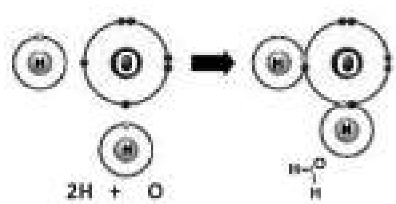


d. **Structure of Water molecule (H_2O)**

Water (H_2O) is formed by the chemical combination of two atoms of hydrogen and one atom of oxygen. The electronic configuration of hydrogen is 1 and that of oxygen is 2,6. When hydrogen and oxygen combine, hydrogen tries to form duplet and oxygen tries to form octet for stability. Since, giving and taking electrons do not make them duplet and octet, they share electrons. Each hydrogen atom shares its one electron with oxygen and the oxygen atom too shares its one each electron with each of the hydrogen atoms forming a single covalent bond amongst each other. The structure of water molecule is given below.



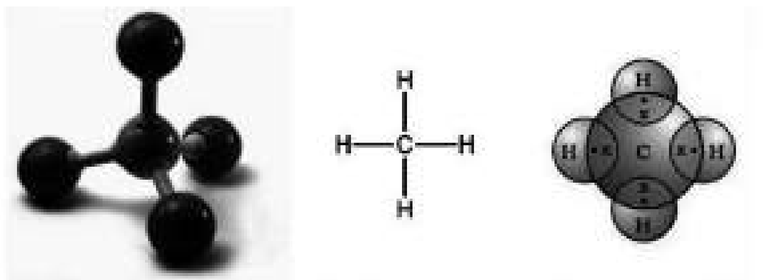
3-d structure of water



Formation of H_2O molecule

d. Structure of Methane molecule (CH_4)

The molecule of methane (CH_4) consists of one carbon atom surrounded by four hydrogen atoms. Carbon has four electrons in its valence shell and hydrogen has 1. Therefore, carbon needs four electrons to gain stability and each hydrogen needs one electron. As giving and taking is not possible, each hydrogen atom shares their one electron with the carbon atom sharing altogether four electrons with carbon. Similarly, carbon also shares one electron with each of the hydrogen atoms to form a single covalent bond. After the bond formation, there are four pairs of shared paired electrons.



3-D structure of CH_4 Bond in methane   2-d structure of CH_4

e. Structure of Ammonia molecule (NH_3)

Ammonia is composed of one atom of nitrogen and three atoms of hydrogen. The electronic configuration of nitrogen is 2,5 and that of hydrogen is 1. Each hydrogen atom needs one electron to become duplet while nitrogen atom needs three electrons to turn octet. So, nitrogen shares one each electron with three hydrogen atoms and hydrogen atoms also share their electrons with the nitrogen forming a single covalent bond between them. After sharing, there will be three pairs of shared electrons.

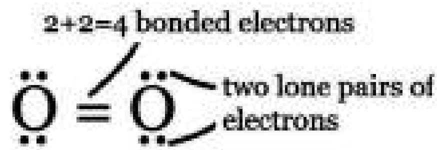
3-d structure of ammonia

Examples of Coordinate bonds

The structure of the ozone molecule (O_3) and sulphur trioxide molecule (SO_3) is discussed below. In these molecules, there is presence of coordinate bond. With the help of these examples, notice the difference between coordinate bond and normal covalent bond.

a. **Structure of Ozone molecule (O_3)**

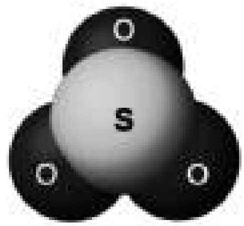
The ozone molecule consists of three oxygen atoms. It is also called trioxygen molecule. The electronic configuration of oxygen is 2,6. Each oxygen atom needs two more electrons to gain stability by achieving octet in its valence shell. The actual structure of ozone is angular with a O-O-O bond angle of 116.8^0. But, for simplicity, let us consider that the structure of ozone is planar and linear i.e. two-dimensional straight line that joins the three oxygen atoms together. The process of formation of ozone and its structure is discussed below:

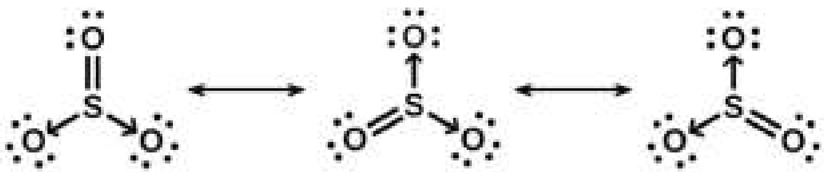


The ozone molecule is formed by the combination of oxygen molecule (O_2) with atomic oxygen (O). The two oxygen atoms in the molecule of oxygen make double covalent bond sharing two pairs of electrons. Each oxygen atom has two lone pairs of electrons. When one more oxygen atom comes to join with the oxygen molecule, one of the oxygen atoms from the oxygen molecule donates one lone pair of electrons to the third oxygen atom to complete octet of the third atom. So, in ozone molecule, there are two types of bonds. They are covalent bond and coordinate bond. But, the actual structure of the ozone molecule is angular. Hence, the required structure would be:

Structure of Sulphur trioxide molecule (SO_3)

The structure of sulphur trioxide is trigonal planar, i.e. it forms a two dimensional triangle. The angle between O-S-O in SO_3 is 120 degrees. In SO_3 molecule, one of the bonds of sulphur with oxygen should be a double bond. If this is the case, then S and O atoms satisfy their octets. Now, two oxygen atoms come to combine with SO. Here, the sulphur atom contributes two each electron to the bond with two oxygen atoms. Thus, two coordinate bonds are formed between two S-O atoms. Since, the sulphur can form double bond with any oxygen atom, the structure of SO_3 is also a resonant structure with three structural possibilities.





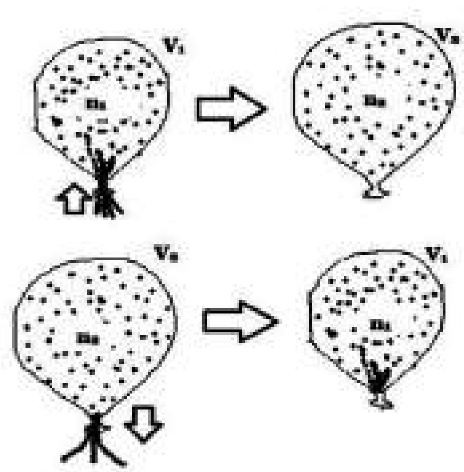
Avogadro's law

The Boyle's law gives the relationship between pressure and volume of a gas. Similarly, the Charles law gives the relationship between temperature and volume of a gas. In the same way, Avogadro's Law gives the relationship between volume of a gas with its amount (moles).

Avogadro's law states that, "U**nder constant temperature and pressure, equal volumes of all gases contain equal number of molecules irrespective of the physical and chemical properties of the gases**."

Mathematical derivation of Avogadro's law

Consider a flexible gas holder viz. a balloon which has a certain amount of gas (say $n_1$mol). The volume of the gas is V_1. Now, what happens to the volume of the air inside the balloon if more air is blown to it? Yes, the volume increases. Likewise, if we release the neck gently to let air out of the balloon, the volume decreases. Thus, the more air we blow into the balloon, the more is the volume occupied by the gas and vice versa. Thus, **the volume of a gas is directly proportional to the amount of substance (n).** i.e.



$V \propto n$

or, V = nk where k is a proportionality constant.

i.e. $\frac{V}{n} = k$

Let, us consider a sample of gas having volume V_1 and amount of gas n_1. When the amount of gas is increased in the sample to n_2, let the volume of the gas be V_2. Then, we know that,

$$\frac{V_1}{n_1} = k \quad \text{------------------- (i)}$$

and $$\frac{V_2}{n_2} = k \quad \text{------------------------(ii)}$$

Equating equations (i) and (ii),

$$\frac{V_1}{n_1} = \frac{V_2}{n_2}$$

or, $\mathbf{V_1 n_2 = V_2 n_1}$

Explanation of Avogadro's Law:

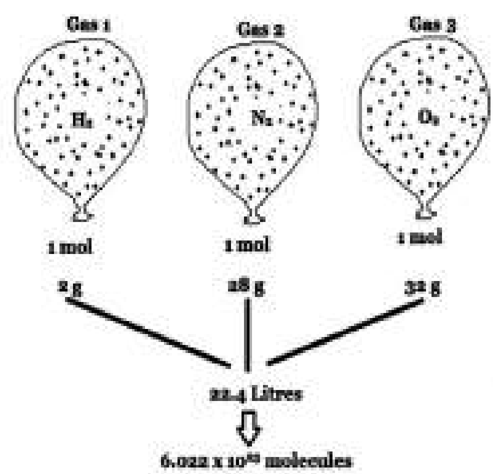
Consider three gases, viz. H_2, N_2 and O_2 at equal temperature and pressure. If we take one mole of all these gases, i.e. 2g of Hydrogen, 28 g of Nitrogen and 32 g of Oxygen respectively, at STP, all these three gases will have equal volume i.e. 22.4 litres and contain equal number of molecules, i.e. 6.022×10^{23}. The volume occupied by all ideal gases of 1 mol amount is called molar volume. One molar volume is equal to 22.4 Litres.

Using the relationship between volume and amount of substance for these three gases,

$$\frac{V_1}{n_1} = \frac{V_2}{n_2} = \frac{V_3}{n_3}$$

i.e. $$\frac{V_1}{1mol} = \frac{V_2}{1mol} = \frac{V_3}{1mol}$$

or, $V_1 = V_2 = V_3 = 22.4$ litres



The conclusion of this problem is:

a. 1 mol of Hydrogen gas = 22.4 L at STP = 1 g = 6.022×10^{23} molecules.

b. 1 mol of Oxygen gas = 22.4 L at STP = 32 g = 6.022×10^{23} molecules.

c. 1 mol of Nitrogen gas = 22.4 L at STP = 28 g g = 6.022×10^{23} molecules.

d. 1 mol of Carbon Dioxide gas = 22.4 L at STP = 44 g = 6.022×10^{23} molecules and so on.

Chemical Arithmetic regarding Avogadro's law

Example 1

1. **A sample of 10 L of a gas contains 1.3 mol. Find the amount of gas present in 20 L of that sample if pressure and volume are kept constant.**

   Solution:

   Given,

   Initial volume of the gas (V_1)= 10 L

   Initial amount of gas (n_1) = 1.3 mol

   Final volume of the gas (V_2) = 20 L

   Final amount of the gas (n_2) = ?

   According to Avogadro's law,

   $V_1 n_2 = V_2 n_1$

   or, 10 L x n_2 = 20 L x 1.3 mol

   or, $n_2 = \frac{20 \times 1.3}{10} mol$

   or, n_2 = 2.6 mol

   n_2 = 2.6 mol

   Therefore, the amount of the gas is 2.6 mol.

Example 2

2. **Calculate the volume of 5 mol of nitrogen gas at STP.**

   Solution:

   Amount of nitrogen gas (n) = 5 mol

   Volume of nitrogen (V) = ?

   According to Avogadro's Law,

   1 mol occupies the volume of 22.4 L.

   Hence, 5 mol occupies the volume of 22.4 x 5 L = 112 L at STP

   Therefore, the volume of 5 mol of nitrogen gas is 112 L.

Summary

1. The force that exists between atoms in a compound is called chemical bond.
2. The main types of chemical bonds are electrovalent or ionic bond, covalent bond and coordinate bond.
3. The type of bond which is formed between an electropositive and electronegative atom is called an ionic bond. They are formed by giving and taking of electrons.
4. The type of bond that is formed by sharing of electrons is called covalent bond.
5. Coordinate bond is a type of covalent bond between two atoms in which the bonding electrons are supplied by one of the two atoms.
6. Sigma bonds (σ bonds) are the strongest type of covalent chemical bonds which are formed by the head to head overlapping between atomic orbitals.
7. Pi bonds (π bonds) are those covalent bonds which are formed by side wise overlapping of the atomic orbitals. They are weaker bonds.
8. Avogadro's law is the law stating that equal volumes of all gases at the same temperature and pressure contain equal numbers of molecules.

Exercise

A. **Tick (√) the best alternatives.**

1. The type of bond between H and O in H_2O is
   a. Covalent bond b. Ionic bond
   c. Coordinate bond d. Both a and b
2. The type of bond in $AlCl_3$ is
   a. Covalent bond b. Ionic bond
   c. Coordinate bond d. None
3. The type of bond in O_3 is
   a. Covalent bond b. Ionic bond
   c. Coordinate bond d. Covalent with coordinate bond

4. Sigma bonds are formed by :
   a. Side wise overlapping of the orbitals
   b. Head wise overlapping of the orbitals
   c. No overlapping of orbitals
   d. Both a and b
5. In triple covalent bond, there is
   a. One sigma bond and one pi bond
   b. Two sigma bonds and two pi bonds
   c. One sigma bonds and two pi bonds
   d. Two sigma bonds and one pi bond

B. **Write very short answers to the following questions.**
   a. What is bonding?
   b. Define covalent bond.
   c. Write any two examples of covalent compounds.
   d. Name any two covalent bonds.
   e. What is coordinate bond?
   f. Write any two examples of coordinate compounds.
   g. Define sigma bond.
   h. What is pi bond?
   i. In between sigma and pi bond which one is stronger?

C. **Write short answers to the following questions.**
   a. Write any two differences between sigma and pi bonds.
   b. Differentiate between Electrovalent bond and covalent bonds
   c. Differentiate between Electrovalent compounds and covalent compounds
   d. Differentiate between Structure of NaCl and CH_4
   e. Define pi bonds with any two examples.
   f. Draw the molecular structure of NaCl.
   g. Sketch the structure of CH_4.
   h. Discuss the formation of bond in O_3 and SO_3.

i. Except inert gases other elements are chemically unstable. Why?

j. Why are inert gases chemically stable?

k. Electrovalent bond is also called ionic bond. Why?

l. Sodium chloride is an electrovalent compound. Why?

m. Water (H_2O) is a covalent molecule. Why?

n. Sulphur trioxide is a coordinate covalent compound. Why?

Activity

1. Take some toothpicks and some potatoes. Suppose toothpicks are bonding and potatoes are atoms. With the help of these things make model of single bond, double bond and triple bonds.
2. Make a model of ammonia, methane, carbon dioxide and water with the help of toothpicks and potatoes.

Glossary

Bond : the force of attraction which binds two or more elements together to make a stable chemical compound

Octet : the electronic configuration in which there are 8 electrons in the valence shell of an atom.

Duplet : the state of electronic configuration in which an atom has only one shell (K-shell) with 2 electrons in it.

Electrovalent bond : the chemical bond which is formed in between two opposite charges as a result of transfer of electrons

Covalent bond : the chemical bond which is formed by mutual sharing of electrons in between two or more non-metal atoms

Coordinate bond : the type of chemical bond in which one of the combining atoms contributes both of the shared electrons

Covalent compounds : the chemical compounds which are formed as a result of covalent bonding

Unit 10

Electrochemistry

Svante Arrhenisu (1859-1927)

Svante Arrhenius was born on February 19, 1859, in Vik, Sweden and died on October 2, 1927. Arrhenius propounded many theories in chemistry, astronomy, earth's ecology, etc. He described Arrhenius theory of ionization.

10.1 Introduction

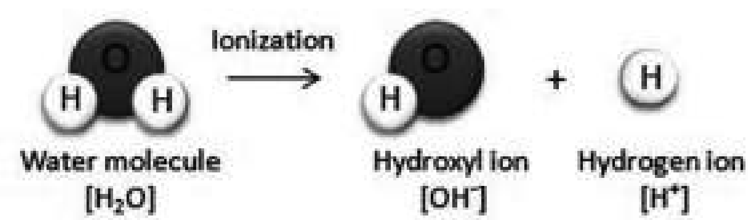
In various physical and chemical processes, electricity is either consumed or produced. Different types of chemical reactions occur in solution as well as in between electrodes. For example, chemical reactions in battery, acid base reaction, etc. Some chemical reactions also occur by passing electricity in the solution. For example electrolysis of water, electroplating, electrotyping, etc. These chemical reactions are studied in electrochemistry. Thus, **electrochemistry is a branch of physical chemistry which deals with the interaction between electrical energy and chemical change.** In this unit, we will describe the nature of water, pH and pOH and importance of neutralization reactions.

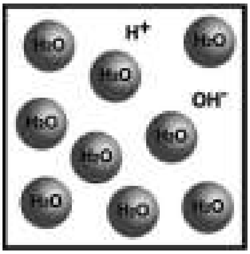
10.2 Ionic Product of Water

Pure water is a weak electrolyte. It undergoes self-ionization. In this process water molecule splits into hydrogen ion (H^+) and hydroxide ion (OH^-).

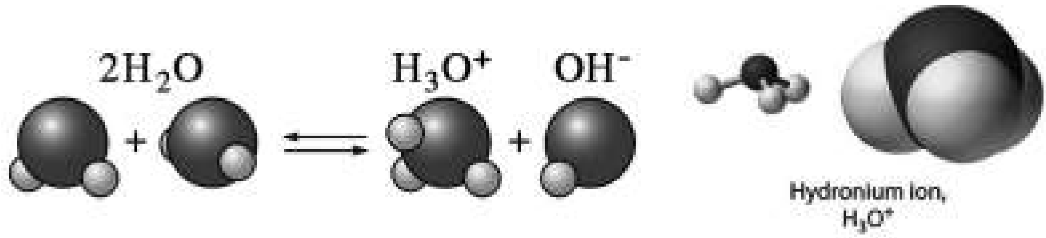
The equation can be shown as:

$H_2O \leftrightharpoons H^+ + OH^-$





As we know hydrogen ion is very reactive and it reacts further with water molecules to form hydronium ion(H_3O^+).



By applying the law of mass action, the ionization constant of water (K) can be given as:

$$K = \frac{[H^+][OH^-]}{[H_2O]}$$

Or, $[H^+][OH^-] = K[H_2O]$

Since dissociation takes place to a very small extent, the concentration of undissociated water molecules, $[H_2O]$, may be regarded as constant. Thus, the product of ionization constant of water and concentration of water $[H_2O]$ gives another constant which is termed as ionic product of water (K_w).

$$\boxed{[H^+][OH^-] = K_w}$$

Water undergoes self-ionization to give hydrogen ion and hydroxide ion. So, water can behave as both acid as well as base. Any substance which increases the concentration of hydrogen ion (H^+) would make water acidic. Similarly, any substance which increases the concentration of hydroxide (OH^-) ion, would make water basic. But, in pure water, the hydrogen ion concentration is always equal to the hydroxide ion concentration. While ionization, water molecule gives hydrogen ion and hydroxide ion in equal amount. Thus, **the product of molar concentration of hydrogen ion (H^+) and hydroxide ion (OH^-) produced by self-ionization of water at a particular temperature is called ionic product of water**. The value of ionic product of water (K_w) increases with the increase of temperature. It means that the concentration of H^+ and OH^- ions increases with increase in temperature.

Temperature (°C)	Value of K_w
0	0.11×10^{-14}
10	0.31×10^{-14}
20	0.86×10^{-14}
25	1.00×10^{-14}
40	2.91×10^{-14}
60	9.61×10^{-14}
100	7.50×10^{-14}

The value of K_w at 25°C is 1×10^{-14}. Since pure water is neutral in nature, H^+ ion concentration must be equal to OH^- ion concentration.

$[H^+] = [OH^-] = x$

or, $[H^+][OH^-]=x^2= 1 \times 10^{-14}$

or, $x = 1 \times 10^{-7}$ m/l

or, $[H^+] = [OH^-] = 1 \times 10^{-7}$ mole/ liter at 25°C temperature

This shows that at 25°C, in one liter water (approximately 55.5 moles), only 10^{-7} moles of water is in ionic form. When an acid or a base is added to water, the ionic concentration product, $[H^+][OH^-]$, remains constant, i.e., equal to K_w but concentrations of H^+ and OH^- ions do not remain equal. The addition of acid increases the hydrogen ion concentration while that of hydroxyl ion concentration decreases,

i.e., $[H^+] > [OH^-]$ (In acidic solution)

Similarly, when a base is added, the OH^- ion concentration increases while H^+ ion concentration decreases, i.e., $[OH^-] > [H^+]$ (In alkaline or basic solution)

In neutral solution, $[H^+] = [OH^-] = 1 \times 10^{-7}$ m/l

In acidic solution,

$$[H^+] > [OH^-]$$

or, $[H^+] > 1 \times 10^{-7}$ m/l and $[OH^-] < 1 \times 10^{-7}$ m/l

In alkaline solution,

$[OH^-] > [H^+]$

or $[OH^-] > 1 \times 10^{-7}$ m/l and $[H^+] < 1 \times 10^{-7}$ m/l

Thus, if the hydrogen ion concentration is more than 1×10^{-7} m/l, the solution will be acidic in nature and if less than 1×10^{-7} m/l, the solution will be alkaline.

$[H^+]$	Nature of Water
10^{-0}, 10^{-1}, 10^{-2}, 10^{-3}, 10^{-4}, 10^{-5}, 10^{-6}	Acidic
10^{-7}	Neutral
10^{-14}, 10^{-13}, 10^{-12}, 10^{-11}, 10^{-10}, 10^{-9}, 10^{-8}	Alkaline

We shall have the following table if OH^- ion concentration is taken into account.

$[OH^-]$	Nature of Water
0^{-14}, 10^{-13}, 10^{-12}, 10^{-11}, 10^{-10}, 10^{-9}, 10^{-8}	Acidic
10^{-7}	Neutral
10^{-0}, 10^{-1}, 10^{-2}, 10^{-3}, 10^{-4}, 10^{-5}, 10^{-6}	Alkaline

From the above discussion, it can be concluded that every aqueous solution, whether acidic, neutral or alkaline contains both H^+ and OH^- ions. The product of their concentrations is always constant, i.e., equal to 1×10^{-14} at 25°C. If one increases, the other decreases accordingly. So that, the product remains 1×10^{-14} m/l at 25°C.

For example, if $[H^+] = 10^{-2}$ m/l, then $[OH^-] = 10^{-12}$ m/l

Thus, the product of $[H^+][OH^-] = 10^{-2} \times 10^{-12} = 10^{-14}$, the solution is acidic.

Similarly, if $[H^+] = 10^{-10}$ m/l, then $[OH^-] = 10^{-4}$ m/l;

Thus, the product of $[H^+][OH^-] = 10^{-10} \times 10^{-4} = 10^{-14}$, the solution is alkaline.

10.3 pH and pOH of a solution

Generally, we see three types of solutions. They are acidic, basic and neutral. The nature of a solution, i.e., acidity, alkalinity or neutral can be expressed in terms of hydrogen ions. The concentration of hydrogen ion and hydroxyl ion is very less in the solution. So, it is expressed as a negative power to the base 10. These kinds of numbers are very difficult to use. To remove the difficulty, Sorensen in 1909 AD, introduced the popular term called pH and pOH. They are used to express acidic and basic strength of an aqueous solution.

pH and pH Scale

The acidic and basic strength of an aqueous solution can be measured in term of hydrogen ions concentration. Thus, **the negative logarithm of molar concentration of hydrogen ions is called pH**. From the definition,

$$pH = -\log[H^+]$$

As we have already discussed that, the molar concentration of hydrogen ions in pure water at 25°C is 1×10^{-7} m/l . Therefore, the pH of pure water can be expressed as,

$$pH = -\log[H^+]$$

$$pH = -\log[10^{-7}]$$

$$= 7$$

Thus, pH value of pure water is 7. It also indicates the neutral solution. If molar concentration of hydrogen ion is more than 10^{-7} m/l, or pH value less than 7, it indicates acidic solution. Similarly, if molar concentration of hydrogen ion is less than 10^{-7} m/l, or pH value more than 7, it indicates basic solution.

If pH = 7 (It indicates neutral solution)

$pH < 7$ (It indicates acidic solution)

$pH > 7$ (It indicates basic solution)

To express the nature of solution in terms of hydrogen ions concentration, a simple and convenient scale is introduced. It is called pH scale. Thus, t**he scale of hydrogen ions concentration which is used to express acidic and basic strength of an aqueous solution is called pH scale**. It is calibrated with the value from 1 to 14.

pH= 1	pH= 2	pH= 3	pH= 4	pH= 5	pH= 6	pH= 7	pH= 8	pH= 9	pH= 10	pH= 11	pH= 12	pH= 13	pH= 14

Acidic solution   Neutral solution   Basic solution

Example 1 : Calculate pH of a solution whose hydrogen ions concentration is 10^{-4} m/l.

Solution:

Hydrogen ions concentration = 10^{-4} m/l.

So, pH = $-\log[H^+]$

pH = $-\log[10^{-4}]$

= 4

Example 2 : Calculate pH of aqueous solution of nitric acid whose hydrogen ions concentration is 10^{-3} m/l.

Solution:

Nitric acid is a strong acid which undergoes complete ionization to give 10^{-3} m/l hydrogen ions concentration.

So, hydrogen ions concentration = 10^{-3} m/l.

pH = $-\log[H^+]$

pH = $-\log[10^{-3}]$

= 3

pOH and pOH Scale

Like in pH scale, the acidic and basic strength of an aqueous solution can also be measured in terms of hydroxyl ions concentration (OH^-). Thus, **the negative logarithm of molar concentration of hydroxyl ions is called pOH**. From the definition,

pOH = $-\log[OH^-]$

As we discussed in pH and pH scale, the molar concentration of hydroxyl ions in pure water at 25°C is 1×10^{-7} m/l . Therefore, the pOH of pure water can be expressed as:

pOH = $-\log[OH^-]$

pOH = $-\log[10^{-7}]$

= 7

Thus, pOH value of pure water is 7. This value also indicates the neutral solution. If molar concentration of hydroxyl ions is more than 10^{-7} m/l, or pOH value less than 7, it indicates basic solution. Similarly, if molar concentration of hydroxyl ions is less than 10^{-7} m/l, or pOH value more than 7, it indicates acid solution.

If pOH = 7 (It indicates neutral solution)

pOH< 7 (It indicates basic solution)

pOH> 7 (It indicates acidic solution)

To express the nature of solution in terms of hydroxyl ions concentration, a simple and convenient scale is introduced. It is called pOH scale. Thus, **the scale of hydroxyl ions concentration which is used to express acidic and basic strength of an aqueous solution is called pOH scale**. It is calibrated with the value from 1 to 14.

pOH =1	pOH =2	pOH =3	pOH =4	pOH =5	pOH =6	pOH =7	pOH =8	pOH =9	pOH =10	pOH =11	pOH =12	pOH =13	pOH =14

Basic solution Neutral solution Acid solution

Example 1 : Calculate pOH of a solution whose hydroxyl ions concentration is 10^{-2} m/l.

Solution:

Hydroxyl ions concentration = 10^{-2} m/l.

So, pOH $= -\log[H^+]$

pOH $= -\log[10^{-4}]$

$= 4$

So, it is a basic solution.

Example 2 : Calculate pOH of aqueous solution of sodium hydroxide whose hydroxyl ions concentration is 10^{-1} m/l.

Solution:

Sodium hydroxide is a strong babe. It undergoes complete ionization to give 10^{-1} m/l hydroxyl ions concentration.

So, hydroxyl ions concentration = 10^{-1} m/l.

pOH $= -\log[OH^-]$

pOH $= -\log[10^{-1}]$

$= 1$

So, it is a basic solution.

Relation between pH and pOH

For pure water, the ion product of water is given as

$[H^+]\,[OH^-] = K_w$

We know that the value of K_w is 10^{-14} at 25°C.

So, taking the -log on both sides of the equation,

$- \log [H^+] + (- \log [OH^-]) = - \log [10^{-14}]$

$\boxed{pH + pOH = 14}$

This is a relation of pH and pOH.

So, for pure water the sum of pH and pOH is equal to 14. If a solution has pH value two than the pOH value will be 12.

Example 1: Calculate pOH of a solution whose hydrogen ions concentration is 10^{-2} m/l.

Solution:

Hydrogen ions concentration = 10^{-2} m/l.

According to the formula of pH,

$pH = - \log [H^+]$

$pH = - \log [10^{-2}]$

PH = 2

Now, pH + pOH = 14

2 + pOH = 14

pOH = 12

The solution has pH value less than 7 or pOH value more than 7. So, it is an acidic solution.

Example 2 : Calculate pOH and pH value of aqueous solution of potassium hydroxide whose hydroxyl ions concentration is 10^{-2} m/l.

Solution:

Potassium hydroxide is a strong babe. It undergoes complete ionization to give 10^{-2} m/l hydroxyl ions concentration.

So, hydroxyl ions concentration = 10^{-2} m/l.

$pOH = -\log[OH^-]$

$pOH = -\log[10^{-2}]$

= 2

Now, pH + pOH = 14

pH + 2 = 14

pH = 12

So, it is a basic solution as it has pH value more than 7 and pOH value less than 7.

10.4 pH meter

Whether we are working in a chemistry laboratory or in our garden, sometimes we need to find out the exact value of pH of the chemical substances or soil. It is also required to measure the acidity or alkalinity of the certain liquids at our home or laboratory. To find their pH value we use pH paper. During pH calculation through pH paper, the colour change in pH paper is matched with the colour chart of the pH scale. The matching of colour of pH paper with pH colour chart may not give exact result. So, we use a scientific device to measure the pH value of the solution. It is called pH meter. Thus, **the scientific device which is used to measure the exact pH value of a solution without colour matching in the pH chart is called pH meter.**

10.5 Neutralization reaction

When we keep an acid and a base together, the acid loses its acidic properties and the base also loses its basic properties to give neutral products. It is called neutralization reaction. Generally, in neutralization reaction, an acid and a base react together to form salt and water. Actually, neutralization involves the combination of H^+ ions and OH^- ions to generate water. The products of neutralization reaction of a strong acid and a strong base have a pH value equal to 7. The neutralization products of a strong acid and a weak base have pH value less than 7. Similarly, when a strong base reacts with a weak acid, it gives the products having pH value greater than 7.

The most common strong acids are HCl, H_2SO_4, HNO_3, etc. Similarly, the strong bases are LiOH, NaOH, KOH, RbOH, $Ca(OH)_2$, $Sr(OH)_2$, $Ba(OH)_2$, etc.

The neutralization reaction between strong acid and strong base

Let us observe the reaction between hydrochloric acid and sodium hydroxide.

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

In the above reaction, hydrochloric acid is a strong acid and sodium hydroxide is a strong base. When they are mixed together, they give salt and water. The actual combination of ions is as follows.

$$H^+ + Cl^- + Na^+ + OH^- \leftrightharpoons NaCl + H_2O$$

When a strong acid and a strong base fully neutralize, the pH of the resultant product is 7. At this point of neutralization, there are equal amounts of hydrogen and hydroxyl ions. There is no excess amount of NaOH and HCl. Therefore, when a strong acid completely neutralizes a strong base, the pH of the salt solution will always be 7.

The neutralization reaction between weak acid and weak base

Let us observe the reaction between carbonic acid and ammonium hydroxide.

$H_2CO_3(aq) + NH_4OH(aq) \leftrightharpoons (NH_4)_2CO_{3(aq)} + H_2O(l)$

In the above reaction, carbonic acid is a weak acid and ammonium hydroxide is a weak base. When they are mixed together, they give salt and water. The actual combination of ions is as follows.

$2H^+ + CO_3^{--} + NH_4^+ + OH^- \leftrightharpoons (NH_4)_2CO_{3(aq)} + H_2O(l)$

When a weak acid and a weak base fully neutralize, the pH of the resultant product is 7. At this point of neutralization, there are equal amounts of hydrogen and hydroxyl ions. There is no excess amount of NH_4OH and H_2CO_3. Therefore, when a weak acid completely neutralizes a weak base, the pH of the salt solution will always be 7.

Summary of the neutralization reactions

Acid	Base	Products	pH value
Strong acid	Strong base	Neutral salt and water	7
Strong acid	Weak base	Acidic salt and water	Less than 7
Weak acid	Strong base	Basic salt	More than 7
Weak acid	Weak base	Neutral salt and water	7

Applications of the neutralization reaction

Soil Test

We grow different kinds of crops in the soil. All kinds of crops cannot be grown in the same soil. This is because different kinds of crops need different kinds of nutrients and a certain pH value for their healthy growth. It would be better to grow the crops after knowing the concentration of the nutrients and pH value of the soil. After continuous growing of the same type of crop as well as due to different kinds of environmental pollutions, such as acid rain, mixing of different kinds of chemicals, chemical fertilizers, etc. the concentration of the nutrients as well as the pH value of the soil gets changed. Therefore, it is necessary to test the soil before sowing the crops.

pH of the Soil

The pH value of a soil is the measure of its alkalinity or acidity based on pH scale from 1 to 14. Zero represents severe acidity, fourteen is extreme alkalinity and seven represents neutral. The pH of the cropland soil should be in the range of 6.0 - 7.5. If pH becomes imbalanced, it can directly affect the availability of nutrients in the soil. The derived products of limestone are applied to the soil to reduce the acidity (sour) of the soil. Similarly, alkaline (sweet) soil requires application of sulphur products to reduce its alkalinity.

a. Neutralization of the soil

Neutralization of soil is necessary in order to promote the growth plants. The ability of plants to take nutrients from the soil into their roots is affected by the pH value of the surrounding soil. Acid rain can cause soil to become acidic. The acid present in the soil must be neutralized with lime, or calcium oxide (CaO). It is called liming of the soil. Some plants get benefit from the liming soil. In the liming process, agricultural limestone-calcium carbonate ($CaCO_3$) that may also contain magnesium carbonate ($MgCO_3$) neutralizes acid in the soil. Soil may be too basic or alkaline, especially in areas where there is little precipitation. Substances that can act as acids such as calcium sulphate, it is also called gypsum ($CaSO_4$), and sulphur (S_2) can be applied to the soil for neutralization.

b. Treatment of Hyperacidity

The hyperacidity in stomach of human beings is caused due to the excess secretion of hydrochloric acid in stomach. It can cause heart burns, sour fluid vomiting in the mouth and bombast. It can be neutralized by using bases present in the tablets called antacids. Antacids are a group (class) of medicines which help to neutralise the acid content of our stomach. They include aluminium hydroxide, magnesium hydroxides, magnesium carbonate, magnesium trisilicate, etc. These antacids come in various brand names and are available as tablets and liquids. Some antacids are combined with another medicine called simeticone which helps to reduce wind (flatulence).

c. Treatment of insect stings:

The sting of bee injects venom in our body which is an acidic. This acid can be neutralized by using chemicals which have basic properties such as sodium bicarbonate or baking soda. Likewise, wasp sting inject alkaline venom in our body. So, baking soda does not work in this situation. Instead, acids should be used to neutralize this alkali. The use of weak acids like vinegar can relieve the pain.

Summary

1. Electrochemistry is a branch of physical chemistry which deals with the interaction between electrical energy and chemical change.
2. Pure water is a weak electrolyte. It undergoes self-ionization. In this process water molecule splits into hydrogen ion (H^+) and hydroxide ion (OH^-).
3. The product of molar concentration of hydrogen ion (H^+) and hydroxide ion (OH^-) produced by self-ionization of water at a particular temperature is called ionic product of water.
4. The value of ionic product of water (K_w) increases with the increase of temperature. It means that the concentration of H^+ and OH^- ions increases with increase in temperature.
5. The value of K_w at 25°C is 1×10^{-14}. Since pure water is neutral in nature, H^+ ion concentration must be equal to OH^- ion concentration.
6. The nature of a solution, i.e., acidity, alkalinity or neutral can be expressed in terms of hydrogen ions.
7. The negative logarithm of molar concentration of hydrogen ions is called pH.
8. If molar concentration of hydrogen ion is more than 10^{-7} m/l, or pH value less than 7, it indicates acidic solution.
9. If molar concentration of hydrogen ion is less than 10^{-7} m/l, or pH value more than 7, it indicates basic solution.
10. The scale of hydrogen ions concentration which is used to express acidic and basic strength of an aqueous solution is called pH scale.
11. The negative logarithm of molar concentration of hydroxyl ions is called pOH.
12. If molar concentration of hydroxyl ions is more than 10^{-7} m/l, or pOH value less than 7, it indicates basic solution.
13. If molar concentration of hydroxyl ions is less than 10^{-7} m/l, or pOH value more than 7, it indicates acid solution.
14. The scale of hydrogen ions concentration which is used to express acidic and basic strength of an aqueous solution is called pH scale.
15. For pure water, the sum of pH and pOH is equal to 14. If a solution has pH value five than the pOH value will be ten.
16. The scientific device which is used to measure the exact pH value of a solution without colour matching in the pH chart is called pH meter.

17. When we keep an acid and a base together, the acid loses its acidic properties and the base also loses its basic properties to give neutral products. It is called neutralization reaction.
18. In the liming process, agricultural limestone, i.e., calcium carbonate ($CaCO_3$) that may also contain magnesium carbonate ($MgCO_3$) neutralizes acid in the soil.
19. Substances that can act as acids such as calcium sulphate or gypsum ($CaSO_4$), and sulphur (S_2) can be applied to the soil for neutralization.
20. Antacids are a group (class) of medicines which help to neutralise the acid content of our stomach.
21. Antacids include aluminium hydroxide, magnesium hydroxides, magnesium carbonate, magnesium trisilicate, etc.

Exercise

A. Tick (√) best alternatives from the followings.

1. The ionic product of pure water is:
   - a. 10^{-7}mol/l
   - b. 10^{-14}mol/l
   - c. 10^{-10}mol/l
   - d. 10^{-5}mol/l
2. pH is the measure of:
   - a. Percentage of hydrogen atoms
   - b. Hydrogen and hydroxyl ions
   - c. Hydrogen ion concentration
   - d. Percentage of water molecules
3. pOH is the measure of:
   - a. Percentage of hydrogen atoms
   - b. Hydrogen and hydroxyl ions
   - c. Hydrogen ion concentration
   - d. Hydroxyl ion concentration
4. Pure water is:
   - a. Slightly acidic
   - b. Slightly alkaline
   - c. Neutral
   - d. Highly acidic
5. Which is an antacid?
   - a. Sodium chloride
   - b. Lactic acid
   - c. Citric acid
   - d. Magnesium hydroxide

B. Give very short answers to the followings.

1. What is ionic product of water?
2. Define pH and pOH.
3. What is pH range? Show with the pH scale.
4. Write down the value of hydrogen ion concentration in pure water.
5. Write down the equation for the ionic product of water.
6. What is pH scale? Write down the pH value of pure water.
7. Define neutralization with the help of example.
8. What will be the pH value when strong acid reacts with weak base?
9. Write down any two applications of the neutralization in our daily life.
10. What are antacids? Write down any two examples.

C. Give short answers to the following questions.

1. Draw a pH scale showing acid, base and neutral range.
2. Calculate the pH of 10^{-3} molar solution of HCl.
3. If the pH of a solution is 4, what is its pOH?
4. What is pOH scale? Show with pOH scale.
5. What is pH meter? Why is it better than pH paper?
6. What treatment can be done in insect stings?

D. Give long answers to the following questions.

1. What is hyperacidity? Describe the process to reduce it.
2. What is soil pH? What can be done to neutralize it?
3. Calculate pOH and pH value of aqueous solution of potassium hydroxide whose hydroxyl ions concentration is 10^{-3} m/l.
4. Calculate pH and pOH of a solution whose hydrogen ions concentration is 10^{-4} m/l.
5. Calculate pOH and pH of a solution whose hydroxyl ions concentration is 10^{-3} m/l.
6. Calculate pOH and pH of aqueous solution of calcium hydroxide whose hydroxyl ions concentration is 10^{-3} m/l.

Activity

1. Collect sample of soil from your surroundings and calculate its pH value using pH paper, and pH colour chart.
2. Collect a citrus fruit and extract juice from it. With help of pH paper and pH colour chart find out its pH value.
3. Make an aqueous solution of detergent and find out its pH value with the help of pH paper and pH colour chart.

Glossary

pH	:	the negative logarithm of molar concentration of hydrogen ions
pOH	:	the negative logarithm of molar concentration of hydroxyl ions
Hyperacidity	:	acidity more than normal in the stomach due to HCl
Liming	:	use of calcium oxide in the soil
Aqueous solution	:	the solution prepared in distilled water

Unit 11

Organic Chemistry


V.V. Markovnikov (1838-1904)

Vladimir Vasilyevich Markovnikov was born 1838 AD in Nizhny Novgorod, Russia and died on February 1904, in Moscow. He formulated a theory called Markovnikov addition theory. It is used in addition of hydrogen halides to the carbon-carbon double bond.

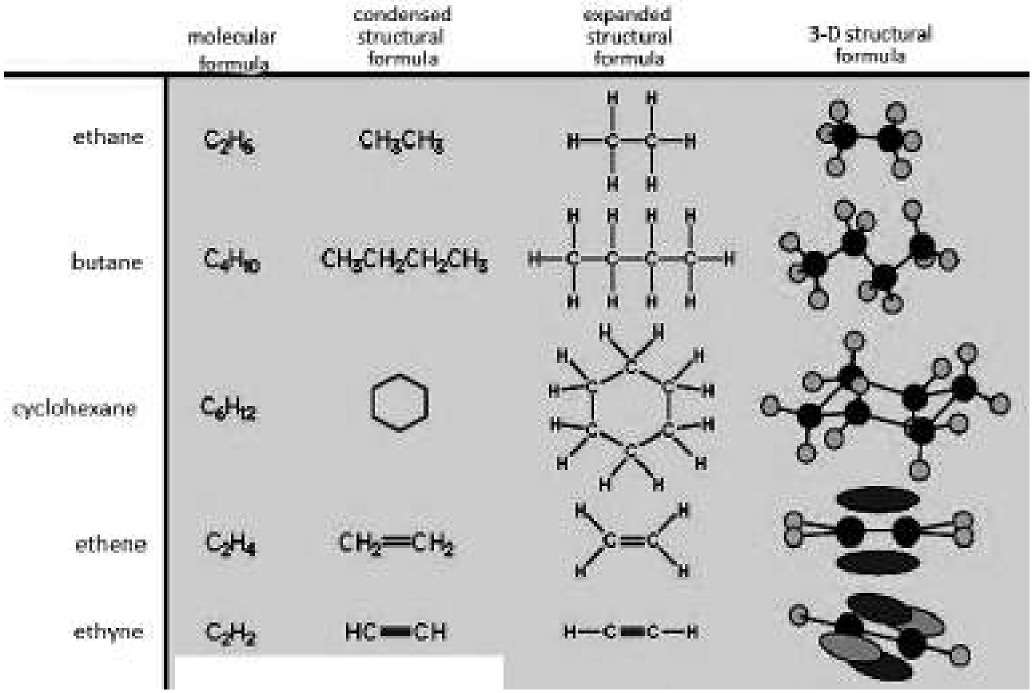
11.1 Introduction

The branch of chemistry in which we study about structure, properties, composition, reactions, and preparation of carbon-containing compounds is called organic chemistry. It includes not only hydrocarbons but also compounds with any number of other elements, like, nitrogen, oxygen, halogens, phosphorus, silicon, sulphur, etc. Organic chemistry was originally limited to the compounds which were produced by the living organisms. But, nowadays it has been extended to the chemical compounds which are made by human beings, such as monomers, polymers, plastics, drugs, etc. The range of application of organic compounds is enormous. It includes pharmaceuticals, petrochemicals, food, explosives, paints, cosmetics etc. In this unit we will discuss about different kinds of organic reactions including nucleophilic and electrophilic reactions.

11.2 Organic compounds

The chemical compounds which contain carbon as one of the elements are called organic compounds. For example, methane, ethane, alcohol, ether, carbohydrates, proteins, etc. It includes vast number of chemical compounds in which one or more atoms of carbon are covalently linked to the atoms of other elements, like hydrogen, nitroge, sulphur, etc. There are few carbon containing compounds which are not classified into organic compounds. For example, carbides, carbonates, cyanides etc. The chart below shows some of the organic compounds along with their molecular formula, structural formula, condensed formula and the three dimensional structure.

Organic chemistry is a highly creative science in which chemists create new molecules and explore the properties of existing compounds. It is the most popular field of study for chemists, scientists, researchers, biotechnologists, doctors, etc.



Uses of organic compounds

Organic compounds are very important to us. They are the pillars to the economic growth of the country. Economic growth becomes fast if the country can manufacture different kinds of organic compounds like rubber, plastics, fuel, pharmaceutical, cosmetics, detergent, coatings, dyes, agrochemicals, etc. On the basis of study of organic compounds, the organic chemistry has many branches, such as biochemistry, biotechnology, medicine; pharmaceutical, etc. Different kinds of organic compounds are present in the body of living beings. Such as carbohydrates, proteins, vitamins, hormones, etc. Absence of one or more organic compounds in the body leads serious problem in the health. So, we can say that organic compounds are making human life easy, comfortable and luxurious.

Organic Reactions

Those chemical reactions which involve organic compounds are called organic reactions. The basic organic reactions are addition reactions, elimination reactions, substitution reactions, pericyclic reactions, rearrangement reactions, photochemical reactions, redox reactions, etc. Organic reactions are used to synthesize new organic molecules. The production of many man-made chemicals such as drugs, plastics, food additives, fabrics depend up on organic reactions.

The oldest organic reactions are combustion of organic fuels and saponification of fats to make soap. But, the modern organic chemistry starts with the Wohler synthesis. In the history of the Nobel Prize in Chemistry, the awards have been given for the invention of specific organic reactions such as Grignard reaction , Diels-Alder reaction, Wittig reaction , olefin metathesis, etc . In this unit we will discuss some basic organic reactions. They are:

i. Substitution reaction

ii. Addition reaction

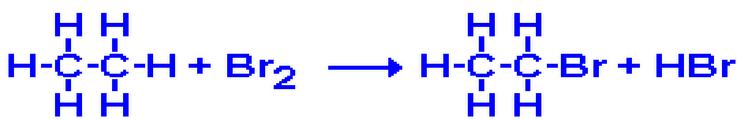
iii. Elimination reaction

i. **Substitution reactions**

Those chemical reactions in which an atom or groups of atoms in a molecule are replaced by another atoms or group of atoms are called substitution reactions. In substitution reaction, the saturation and unsaturation of the organic compounds do not change. For example,

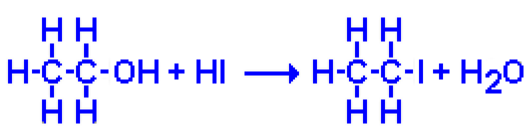
Example: 1

The gas ethane (CH_3CH_3) reacts with bromine vapour in presence of light to form bromoethane (CH_3CH_2Br) and hydrogen bromide (HBr). In this process, a hydrogen atom in ethane has been substituted by a bromine atom. So, reaction of ethane with bromine vapour is an example of substitution reaction.



Example: 2

Ethanol (CH_3CH_2OH) reacts with hydrogen iodide (HI) to form iodoethane (CH_3CH_2I) and water. Here, a group of atoms (i.e. OH), has been replaced by an iodine atom.

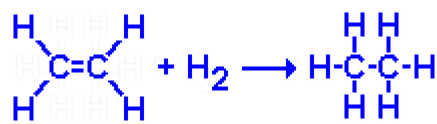


2. **Addition reactions**

Those chemical reactions in which a molecule reacts with organic molecule (which has one or more multiple covalent bonds) to form a product molecule are called addition reactions. After addition reaction, molecular mass becomes equal to the sum of molecular mass of the reacting molecules. In addition reactions, the saturation of the organic molecules decreases. It means that alkyne becomes alkene and alkene becomes alkane.

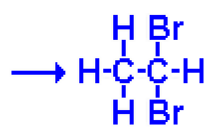
Example 3

Ethene ($CH_2{=}CH_2$) has a double covalentbond. It combines with one hydrogen molecule (H_2) to given ethane (CH_3CH_3). In this reaction two hydrogen atoms combine with two different carbon atoms in presence of platinum as a catalyst.



Example 4

Ethyne (C_2H_2) has a triple covalent bond. It combines with two molecules of hydrogen bromide to give 1,1 dibromoethane (CH_3CHBr_2). In this reaction, two reactant molecules are combined together to give a final product. So, it is a kind of addition reaction.

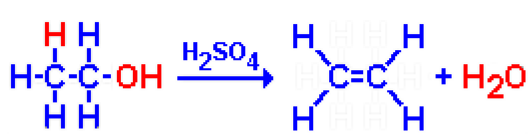
H–C ≡ C – H + 2HBr 

3. **Elimination reactions**

Those chemical reactions in which atoms or group of atoms are eliminated from the organic molecule to give unsaturated product molecule are called elimination reactions. After elimination reaction, the compound with multiple covalent bonds is formed. It has less molecular mass than the parent organic molecule.

Example 5

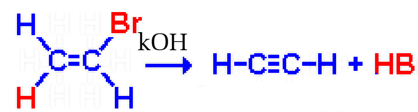
Ethanol (CH_3CH_2OH), when mixed with concentrated sulphuric acid (H_2SO_4), it loses one hydrogen atom and one hydroxyl group. As a result ethene ($CH_2{=}CH_2$) and water are formed.



Example 6

Bromoethene ($CH_2{=}CHBr$) when reacts with potassium hydroxide, it loses one hydrogen atom and one bromine atom. As a result, ethyne (C_2H_2) and hydrogen bromide are obtained.

11.3 Electrophiles and Nucleophiles



We know that opposite charges attract each other and like charges repel each other. Similarly, in chemical reactions, electrons flow from "electron rich" area to "electron poor" area. The special name is given to the types of chemical species that are considered "electron rich" and "electron poor". They are called **nucleophiles** and **electrophiles**.

Nucleophiles

Nucleophile is a Greek word. It means "nucleus loving", or "positive-charge loving". They attack on the region of low electron density of the reactant molecule. Thus, **a nucleophile is an atom or group of atoms that provides a pair of electrons to form a new covalent bond**. Nucleophiles can donate an electron pair to an acceptor compound because they have either negative charge or a lone pair of electrons or multiple covalent bonds. The properties of nucleophiles due to which they donate electrons are called nucleophilicity. Nucleophiles are either negatively charged species or neutral molecules with one pair of available electrons for donation. They are generally denoted by the symbol "Nu". There are two types of nucleophiles. They are:

i. **Negatively charged nucleophiles:**These nucleophiles have negative charge on them. For example,

Name of negative nucleophiles	Symbols
Hydroxide ion	OH^-
Chloride ion	Cl^-
Bromide ion	Br^-
Iodide ion	I^-
Cyanide ion	CN^-
Oxide ion	O^{2-}
Sulphide ion	S^{2-}
Nitride ion	N^{3-}
Alkoxide ion	RO^-

ii. **Neutral nucleophiles:** These are neutral molecules which have a pair electrons for donation. For example,

Name of neutral nucleophiles	**Symbols**
Water	H_2O
Ammonia	NH_3
Alcohol	R-OH
Ether	R-O-R
Amines	$R\text{-}NH_2$

Electrophiles

Electrophile is a Greek word. It means "electrons loving", or "negative-charge loving". They attack on the region of high electron density of the reactant molecule. Thus, a**n electrophile is an atom or group of atoms that accepts a pair of electrons to form a new covalent bond**. Electrophile can accept an electron pair from the donor compound because they have either positive charge or neutral molecules with electrons deficient centers. The property of an electrophile due to which it accepts electrons is called electrophilicity. Electrophiles are either positively charged species or neutral molecules with electron deficient centers. They are generally denoted by the symbol "E". There are two types of electrophiles. They are:

i. **Positively charged electrophiles:** These electrophiles have positive charge on them. For example,

Name of positively charged electrophiles	**Symbols**
Hydronium ion	H_3O^+
Hydrogen ion	H^+
Ammonium ion	NH_4^+
Nitronium ion	NO_2^+
Carbonium ion or carbocation	CH_3^+

ii. **Neutral electrophiles:** These electrophiles have electron deficient centers. For example,

Name of neutral electrophiles	Symbols
Sulphur trioxide	SO_3
Boron trifluoride	BF_3
Boron trichloride	BCl_3+

11.4 Markonikov's rule

This is an empirical rule based on Markonikov's experimental observations on the addition of hydrogen halides to alkenes. The rule states that "***when an unsymmetrical alkene reacts with a hydrogen halide to give an alkyl halide, the hydrogen adds to the carbon of the alkene that has the greater number of hydrogen atoms and the halogen to the carbon of the alkene with the fewer number of hydrogen atoms***". Alternatively, the rule can be stated that the hydrogen atom is added to the carbon with the greatest number of hydrogen atoms while the X (halogen) component is added to the carbon with the least number of hydrogen atoms.

Example: When propene reacts with hydrogen bromide, it gives 2-bromo propane. In this reaction, we do not get 1-bromo propane. According to Markovnikov's rule, first of all hydrogen bromide breaks down into hydrogen ion and bromide ion. Hydrogen ion goes to the first number of carbon where the number of hydrogen atoms are two. At the same time bromide ion goes to the second number of carbon where there is only one hydrogen atom.

$$CH_3\text{-}CH{=}CH_2 \ + \ HBr \longrightarrow CH_3\text{-}CHBr\text{-}CH_3$$

11.5 Saytjeff's rule:

Saytzeff rule also known as Zaitsev Rule. It predicts the alkene products formed after elimination reaction. According to this rule, in an elimination reaction, the less hindrances and more stable products are favoured over less stable and more hindrance products. Thus, Saytzeff Rule states that **the most substituted product is more stable and is favoured over other products**. It occurs in different elimination reactions.

Alkyl groups such as methyl, ethyl, propyl, etc. are electron donating groups. These groups increase the electron density on the pi-bond of the alkene. As these groups are larger in size, they create more hindrance by making the alkene less stable.

Example:

Reaction (i) $CH_3\text{-}CH_2\text{-}CH_2\text{–}CH_2\text{-}CHBr\text{–}CH_3 \rightarrow CH_3\text{-}CH_2\text{-}CH{=}CH\text{-}CH_3$ (2-Penene) It is a major product.

Reaction (ii) CH_3-CH_2-CH_2–CH_2-CHBr–CH_3 $\rightarrow$ CH_3-CH_2-CH_2-CH=CH_2 (1-Penene)
It is a minor product

In the above examples, when hydrogen and bromine are removed from 2-bromopentane, two products are formed. They are 2-Pentene and 1-Pentene. According to Saytzeff Rule,2-Pentene is more possible than1-Pentene. This is because 2-Pentene is less sterically hindered as compared to 1-Pentene.

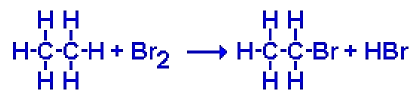
Summary

1. The branch of chemistry in which we study about structure, properties, composition, reactions and preparation of carbon-containing compounds is called organic chemistry.
2. The chemical compounds which contain carbon as one of the elements are called organic compounds.
3. The examples of organic compounds are protein, fat, vitamins, carbohydrates, methane, ethane, etc.
4. Those chemical reactions which involve organic compounds are called organic reactions.
5. The basic organic reactions are addition reactions, elimination reactions, substitution reactions, pericyclic reactions, rearrangement reactions, photochemical reactions, redox reactions, etc.
6. Those chemical reactions in which an atom or groups of atoms in a molecule are replaced by another atoms or group of atoms are called substitution reactions.
7. Those chemical reactions in which a molecule reacts with organic molecule (which has one or more multiple covalent bonds) to form a product molecule are called addition reactions.
8. Those chemical reactions in which atoms or group of atoms are eliminated from the organic molecule to give unsaturated product molecule are called elimination reactions.
9. Nucleophile is an atom or group of atoms that provides a pair of electrons to form a new covalent bond.
10. An electrophile is an atom or group of atoms that accepts a pair of electrons to form a new covalent bond.
11. When an unsymmetrical alkene reacts with a hydrogen halide to give an alkyl halide, the hydrogen adds to the carbon of the alkene that has the greater number of hydrogen atoms and the halogen to the carbon of the alkene with the fewer number of hydrogen atoms".
12. Saytzeff Rule states that the most substituted product is more stable and is favoured over other products.

Exercise

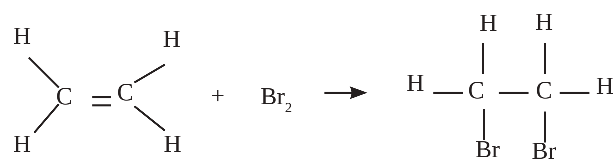
A. **Tick (√) the best alternatives from the followings.**

1. The given is an example of ………………… reaction.



   a. Elimination reaction
   b. Substitution reaction
   c. Addition reaction
   d. Decomposition reaction

2. What type of chemical reaction is given in the example?



   a. Elimination reaction
   b. Substitution reaction
   c. Addition reaction
   d. Neutralization reaction

3. A nucleophile is a reactant that …………..a pair of electrons to form a new covalent bond.
   a. Gives b. Bonds
   c. Accepts d. Pairs

4. An electrophile is a species that …………..a pair of electrons to form a new covalent bond.
   a. Gives b. Bonds
   c. Accepts d. Pairs

5. According to the Sayt Jeff's rule, the most substituted product ……………. .
   a. is less stable b. is not formed in the reaction
   c. is more stable d. has more hindrance

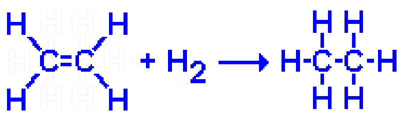
B. **Give very short answers to the following questions.**

1. Define organic chemistry.
2. What are organic compounds? Write down any two examples.
3. Define substitution reaction.
4. What types of reactions are called addition reaction?
5. Define an elimination reaction.
6. What are nucleophiles?

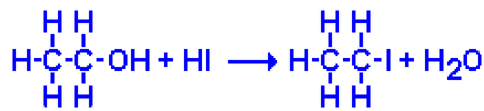
7. State Markonikov's rule.
8. Define electrophiles.
9. State Saytzeff Rule.
10. What kinds of compounds are called most substituted compounds?

C. **Give short answers to the following questions.**

1. Write the type of reactions with reason.



(i)



(ii)

2. Write any two differences between electrophilic substation and electrophilic addition reactions.
3. State any two differences between electrophiles and nucleophiles.
4. Which rules do the given reactions below display? State the rule too.

   (i) $CH_3\text{-}CH_2\text{-}CH_2\text{–}CH_2\text{-}CHBr\text{–}CH_3 \rightarrow CH_3\text{-}CH_2\text{-}CH{=}CH\text{-}CH_3$

   (ii) $CH_3\text{-}CH{=}CH_2 + HBr \rightarrow CH_3\text{-}CHBr\text{-}CH_3$
5. What are positive and neutral electrophiles? Explain with examples.
6. Define positive and neutral electrophiles with the help of examples.

D. **Give long answers to the following questions.**

1. Define addition reactions with the help of examples.
2. What are substitution reactions? Explain with examples.
3. What types of reaction are called elimination reactions? Describe with the help of reactions.
4. Explain Markovnikov's rule and Saytjeff's rule with the help of chemical reactions.

Project work

1. Observe in your home, surrounding and school and make a list of different kinds of organic compounds present there.
2. Make a list of organic and inorganic compounds present in your home, surrounding and school. Also segregate them into edible and non-edible compounds.

Unit 12

Metals and Metallurgy


***Georgius Agricola* (1495-1555)**

Georgius Agricola was born on 1495 AD and died on 1555 AD. He was the pioneer chemist who studied about the earth, rocks, minerals and fossils. He was the first person who studied about metallurgy and mineralogy.

12.1 Aluminium

Introduction

Name	Aluminium
Symbol	Al
Atomic weight	27
Atomic number	13
Electronic configuration	$1s^2$,$2s^2 2p^6$,$3s^2 3p^1$
Position in periodic table	Period-3 and group-IIIA (13)
Valency	3

Occurrence

Aluminum is the most abundant metal on the earth's crust. It occurs about 68.1% in nature. It is highly reactive metal. So, it does not occur in Free State in the nature. In combined state it is found in different types of compounds, like oxides, sulphides, fluorides, etc. The main ores of aluminium are given below.

Ores of aluminium

Bauxite: $Al_2O_3.2H_2O$

Felspar: $K(AlSiO_3.O_8)$

Cryolite: Na_3AlF_6

Extraction of aluminium

Bauxite is the chief ore of aluminium. We can extract aluminium from this ore. The extraction of aluminium involves the following steps.

i. **Collection of ore:** As we know that bauxite is a chief ore of aluminium. So, bauxite ore is collected from the mines using different kinds of tools and techniques.

ii. **Powdering of ore:** The collected bauxite ore has big lumps. So, it is crushed and ground to make fine powder. For crushing and grinding, we use crusher machines, stamp mill and pulverizing machine.

iii. **Heating of ore with sodium hydroxide:** The powdered ore is mixed with sodium hydroxide and heated to form sodium metal aluminate ($NaAlO_2$). This metal aluminate is treated with water and dilute hydrochloric acid to form Aluminium hydroxide [$Al(OH)_3$]. Now, aluminium hydroxide is dried and strongly heated to get aluminium oxide which is called alumina (Al_2O_3).

iv. **Purification of aluminium:** At the end of metallurgical process, the molten alumina is purified by using electrolysis process. As a result, 99% pure aluminium is obtained.

Properties of aluminium

1) **Physical properties**

i. Aluminium is a shiny bluish white metal.

ii. It is a light metal having specific gravity 2.7.

iii. Its melting point is 660°C and boiling point is 1800°C.

iv. It is malleable, ductile and good conductor of heat and electricity.

2) **Chemical Properties**

a) **Reaction of aluminium with air**

Aluminium does not react with dry air, but in the presence of moisture, it reacts with oxygen to form aluminium oxide.

$$4Al + 3O_2 \xrightarrow{\Delta} 2Al_2O_3$$

b) **Reaction with water**

Aluminium does not react with cold water but it reacts with steam to produce aluminium hydroxide.

$$2Al + 6H_2O \rightarrow 2Al(OH)_3 + 3H_2\uparrow$$

c) **Reaction with halogens**

Aluminium reacts with halogens to form respective halides.

$$2Al + 3Cl_2 \rightarrow 2AlCl_3 \quad \text{(Aluminium chloride)}$$

$2Al + 3Br_2 \rightarrow 2AlBr_3$ (Aluminium bromide)

$Al + 3I_2 \rightarrow 2AlI_3$ (Aluminium Iodide)

d) **Reaction with nitrogen**

Aluminium reacts with nitrogen to form aluminium nitride.

$2Al + N_2 \rightarrow 2AlN$ (Aluminium nitride)

e) **Reaction with alkalis**

In presence of water, aluminium reacts with sodium hydroxide to give sodium metal aluminate.

$2Al + 2NaOH + 2H_2O \rightarrow 2NaAlO_2 + 2H_2\uparrow$

f) **Reaction with Acids**

i) **Reaction with hydrochloric acid**

Aluminium reacts with dilute as well as concentrated hydrochloric acid to give aluminium chloride and hydrogen gas.

$2Al + 6HCl \rightarrow 2AlCl_3 + 3H_2\uparrow$

ii) **Reaction with sulphuric acid**

Aluminium reacts with dilute sulphuric acid to give aluminiumsulphate and hydrogen gas.

$2Al + 2H_2SO_4 \rightarrow Al_2(SO_4)_3 + 3H_2\uparrow$

Aluminium reacts with concentrated sulphuric acid to give aluminium sulphate, sulphur dioxide and water.

$2Al + 6H_2SO_4 \rightarrow Al_2(SO_4)_3 + 3SO_2 + 6H_2O$

iii) **Reaction with nitric acid**

Aluminium does not react with dilute as well as concentrated nitric acid. It is because aluminium makes a protective oxide layer over its surface.

Uses of aluminium

i. Aluminium is a good conductor of electricity. So, it is used to make electric transmission wires.

ii. It is used to make different types of household utensils, pictures frames, etc.

iii. It is used for making aluminium foils for wrapping foods, pharmaceutical products, biscuits, chocolates, cigarettes, etc.

iv. It is used to make different parts of aircrafts, ships, cars, etc. as it is a light metal.

v. It is used to make alloys and coins.

12.2 Silver

Introduction

Name	Silver
Latin name	Argentum
Symbol	Ag
Atomic weight	107.9
Atomic number	47
Electronic configuration	$[Kr]4d^{10}5s^1$
Position in periodic table	Period-5 and group-IB (11)
Valency	1

Occurrence

Silver is less reactive than iron and aluminium. So, it occurs in Free State in the earth. It is more reactive than gold. So, it also occurs in combined state making different compounds. The various compounds of silver which are present in the earth are called ores of silver.

Ores of Silver

The main ores of silver are:

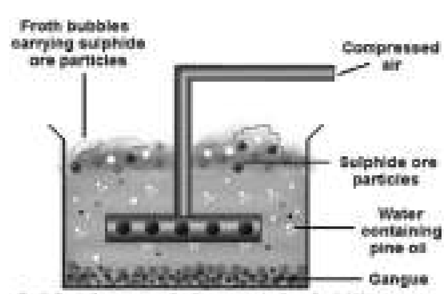
a. Argentite or silver glance (Ag_2S)
b. Horn silver (AgCl)
c. Ruby silver
d. Silver galena
e. Silver copper glance

Extraction of Silver

Silver is mainly extracted from the argentite ore by using Cyanide's process. The cyanide's process has the following steps:

i. **Mining:** In this process, argentite ore of silver is mined from the earth. The mined silver ore contains large amount of gangue and foreign particles. Now the ore is concentrated to remove the unwanted impurities.

ii. **Concentration of the argentite ore:** Argentite is a sulphide ore. So we use froth floatation process to remove its impurities. In this process, the mixture of argentite ore, water and pine oil is kept in a container. The mixture is now strongly agitated with current of air. In pine oil, the sulphide ore becomes wet. The wetted sulphide ore is lighter than water. So, it becomes precipitate and



Froth floatation process for the concentration of sulphide ores.

floats on the top of water in the form of froth. At the same time, the unwanted earthy particles become wet with water. These wetted impurities have more density than water. So, they settle down at the bottom of the container. Now, the froth containing argentite ore is separated like cream from the mill. Thus, the ore and impurities can be separated using froth floatation.

iii. **Treatment with sodium cyanide:** The concentrated ore is then treated with 0.4 to 0.7% solution of sodium cyanide (NaCN) and agitated with a strong blow of air. The solution of sodium cyanide reacts with argentite ore and gives a complex compound sodium argentocyanide $Na[Ag(CN)_2]$ and sodium sulphide (Na_2S). The sodium sulphide then converted into sodium sulphate due to the reaction with oxygen. The necessary chemical reactions are:

$$Ag_2S + 4NaCN \rightarrow 2Na[Ag(CN)_2] + Na_2S$$

$$Na_2S + 2O_2 \rightarrow Na_2SO_4$$

iv. **Precipitation:** The sodium argentocyanide is then treated with scrap zinc. Zinc has more electropositive nature than silver, so it displaces silver from the solution. The silver is then separated and fused with borax or potassium nitrate to get compact silver mass.

$$Zn + 2Na[Ag(CN)_2] \rightarrow Na_2[Zn(CN)_4] + 2Ag$$

v. **Silver Refining:** Silver obtained in these steps also contains some impurities. So, it is purified by using electro refining process.

Properties of silver

1) **Physical properties**

a) Silver is a shining white metal.

b) It is malleable, ductile and good conductor of heat and electricity.

c) The melting point of silver is 956°C and boiling point is 1955°C.

d) Its specific gravity is 10.52

2) **Chemical properties**

a) **Reaction of silver with halogens**

Silver reacts with halogens to give corresponding halides.

$2Ag + Cl_2 \rightarrow 2AgCl$ (Silver chloride)

$2Ag + Br_2 \rightarrow 2AgBr$ (Silver bromide)

$2Ag + I_2 \rightarrow 2AgI$ (Silver iodide)

b) **Reaction of silver with sulphur**

Silver reacts with sulphur to give silver sulphide.

$$2Ag + S \xrightarrow{\Delta} Ag_2S$$

c) **Reaction of silver with acids**

i) **Reaction with hydrochloric acid:**

Silver does not react with dilute hydrochloric acid. But, in the presence of oxygen, silver reacts with concentrated hydrochloric acid to give silver chloride and water.

$4Ag + 4HCl + O_2 \rightarrow 4AgCl + 2H_2O$

ii) **Reaction with sulphuric acid:**

Silver does not react with dilute sulphuric acid. But it reacts with concentrated H_2SO_4 to give silver sulphate, sulphur dioxide and water.

$2Ag + 2H_2SO_4 \rightarrow Ag_2SO_4 + SO_2 + 2H_2O$

iii) **Reaction with nitric acid:**

Silver reacts with dilute nitric acid to give silver nitrate, nitric oxide and water.

$3Ag + 4HNO_3 \rightarrow 3AgNO_3 + NO + 2H_2O$

Similarly, silver reacts with concentrated nitric acid to give silver nitrate, nitrogen dioxide and water.

$Ag + 2HNO_3 \rightarrow AgNO_3 + NO_2 + H_2O$

Uses of silver

i) Silver is used to make different types of ornaments, coins and utensils.

ii) It is used for silver plating.

iii) It is used for filling in the teeth.

iv) It is used for silvering mirror.

v) Silver bromide (AgBr) is used in photography.

12.3 Gold

Introduction

Name	Gold
Latin name	Aurum
Symbol	Au
Atomic weight	197.2
Atomic number	79
Electronic configuration	[Xe] $4f^{14}$ $5d^{10}$ $6s^1$
Position in periodic table	Period-6 and group-IB (11) and block: d-block
Valency	1 and 3

Occurrence

Gold is very less reactive metal. It is less reactive than iron, silver, aluminium and copper. In general conditions, it does not react with most of the substances like air, water, acid, base, etc. So, it is also called a noble metal and occurs in Free State. It is found in reef gold mixed with quartz or alluvial soil.

Ores of Gold

The main ores of gold are:

i. Alluvial Soil or quartz veins

ii. Calverite ($AuTe_2$)

iii. Petzite $(Ag.Au)_2$

Extraction of Gold:

Gold is mainly extracted by using Sluicing and Cyanide process. These processes are illustrated in the following steps:

1. **Sluicing**

   In sluicing process, the alluvial soil is washed using high pressure water. The mixture of alluvial soil and water runs downward in the pan or long channel called sluice. Water and foreign materials constantly flow down along the sluice while the heavier gold particles settle down in the cavities of sluice.

2. **Cyanide process:**

   The cyanide process includes extraction of gold by using solution of sodium cyanide. A strong jet of air is blown to the mixture of impure gold and sodium cyanide solution. Now, the gold gets dissolved in the sodium cyanide solution to form sodium aurocyanide and sodium hydroxide. After that, gold is extracted by adding zinc dust in the solution of sodium aurocyanide. Since zinc is more reactive than gold, so it displaces gold in the form of precipitate.

   $$4Au + 8NaCN + 2H_2O \rightarrow 4\,Na[Au(CN)_2] + 4NaOH$$

   $$Zn + 2Na[Au(CN)_2] \rightarrow Na_2[Zn(CN)_4] + 2Au$$

   The gold thus obtained may contain impurities of zinc and lead. So, it is purified by using several refining processes to get pure gold.

Properties of Gold

1. **Physical properties**

a. Gold has yellowish metallic luster.

b. It is a good conductor of heat and electricity.

> ***Do you know?***
> The specific gravity of gold is 19.3. It means it is 19.3 times heavier than pure water at 4^0C.

c. It is highly malleable and ductile.

d. Its melting point is 1063^0C and boiling point is 2530^0C.

e. Its specific gravity is 19.3.

2. **Chemical properties**

i. Reaction of gold with halogens

Gold reacts with halogens to give gold halides.

$2Au + 3Cl_2 \rightarrow 2AuCl_3$ (Gold chloride)

ii. Reaction of gold with aquaregia

Gold is very less reactive metal. It does not react with individual acid. But, it reacts with aquaregia. Aquaregia is a homogenous mixture of three part of concentrated hydrochloric acid and one part of concentrated nitric acid. Gold reacts with aquaregia to give auric chloride, nitrosyl chloride and water. The chemical reaction is given below.

$\{3HCl + HNO_3 \rightarrow NOCl + 2H_2O + 2Cl\} \times 3$

$9HCl + 3HNO_3 \rightarrow 3NOCl + 6H_2O + 6Cl$

$2Au + 6Cl \rightarrow 2AuCl_3$

$2Au + 9HCl + 3HNO_3 \rightarrow 2AuCl_3 + 3NOCl + 6H_2O$

Uses of Gold

1. Gold is used extensively for making ornaments and jewelleries.
2. It is used for making gold coins.
3. It is used for gold plating of the precious metals, idols, statues, roof tops, temples etc.
4. It is used for gold leaf electroscope.
5. It is used in photography, medicines and filling the teeth.

12.4 Alloy

An alloy is a homogeneous mixture of two or more metals or metals and non-metals. In alloy mixing elements do not loss their identity but we get a resultant metallic substance which has different properties than its components. An alloy with two components is called a binary alloy. The alloy with three components is called a ternary alloy and similarly, if alloy has four components, it is called a quaternary alloy.

Alloys are usually designed to have properties that are more desirable than their components. For example, steel is an alloy which has iron as one of the main components. Here, steel is stronger than iron. Similarly, brass is more durable than copper but more attractive than zinc.

Unlike pure metals, many alloys do not have fixed melting point. Instead, they have a melting range. In alloy, the temperature at which melting begins is called the solidus and that at which melting becomes complete is called the liquids. Some special types of alloys can be designed which have single melting point. They are called eutectic mixtures.

Sometimes an alloy is just named according to the base metal. For example, 14 karat (58%) gold is an alloy of gold with other elements. Similarly, alloy of silver used to make jewellery and alloy of aluminiumis used to make different structures.

Some examples of alloys

1. Steel: Iron (80–98%), carbon (0.2–2%), plus other metals such as chromium, manganese, and vanadium.
2. Stainless steel: Iron (50%)+, chromium (10–30%), plus smaller amounts of carbon, nickel, manganese, molybdenum, and other metals.
3. Brass: Copper (65–90%), zinc (10–35%).
4. Amalgam: Mercury (45–55%), plus silver, tin, copper, and zinc.
5. Bronze: Copper (78–95%), tin (5–22%), plus manganese, phosphorus, aluminum, or silicon.
6. Bell Metal: Copper (78%) and (tin 22%)
7. White gold (18 carat gold): Gold (75%), palladium (17%), silver (4%), copper (4%)
8. Nichrome: Nickel (80%), chromium (20%)

	Name of alloy	**Composition**
1.	Steel	Iron (80–98%), carbon (0.2–2%), other metals such as chromium, manganese, and vanadium
2.	Stainless steel	Iron (more than 50%), chromium (10–30%), small amount of carbon, nickel, manganese, molybdenum and other metals.
3.	Brass	Copper (65–90%) and zinc (10–35%)
4.	Amalgam	Mercury (45–55%), silver, tin, copper and zinc
5.	Bronze	Copper (78–95%), tin (5–22%), manganese, phosphorus, aluminum, or silicon
6.	Bell Metal	Copper(78%) and tin(22%)

7.	White gold (18 carat gold)	Gold (75%), palladium (17%), silver (4%) and copper (4%)
8.	Nichrome	Nickel (80%) and chromium (20%)

12.5 Amalgam

Amalgam is an alloy of mercury. It is made up of 45–55% mercury with silver, tin, copper and zinc. Amalgams can be formed with many metals. Among them, the silver-mercury amalgam is used in dentistry to fill teeth and gold-mercury amalgam is used in the extraction process of gold.

Summary

1. Aluminum is the most abundant metal on the earth's crust. It occurs about 68.1% in nature.
2. The ores of aluminium are Bauxite: $Al_2O_3.2H_2O$, Felspar: $K(AlSiO_3.O_8)$, and Cryolite: Na_3AlF_6
3. Bauxite is the chief ore of aluminium.
4. Aluminium is used for making aluminium foils for wrapping foods, pharmaceutical products, biscuits, chocolates, cigarettes, etc.
5. Silver is less reactive than iron and aluminium. So, it occurs in Free State in the earth.
6. The main ores of silver are Argentite or silver glance (Ag_2S), Horn silver (AgCl), Ruby silver, Silver galena and Silver copper glance.
7. Gold is very less reactive metal. It is less reactive than iron, silver, aluminium and copper.
8. In general conditions, gold does not react with most of the substances like air, water, acid, base, etc. So, it is also called a noble metal and occurs in Free State.
9. The main ores of gold are Alluvial Soil or quartz veins, Calverite ($AuTe_2$), and Petzite $(Ag.Au)_2$
10. Aquaregia is a homogenous mixture of three part of concentrated hydrochloric acid and one part of concentrated nitric acid.
11. Gold reacts with aquaregia to give auric chloride, nitrosyl chloride and water.
12. An alloy is a homogeneous mixture of two or more metals or metals and non-metals.
13. Amalgam is an alloy of mercury. It is made up of 45–55% mercury with silver, tin, copper and zinc.

Exercise

A. **Tick (√) the best alternatives from the followings.**

1. .Gold is extracted by:
   a. Argento- cyanide process b. Cyanide process
   c. Alumino process d. Forth floatation process
2. In the extraction of silver ………… is added to the ore.
   a. Potassium permanganate b. Sodium cyanide
   c. Sodium argento cyanide d. Potassium phosphate
3. Gold is called a :
   a. Global metal b. Colossal metal
   c. Noble metal d. Reactive metal
4. Silver is extracted by
   a. Paramagnet method b. Dice method
   c. Sluice method d. Froth flotation method
5. The main ore of aluminium is :
   a. Bauxite b. Argentite
   c. Haematite d. Alluvial soil

B. **Give very short answers to the following questions.**

1. Define metallurgy.
2. What is froth floatation method?
3. Show the sub-shell electronic configuration of aluminium.
4. Define aquaregia.
5. Where do we use sluice method?
6. Mention the period and group of gold and silver.
7. Which element is present in aquaregia to react with gold?
8. Write down the main ore of gold, silver and aluminium.
9. Define alloy and amalgam.
10. Write down any two alloys with their composition.

C. Give short answers to the following questions.

1. How does silver react with the following chemicals? Show the balanced chemical reaction.
   i. Hydrochloric acid
   ii. Nitric acid
   iii. Sulphuric acid
2. Show the reaction of gold with aquaregia and chlorine.
3. Why is aluminium used to make parts of aircraft?
4. Write down any four uses of each silver, aluminium and gold.
5. List any three physical properties of each aluminium, gold and silver.
6. How does aluminium react with the following chemicals? Show the balanced chemical reaction.
   i) Hydrochloric acid
   ii) Nitric acid
   iii) Sulphuric acid

D. Give long answers to the following questions.

1. Describe the metallurgical steps of aluminium.
2. What is chief ore of silver? Describe the extraction of silver.
3. Which metal do you get from alluvial soil? Describe the extraction method of this metal.

Project Work

1. Make a list of different kinds of alloys which are present in your home and school.
2. Iron, copper, silver, gold are common metal. But, mostly jewelers are made from silver and gold. Ask to our relatives and parents about their views of using only gold and silver.

Glossary

Ores	:	those earthy substances which contain metals
Noble metal	:	metal which is passive towards the chemical reaction
Aquaregia	:	the mixture of hydrochloric acid and nitric acid in the ration of 3:1.
Amalgam	:	the alloy of mercury with other metal

Unit 13

Biomolecules

Neuberg was an early pioneer in biochemistry, and he is often called "the father of modern biochemistry". His varied research interests resulted in important contributions to the understanding of fermentation processes, solubility and transport phenomena in cells, the chemistry of carbohydrates, sugars, enzymes, amino acids and photochemistry.

Carl Alexander Neuberg (1877-1956)

13.1 Introduction to biomolecules

All living things are composed of biomolecules. All forms of life like bacteria, algae, plant and animals are made up of similar biomolecules that are responsible for life. All the carbon compounds that we get from living tissues can be called as biomolecules which occur naturally in living organisms.

Biomolecules include Organic compounds like proteins, carbohydrates, lipids and nucleic acids and inorganic compounds like minerals and water. We will learn about the importance and uses of carbohydrate, protein, lipid, enzyme and nucleic acid in human body in this unit.

13.2 Organic compounds

There are four major classes of organic biomolecules:

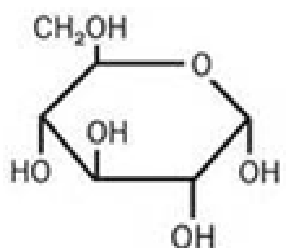
- Carbohydrates
- Lipids
- Proteins
- Nucleic acids

13.2.1 Carbohydrates

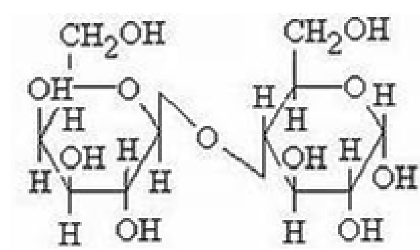
Carbohydrates are the organic molecules which are easily available in nature. They are good source of energy. They are a long chain of sugars and on the basis of the numbers of sugars they can be classified as monosaccharides, oligosaccharides and polysaccharides.

Monosaccharides are simple sugars that are composed of 3-7 carbon atoms. Glucose, fructose and galactose are common monosaccharides. **Oligosaccharides** are the saccharides which are made up a chain of 2-10 monosaccharides. Disaccharide is one of the examples of oligosaccharides which is made up of two monosaccharides. Maltose (Glucose + Glucose), sucrose (Glucose + Fructose) and lactose (Glucose + Galactose) are the examples of disaccharides. **Polysaccharides** are polymers of monosaccharides. The most common polysaccharides are glycogen, starch and cellulose.

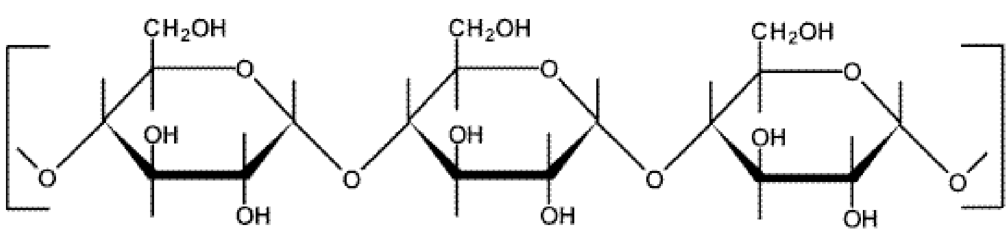
> ***Do you know?***
>
> A normal healthy person needs 500 gm (2050 K Cal) of carbohydrate per day. One gm of carbohydrates gives 4.1Kcal energy. Pregnant woman, athletes, labours, etc. need more amount of carbohydrates.



Monosaccharide [Glucose ($C_6H_{12}O_6$)]



Disaccharide[Lactose ($C_{12}H_{22}O_{11}$)]



Polysaccharide (Starch)

Sources of carbohydrates

Plant contains a large number of carbohydrates. Glucose is the basic carbohydrate from which many other complex carbohydrates are formed in a plant body. Carbohydrates are found in the plants like sugarcane, sugar beet, pine apple, carrot roots, milk, rice, wheat, potatoes, corns, etc.

Importance of carbohydrates

Carbohydrates are an essential part of our diet which is found naturally in grains, fruits and vegetables. The benefits of carbohydrates are as follows:

1. Carbohydrate provides energy to the body.
2. Carbohydrate contains fiber which promotes regular bowel movement so as to prevent constipation.
3. Carbohydrates form structural and protective components, like cell wall of plants and microorganisms.
4. Carbohydrates participate in biological transport and activation of growth factors.
5. Cellulose is used to make paper and cloth.

13.2.2 Lipids

Lipids are composed of long hydrocarbon chains and hold a large amount of energy. They are the energy storing molecules. They are generally esters of fatty acids and are building blocks of biological membranes. Lipids can be classified as simple lipids (fats, oils and waxes), compound lipids (glycolipids, phospholipids and lipoprotein) and derived lipids (cholesterols).

A) Simple lipids: It includes fats (monoglyceride, diglyceride, triglyceride etc.), oils and waxes.

B) Compound/Complex lipids: They are called complex lipid as they are formed with the combination of lipid with the non-lipid molecule. For ex ample glycolipids (lipids + carbohydrates), phospholipids (lipids + phos phates) and lipoprotein (lipids + protein).

C) Derived lipids: These include hydrolysed products of simple and compound lipids. For example cholesterol.

Sources of lipids

Both plant and animal sources contain different types of lipids that can provide both beneficial and harmful effects. Some examples of lipid-rich foods include meats, oils, butter, milk, cheese, nuts, eggs and avocados.

Importance of Lipids

Our body needs some lipids to maintain our health and body systems. Some importance of lipids are as follows:

1. Lipids act as the reserve energy of the body.
2. Lipids serve as source for fat soluble vitamins like A, D, E, K.
3. Layers of fat below the skin of animals provide protection from cold and help to maintain body temperature.

4. Cholesterol is the precursor of bile acids, vitamin D and steroids.
5. Essential fatty acids play important role in pain, fever, inflammation, blood clotting and to prevent from bacterial diseases.

13.2.3 Proteins

Proteins are important molecules in cells and therefore are known as building blocks of life. Proteins provide structure and protection to the body in the form of skin, hair, cartilage, ligaments, muscles, tendons, etc. They are formed from 20 different amino acids.

> **Do you know?**
> Due to deficiency of protein in the body, infants are usually suffered from marasmus and kwashiorkor.

Some biologically important proteins are trypsin, pepsin, collagen, keratin, insulin, thrombin, fibrinogens, albumin, globulin, hormones, etc.

Amino Acid: It is the monomer of protein. It is also called building block of protein as it forms a long chain to build a protein. Amino acids are colourless, crystalline and water soluble (but not soluble in organic solvents like alcohol) biomolecule which has great importance in human's life. There are 20 different types of amino acids in nature which are classified into essential, semi-essential and non-essential amino acids.

A) Essential Amino Acids: Amino acids which are not synthesized in the body and hence essential to supply through diet. There are eight amino acids (leucine, isoleucine, lysine, phenylalanine, tryptophan, valine, methionine and threonine) which are considered as essential amino acids.

B) Non-essential Amino Acids: Amino acids which are synthesized in the body and hence not essential to supply through diet. There are ten amino acids (alanine, serine, glycine, glutamine, tyrosine, aspartic acid, cysteine, proline, glutamic acid and asparagine).

C) Semi-essential Amino Acids: Amino acids which can partly synthesized in adult human body but are necessary to supply through diet in infants are called semi-essential amino acids. There are two amino acids (arginine and histidine) which are considered as semi-essential amino acids.

Sources of protein

Some of the common sources of proteins are groundnuts, beans, whole cereals, pulses, fish, egg, meat, milk, cheese. Egg protein is considered as reference protein as it is absolutely digestible and chemically complete.

> ***Do you know?***
> Egg is a good source of energy. It contains all the nutrients except vitamin C and carbohydrate.

There are two groups of proteins; First class proteins which includes the proteins containing all the types of amino acids (animal proteins like egg, meat, etc.) and Second class proteins which do not contain all the types of amino acids (like plant proteins).

Importance of protein

1. Some proteins are hormones; they regulate many body functions.
2. Some proteins act as enzymes; they catalyze or help in biochemical reactions.
3. Some proteins act as antibodies; they protect the body from infection.
4. Proteins transport different substances in blood of different tissues.
5. Proteins help in contraction of muscle and cells of our body.
6. Proteins helps in healing of wounds and prevent blood clotting.

Enzymes

Enzymes are large biomolecules which are responsible for many activities or chemical reactions necessary for the life. Chemically enzymes are made up of proteins and they act as catalyst for various biological processes. They only increase the rate of reaction without themselves being changed. They are highly specific to the reaction producing specific products. As they are made up of proteins or long chain of amino acids, they get affected by the temperatures and high levels of acids. The degree of temperature and the level of acids required to break down the molecules of enzymes depends on the strength of the bonds between amino acids. And once the bonds are degraded they become inactive or non-functional. The part of enzyme which acts as the catalyst is called active site of the catalyst, whereas remaining part is called substrate. On the basis of the reaction they catalyze, they are classified into 6 groups: oxidoreductase, tansferase, hydrolase, lyases, isomerase and ligases. Some examples of enzymes are pepsin, erepsin, amylase, lipase, lactase, etc.

> ***Do you know?***
> An adult person needs 70-100gm of protein on average per day. 1gm of protein gives 5.65Kcal energy.

Characteristics of Enzymes

Enzymes have the following characteristics:

1. They possess great catalytic power.
2. They are highly specific.
3. Their activity can be controlled.
4. They are made up proteins.
5. They can be denatured by alcohol, heat, concentrated acids and alkaline reagents.
6. The activity of enzyme is high at optimum temperature.

Functions of Enzymes

Enzymes perform wide variety of functions in living organisms which are summarized below:

A) Biological functions of enzymes

1. Enzymes help in the signal transduction and cell regulation.
2. They help in the muscle contraction.
3. Enzymes present in virus are for their pathogenic effect.
4. They help in the digestion in animals.
5. Some enzymes also help in the conversion of large organic molecules into absorbable form.

B) Industrial function of enzymes

1. Some of the enzymes like amylases are used in the production of sugars from starch.
2. Some enzymes like trypsin are used for pre-digestion of baby foods.
3. Enzymes from barley are used in brewing industries.
4. Some enzymes like amylases, proteases, etc are used in beer production.
5. Enzyme like renin is used in dairy industries for manufacturing cheese.
6. Catalase enzyme converts latex into foam rubber.
7. Some enzymes have great value in molecular biology, forensic science, etc.

13.2.4 Nucleic Acids

Nucleic acids are essential biomolecules found in the chromosomes of living cells and viruses. The nucleic acids include the DNA and the RNA. They are the hereditary determinants of living organisms. They are present in most living cells either in free-state or bound to proteins as nucleoproteins. Nucleic acids are made of polymer of nucleotides. Nucleotide consists of nitrogenous base, a pentose sugar and a phosphate group. A nucleoside is made of nitrogenous base attached to a pentose sugar. The nitrogenous bases are purine (adenine and guanine) and pyrimidine (thymine, cytosine and uracil). Among these nitrogenous bases A, C, G and T are found in DNA, whereas A, C, G and U are found in RNA.

Importance of nucleic acids

1. Nucleic acids store and transfer genetic information.
2. DNA controls the synthesis of RNA in the cell.
3. RNA directs synthesis of proteins.

A = Adenine
G = Guanine
T = Thymine
U = Uracil
C = Cytosine

Activity: Make a list of the types of food that are available in your surroundings (home, school, market, etc). After forming the list of foods, write the biomolecules present in corresponding foods. Present your list to the class.

13.3 Inorganic compounds

13.3.1 Minerals

Minerals are inorganic biomolecules which are taken by the body in the form of salts. They have various functions such as formation of tissues and bones, conduction of nerve impulses, formation of RBCs, etc. There are eight major elements essential for growth and development of living beings and are called essential elements. They are sodium, chlorine, potassium, calcium, phosphorus, nitrogen, sulphur and magnesium. Some elements like fluorine, zinc, copper, iodine, iron, manganese, chromium, cobalt, etc. are required in small amounts and are called trace elements. All the minerals are basically obtained from the plants and the plant absorb these minerals from the soil.

13.3.2 Water

Water is the most abundant chemical compound in the cells of living beings. It forms about 60% to 90% of the total chemical content of the cell. Water is found in two forms- free and bound. Nearly 95% of water occurring in a cell is found in the free state which can be used for metabolic process while about 5% occurs in a state bound to other molecules.

Water molecule is very small and readily passes through the cellular membranes. It is composed of two atoms of hydrogen and one atom of oxygen linked up by covalent bonds.

Importance of Water

1. Water is required for several biochemical reactions like hydrolytic digestion of nutrients, photosynthesis, etc.
2. Water is important for the transportation of materials.
3. Water maintains the turgidity of cell and their organelles.

Biomolecules	Organic Biomolecules	**Carbohydrates**	**Monosaccharide:** Glucose, Fructose, Galactose
			Disaccharide: Maltose (glucose+glucose), Sucrose (glucose+fructose)
			Polysaccharide: Glycogen, Starch, Cellulose
		Lipids	**Simple lipids:** Monoglyceride, diglyceride, triglyceride, oils, wax, etc.
			Compound lipids: glycolipid (carbohydrate+lipid), phospholipid (phosphate+lipid), lipoprotein (protein+lipid)
			Derived lipids: cholesterol (hydrolysed products of simple and compound lipids)
		Aminoacids (monomers of protein)	Essential aminoacids: leucine, isoleucine, lysin, phenylalanine, tryptophan, valine, methionine, threonine (8)
			Semi-essential: Arginine, Histidine (2)
			Non-essential: Alanine, serine, glycine, glutamine, tyrosine, cysteine, proline, asparagine, aspartic acid, glutamic acid (10)
		Protein	First class protein: contains all aminoacids (eg: animal proteins)
			Second class protein: contains only few of aminoacids (eg: plant proteins)
		Enzymes	Oxidoreductase (helps in oxidation-reduction reaction)
			Transferase (helps in transferring chemical groups from one compound to another)
			Hydrolase (helps in breaking down water molecule into ions)
			Lyase (helps in removal of functional groups from their substrates)
			Isomerase (helps in making isomers from the substrate of enzyme molecules)
			Ligase (helps in synthesis of monomers to make long chain of polymers)
		Nucleic Acid	DNA (transfers genetic information and controls the synthesis of RNA in cell)
			RNA (helps in protein synthesis)
		Minerals	Essential elements: Na, Cl, K, Ca, P, S, Mg, N
			Trace elements: F, Zn, Cu, I, Fe, Mn, Cr, Co, etc.
		Water	60-90% of total chemical content of body
			About 95% in free state and 5% in combined state

Summary

1. All the carbon compounds that we get from living tissues can be called as biomolecules which occur naturally in living organisms.
2. Biomolecules include organic compounds like proteins, carbohydrates, lipids and nucleic acids and inorganic compounds like minerals and water.
3. Carbohydrates are a long chain of sugars and on the basis of the numbers of sugars they can be classified as monosaccharides, oligosaccharides and polysaccharides.
4. Monosaccharides are simple sugars that are composed of 3-7 carbon atoms. Example: glucose, fructose, galactose, etc.
5. Oligosaccharides are the saccharides which are made up a chain of 2-10 monosaccharides. Example: sucrose, lactose, etc.
6. Polysaccharides are polymers of monosaccharides. Example: glycogen, starch, cellulose, etc.
7. Lipids are the energy storing fat molecules.
8. Lipids can be classified as simple lipids (fats, oils and waxes), compound lipids (glycolipids, phospholipids and lipoprotein) and derived lipids (cholesterols).
9. Proteins provide structure and protection to the body in the form of skin, hair, cartilage, ligaments, muscles, tendons, etc.
10. Amino acid is the monomer of protein. It is of three types; essential, semi-essential and non-essential amino acids.
11. Amino acids which are not synthesized in the body and hence essential to supply through diet are called essential amino acids. There are eight essential amino acids.
12. Amino acids which are synthesized in the body and hence not essential to supply through diet are called non-essential amino acids. There are ten non-essential amino acids.
13. Amino acids which can partly synthesized in adult human body but are necessary to supply though diet in infants are called semi essential amino acids. There are two semi-essential amino acids.
14. Nucleic acids are essential biomolecules found in the chromosomes of living cells and viruses.
15. Minerals are inorganic biomolecules which are taken by the body in the form of salts.
16. Water is the most abundant chemical compound in the cells of living beings. It contributes about 60 to 90% of the total chemical content of the cell.

Exercise

A. **Tick (√) the best alternative from the followings.**

1. Kwashiorkor is due to the deficiency of:
   i) carbohydrate   ii) protein
   iii) lipid   iv) nucleic acid
2. First-class proteins are those which are……
   i) synthesized in the body   ii) not synthesized in the body
   iii) rich in essential amino acids   iv) rich in non-essential amino acids
3. There are ……… types of non-essential amino acids in the nature
   i) 8   ii) 10
   iii) 12   iv) 20
4. Fructose is a:
   i) monosaccharide   ii) disaccharide
   iii) oligosaccharide   iv) polysaccharide
5. Lactose is the combination of which two monosaccharide?
   i) Glucose and Fructose   ii) Glucose and Galactose
   iii) Glucose and Glucose   iv) All of the above

B. **Answer the following short questions.**

1. Define biomolecules with example.
2. Differentiate between essential amino acid and non-essential amino acid.
3. Egg is considered as a reference protein, why?
4. Why should atheletes, labourers, etc. need high carbohydrate in their diet?
5. Write any two importance of each of protein, carbohydrate and nucleic acid each.
6. What roles do the proteins play in our body?
7. Write importance of enzymes.

C. **Answer the following long questions.**

1. Write brief note on carbohydrate.
2. Explain in brief about the importance of biomolecules in our daily life.
3. Enzymes are called bio-catalyst. Explain.

Project work

Define Balanced diet. What should be the nutrients (biomolecules) present in the diet to consider as a balanced diet? Now, collect the names of the food that your neighbours consume for a week and prepare a note whether they are having balanced diet or not. If not also explain what type of complications they are facing. Present your findings to the class.

Glossary

Organic compounds:	The carbon and hydrogen containing compounds which are extracted from living organisms
Inorganic compounds:	The compounds which do not contain carbon and hydrogen both it its molecule and are found in earth's crust
Esters:	The organic compound which contain function group –COOR
Crystalline:	A substance having definite shape
Denaturation:	The process of breaking down of bonds between amino acids (peptide bonds)
Precursor:	A substance from which another substance is formed during metabolic reaction
Steroid:	Fat soluble organic compound having 17 carbon atoms in 4 rings which may have important physiological effect in organisms
Turgidity:	Acquiring essential shape by a cell or organelles by absorbing and keeping required amount of water

Unit 14

Cell Biology

Palade is the father of modern cell biology. He is internationally recognized for his pioneering use of electron microscopy and cell fractionation. He established the pathway for synthesis and transport of proteins along the secretory pathway, illuminating how cells build and transport their protein building blocks. He was also awarded the Nobel Prize for Physiology or Medicine in 1974 for his contributions to the understanding of cell structure, chemistry and function. He share the Nobel prize with Albert Claude and Christian de Duve.

Dr. George Palade (1912-2008)

14.1 Introduction

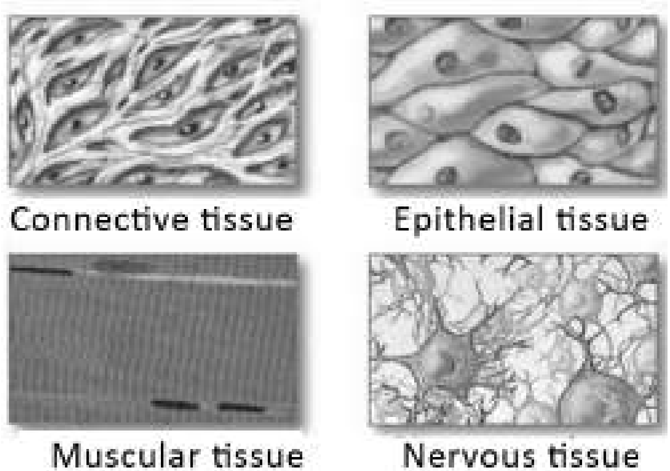
Cell biology is also called cytology. The term cytology was derived from two Greek words; kytos, which means hollow, vessel and logos, which means study. That means **cell biology is a branch of biology which deals with the study of cells, their structures and functions.**

Cell biology explains the structure, organization of the organelles they contain, their physiological properties, metabolic processes, signaling pathways, life cycle, and interactions with their environment. In this chapter, we will learn about the types of animal tissues and plant tissues, structure of roots of monocot and dicot plants, structure of stem of monocot and dicot plants. We will also learn about the somatic cell, process and importance of mitosis cell division, germ cell and process and importance of meiosis cell division in this unit.

14.2 Tissue

In multicellular organisms, a group of similar cells perform specific function; such an aggregation of cells is called tissue. The branch of science which deals with the study of tissue is known as histology. The word tissue is derived from a Latin word tissu which means 'weave'. Cells that form tissues are woven together to form tissues. **A tissue is defined as a group or layer of cells of common origin that perform specific functions**. For example, muscle tissue is a group of muscle cells. Tissue can be divided into two different types: Animal tissue and plant tissue.

14.2.1 Animal Tissue



Animal organs are made up of four basic types of tissues which have distinctive features and specific functions which combine to form functioning organs.

The animal tissues are classified into four types:

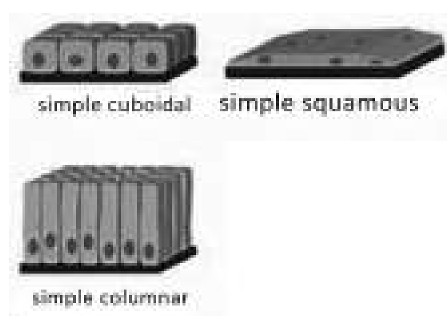
1. Epithelial tissue
2. Connective tissue
3. Muscular tissue
4. Neural tissue (Nervous tissue).

Epithelial Tissue

Epithelial tissue, also called epithelium, forms the outer covering or lining part of the body. It is composed of closely packed cells, arranged in flat sheets. Epithelial tissues form the surface of the skin, line many cavities of the body and cover the internal organs. Epithelial tissue has two major functions in our body. One is to protect the underlying tissues and the other is secretion.

There are two types of epithelial tissues - Simple epithelium (made up of single layer of cells) and Compound/Stratified epithelium (made up of more than one layer of cells).

A) Simple epithelium



It is composed of single layer of cells. It functions as a lining of cavities of body, ducts and tubes. On the basis of structural modification the simple epithelium is divided into three types: Squamous, Cuboidal and Columnar cells.

1) **Simple Squamous epithelium:** It is made up of a single layer of irregular, thin, flattened cells with irregular boundaries. The cells fit closely together forming a pavement, hence also called pavement epithelium. An oval nucleus is present in the middle part of the cell.

   It is found in the walls of blood vessels, Bowman's capsule of nephron and in air sacs of lungs. They are involved in functions like exchange of gases, ultrafiltration, diffusion and protection.

2) **Simple Cuboidal epithelium:** It is made up of a single layer of cube-like cells. Nucleus is round and centric. They are commonly found in ducts of glands and tubules of nephrons. Its main functions are secretion and absorption. It is of two types:

a) **Ciliated cuboidal:** The cuboidal cells having cilia on the free surfaces.

b) **Brush bordered cuboidal:** The cuboidal cells having microvilli at the free ends of the cells.

3) **Simple Columnar epithelium:** It is composed of single layer of tall and slender cells. Each cell has an elongated nucleus at its base. The free surface of these cells may have microvilli. They are found in lining of stomach, intestine and gall bladder. It helps in secretion and absorption. These are of two types:

a) **Ciliated columnar epithelium:** If the columnar epithelium bears cilia on their free surface they are called ciliated epithelium. They help to move particles by secreting mucus in a specific direction. They are present in the inner surface of bronchioles, spinal cord and fallopian tubes.

b) **Brush bordered columnar epithelium:** If the columnar epithelium bears microvilli on their free surface they are called brush bordered epithelium. It is located on intestinal mucosa. It helps to increase the surface area for absorption of nutrients.

*Note: **Pocudostrutified column or epithelium:** Poeudostrafied volumn or epithelium is a columnar tissue formed by a single layer of cells but give the appearane of a being made from multiple layers from cross-section. It can make confusion with the stontified columnar tissue is found monstly in the inner walls of upper respiratory tract (nose, tractea bronchi) and hence also called as respiratory columnar etithelim.*

4) **Glandular epithelium:** It is the modified form of columnar epithelium which is composed of single layer of cells. The cells it contains are highly specialized for the secretion of chemical substances. This epithelium is found in liver, pancreas, mammary glands, sweat glands, etc. and is responsible for secretion of enzymes, hormones, saliva, sweat, etc.

B) Compound or Stratified epithelium

Compound epithelium is made up of more than one layer of epithelial cells. Basement membranes are usually absent in this epithelium. It functions as a protective covering, It covers dry surface of skin, moist surface of buccal cavity, pharynx and inner lining of salivary glands.

Stratified epithelial tissues are of following types:

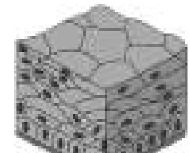


Fig: Stratified Squamous Epithelium

1) **Stratified Squamous Epithelium:** It is much thicker than simple epithelial tissue. It is composed of several layer of cells. The innermost layer is composed of germinative columnar and cuboidal cells. It is of two types:

   a) **Keratinized epithelium:** It is found on outer dry surfaces of the body like skin, hair and nails. It contains keratin (insoluble fibrous protein) with water proof layer. It is also resistant to friction and bacterial infections.

   b) **Non-keratinized epithelium:** It is found on wet surfaces like mouth cavity, tongue, pharynx, oesophagus, vagina, etc. which protects the surface from drying.

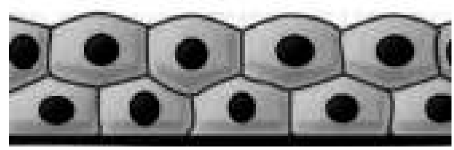


Fig: Stratified Cuboidal Epithelium

2) **Stratified Cuboidal Epithelium:** It consists of cuboidal cells in the outermost layer whereas the innermost layer consists of germinative cells. It forms the lining of ducts of glands like sweat glands, mammary galnds, conjunctiva of eyes, etc. It helps to circulate nutrients and increase absorption. It also protects from pathogens.

3) **Stratified Columnar Epithelium:** It is found in lining of vasa deferentia, male urethra, trachea, bronchi, etc. It helps in protection and secretion.

4) **Transitional Epithelium:** It is found in urinary bladder and ureter. It helps in the expansion of the organ.

Fig: Stratified Columnar Epithelium

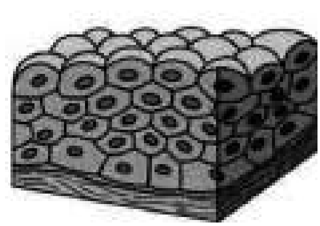


Fig: Transitional Epithelium

Function of Epithelial tissue

1. Epithelial tissue covers the outside of the body and lines organs, blood and lymph vessels and cavities.
2. It protects the internal structures of the body from mechanical injuries and dehydration.
3. It also helps to protect against microorganisms.

4. It absorbs, secretes, and excretes substances. In the intestines, this tissue absorbs nutrients during digestion, in glands it secretes hormones and enzymes, and other substances. Epithelial tissue in the kidneys excretes wastes, and in the sweat glands excretes perspiration (sweat).
5. It contains sensory nerves in the skin, tongue, nose and ears.

Connective Tissue

Connective tissues are the most abundant fibrous tissues. They give shape to organs and hold them in place. Blood, bone, tendon, ligament, adipose and areolar tissues are examples of connective tissues. All connective tissues except blood secrete structural proteins called collagen or elastin.

Connective tissues are classified into four types - Loose connective tissue,Dense connective tissue, Specialized or Supportive connective tissue and Fluid connective tissue.

A) Loose Connective tissues - These tissues have cells and fibres that are loosely arranged in a semi-fluid ground substance. They are of two types: areolar tissue and adipose tissue.

1) Areolar tissue: Areolar tissue is the simplest and the most widely distributed connective tissue found in the animals. It contains matrix having fibres (white collagen fibres, yellow elastic fibres, fibrocytes, etc) and connective tissue cells.It is present beneath the skin, between and around the muscles, blood vessels, alimentary canal, respiratory tract, etc. It forms the basement membrane of epithelium, helps in wound healing, binds and supports the visceral organs.

2) Adipose tissue: This type of tissue contains large number of fat cells (adipocytes); that are specialized to store fats. Fibres are less or absent in adipose tissue. It is found beneath the skin around certain organs like kidney, eyes, heart, etc.

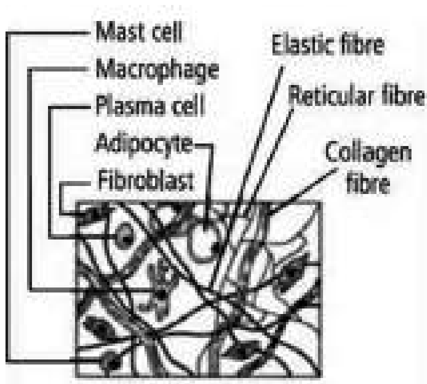


Fig: Areolar tissue

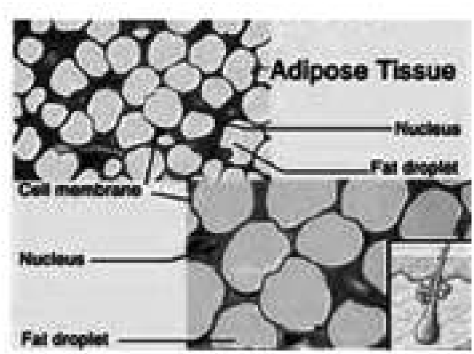


Fig: Adipose tissue

B) Dense Connective tissue - Fibers and fibroblasts are packed compactly in dense connective tissue. Tendons are dense regular tissues that attach skeletal muscle to bones and ligaments attach bone to other bones. Collagen is the dense irregular tissue present in the skin. It can be classified into two types: White fibrous tissue and Yellow elastic tissue.

1) **White fibrous tissue:** It is a modified form of areolar tissue but contains thick layer of white collagen fibres which are tough but non-elastic. White fibres are so arranged that they run parallel to each other forming a tendon which helps to connect muscles with bones. They are found in brain (duramater), spinal cord, pericardium of heart, etc. It provides mechanical strength against stretch.

2) **Yellow elastic tissue:** It is also a modified form of areolar tissue but composed of thick layer of yellow elastic fibres. It forms a ligament which helps to bind bone with bone. It is found in the joints of bones. It provides flexibility and elasticity to the body.

C) Specialized connective tissue: These are the special connective tissues which help in the formation of skeleton of the vertebrates. It protects some delicate organs from mechanical injuries. Cartilage and bones are the two types of specialized connective tissue.

1) **Cartilage:** Cartilage is solid, flexible tissue. The cartilage cell is called chondrocytes. Cartilage is present in the tip of the nose, outer ear joints, between bones of vertebral column.

2) **Bones:** They are hard and non-flexible, rich in calcium salts and collagen fibers. It provides structural frame to the body. The bone cells are osteocytes. The bone marrow in some bones is the site of production of blood cells.

D) Fluid connective tissue: These are the connective tissues which are available in the liquid form and specially helps in the circulation of the materials in the body, regulation of body temperature, minerals, water, etc and protection of the body. It includes Blood and Lymph.

1) **Blood:** It is a red, viscous fluid connective tissue which is slightly alkaline (pH 7.2-7.3) in nature. It contains plasma and blood corpuscles (RBCs, WBCs and platelets). It is the major component which helps in the transportation of food, nutritive materials, waste materials, respiratory gases, etc. to their respective sites. Especially WBCs help to protect animals' body from various diseases. Plasma and platelets help in the coagulation of blood and protect the body from excessive loss of blood.

2) **Lymph:** Lymph is another kind of fluid which is similar to the blood but RBCs and some other proteins of blood are absent. It contains some proteins, water in its plasma. It also helps in the transportation of different materials from the tissues into the blood stream and vice versa.

Functions of Connective Tissue

Connective tissue performs the following important functions in the body.

1. Connective tissue acts as a packing material. It binds tissues with one another. Such as muscles with skin and muscles with bones.
2. Adipose tissue stores fat and conserves body heat.
3. It forms convenient pathways for vessels (blood vessels and lymphatic vessels) and nerves.
4. Loose connective tissue facilitates movements between the adjacent structures, and minimizes friction and pressure effects.
5. Connective tissue helps in the repair of injuries whereby the fibroblasts lay down collagen fibers to form the scar tissue.
6. The macrophages of connective tissue serve a defensive function against the bacterial invasion by their phagocytic activity. They also act as scavengers in removing the cell debris and foreign material.

Muscular Tissue

Muscular tissues are soft tissues made up of long cylindrical fibers, arranged parallel. These fibers are composed of fine fibrils known as myofibrils. Its function is to produce force and cause motion within internal organs. Muscles are of three types skeletal, smooth, and cardiac muscles.

A) **Skeletal Muscle -** Skeletal muscle is also known as striated (striped) muscle. This muscle is called voluntary as the movement of this muscle is under the control of our will. It is closely attached to the bones, and found in the limb, tongue, abdominal wall, etc. It helps in locomotion and gives general support and posture to a body.

B) **Smooth Muscle -** Smooth muscle is also known as unstriated (unstriped) muscle. It is found in the lining of visceral organs like alimentary canal, blood vessels, urinary bladder, etc. so also known as visceral muscle. This muscle is involuntary as the movement of these muscles is not in our control but under the control of autonomic nervous system.

C) **Cardiac Muscle -** Cardiac muscle are specialized tissue present only in the wall of heart. It is responsible for contraction and relaxation of cardiac tissue and distribution of blood. It functions spontaneously without tiring. It resembles both striated and unstraited as it contains alternating light and dark bands and looks like striated muscle but also involuntary in action as unstriated muscle.

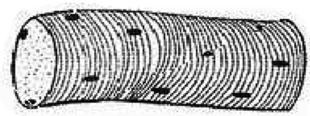


Fig: Skeletal muscle

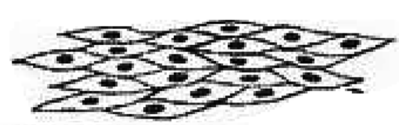


Fig: Smooth muscle

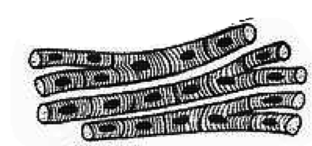


Fig: Cardiac muscle

Function of Muscular Tissue

1. Muscular tissue helps in locomotion. For example, when muscles contract or relax, the fibers pull or relax bones to which they are attached, thus allowing locomotion.
2. The contraction of muscle produces heat, which keeps the body warm during the winter cold months.
3. The muscle provides the framework for the body and also helps in maintaining posture and flexible joints.
4. The circulation of blood is also partly depending on the contraction of muscle tissue.
5. Smooth muscle is responsible for the contractility of hollow organs, such as blood vessels, the gastrointestinal tract, urinary bladder, etc.
6. Cardiac muscle is responsible for contraction and relaxation of cardiac tissue and distribution of blood.

Nervous Tissue

Nervous tissue is the primary tissue that composes the central nervous system and the peripheral nervous system. They are responsible for sensing stimuli and transmitting signals to and from different parts of an organism. Nervous tissue consists of neurons, neuroglia, ependymal cells and neuro secretory cells.

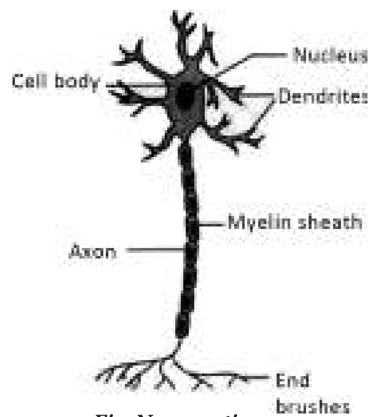


Fig: Nervous tissue

Function of Nervous tissue

1. The function of nervous tissue is to transmit and receive external and internal stimuli.
2. Their function is to provide support, nutrition and insulation for the neurons.
3. Neurons are responsible for the reception, propagation and conduction of nerve impulses.

Activity

Observe the histological slides (permanent slides) of different types of animal tissues under microscope. Draw the diagrames along with the labeling and write comment on the corresponding tissue.

Note: *Don't get confused the stem with shoot. Stem refers to vascular part only whereas shoot refers to new fresh plant growth which includes stem and other structures like leaves or flowers.*

14.3 Stem

A stem is one of the major vascular part of a plant. It consists of nodes and internodes. Stems are solid, fistular, cylindrical, angular, etc. On the basis of position, stem can be divided into three types: aerial, sub-aerial and underground. Here, we will learn about internal structure of stem in both monocot and dicot plants.

Internal structure of Monocot stem

The transverse section of stem of a monocot plant contains following internal structures:

1. Epidermis
2. Hypodermis
3. Ground tissues
4. Vascular bundles

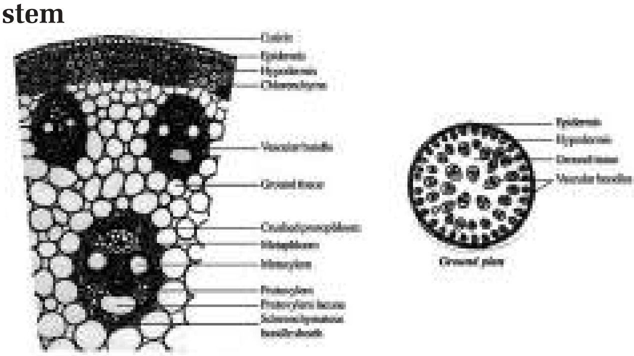


Fig: Internal structure of monocot stem

1. **Epidermis:** Epidermis is the outermost covering of the stem. It is made up of compactly arranged parenchyma cells. Intercellular spaces are absent. It is covered with a thick cuticle without epidermal stem hairs. The epidermis contains numerous minute openings called stomata.

2. **Hypodermis:** Hypodermis is a region that lies below the epidermis. It is made up of two or three layers of sclerenchyma cells. Intercellular spaces are absent. It provides mechanical strength to the plant.

3. **Ground Tissue:** Ground tissue is a major component of the stem. It consists of several layers of loosely arranged parenchyma cells. It is not differentiated into cortex, endodermis, pericycle and pith as in dicot stem. It contains reserve material and stores food.

4. **Vascular Bundles:** Vascular bundles are conjoint, collateral and closed. These are found irregularly scattered in the ground tissue. The bundles are smaller and towards the periphery and are larger in towards the centre. Each vascular bundle has a covering called bundle sheath formed by a single layer of sclerenchyma cells. The vascular bundle consists of xylem towards the inner surface and phloem towards the outer surface. Cambium is absent.

a) **Xylem:**Xylemconsists of vessels, tracheids, xylem parenchyma and xylem fibres. All four types of cells present in xylem are dead and hollow cells. In the xylem, there are two metaxylem and two protoxylem vessels arranged in

'Y' shape. A large water cavity is present in the inner side of the protoxylem. It stores water and also called lacuna or water cavity. The xylems are endarch with outer metaxylem and inner protoxylem.

b) Phloem: Phloem consists of sieve tubes, companion cells and phloem fibres. Phloem parenchyma is absent. The outer phloem is protophloem and is broken. Metaphloem lies in the inner portion and is functional.

Characteristics of Monocot Stem

1. Presence of thick cuticle, single layered epidermis
2. Epidermal hairs are absent.
3. Ground tissues are not differentiated into cortex and pit.
4. Ground tissue is not differentiated as endodermis and pericycle.
5. Epidermis contains stomata with guard cells.
6. Each vascular bundle is oval, conjoint, collateral and closed.

Internal structure of Dicot stem

The transverse section of stem of a dicot plant contains following internal structures:

1. Epidermis
2. Hypodermis
3. Cortex
4. Endodermis
5. Pericycle
6. Vascular bundles
7. Cambium
8. Medullary rays
9. Pith

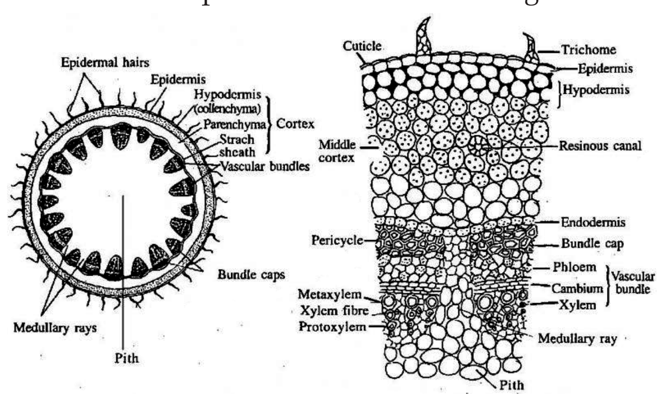


Fig: Internal structure of Dicot stem

1. **Epidermis:** Epidermis is the outermost layer of the stem. It is made up of single layer of compactly arranged parenchyma cells. The cells are thick walled and intercellular spaces are absent. It has multicellular epidermal stem hairs. Externally, it is covered by cuticle, which prevents excessive evaporation of water. The epidermis also bears stomata, which are involved in transpiration.

2. **Hypodermis:** Hypodermis lies below the epidermis. It is made up of four-five layers of compactly arranged collenchyma cells. Intercellular space is absent. The cells are thick walled due to the deposition of cellulose and pectin. It gives mechanical support. The hypodermis consists of chloroplast, which helps in photosynthesis.

3. **Cortex:** Cortex is made up of several layers of loosely arranged parenchymatous cell. Intercellular space is prominent. The cortex may contain chloroplast which helps in photosynthesis. It helps in gaseous exchange and storage of food materials.

4. **Endodermis:** Endodermis is the innermost layer of cortex. It is made up of compactly arranged parenchymatous cells. Intercellular spaces are absent. The endodermis is slightly wavy in appearance. The cells stores starch grains and also called as starch sheath.

5. **Pericycle:** Pericycle is the layer below the endodermis. It is made up of few layers of compactly arranged sclerenchyma and parenchyma cells. Above each vascular bundle, the pericycle forms a distinct cap-like structure known as bundle cap. The bundle cap helps in mechanical support.

6. **Vascular bundles:** Vascular bundles are arranged in ring. It is conjoint, collateral and open. Each bundle is composed of phloem in its outer surface and xylem in its inner surface on the same radius with a strip of cambium in between them.

   a) **Xylem** lies toward the pith. It is composed of xylem vessels, tracheids, xylem fibres (wood fibres) and xylem parenchyma (wood parenchyma). Protoxylem is smaller with spiral thickenings and metaxylem is wider with pitted vessels.

   b) **Phloem** lies outside of the vascular bundle and composed of sieve tubes, companion cells, phloem fibres and phloem parenchyma.

7. **Cambium** is a thin strip of two or three layered cells which are radically arranged. It lies between xylem and phloem. Cambium increases the thickness of the plant through secondary growth. So it is also called secondary meristem.

8. **Medullary Rays:** Medullary rays are parenchymatous cells. It is found between vascular bundles. It helps for the radial conduction of water and food materials. It is extension of pith, hence also called pith rays.

9. **Pith:** Pith is the central part of the stem. It is made up of loosely arranged parenchyma cells. Intercellular spaces are prominent. The pith stores the food.

Characteristics of Dicot Stem

1. Presence of well-defined epidermis with cuticle and multicellular epidermal stem hairs.
2. Presence of slightly wavy endodermis.
3. Pericycle is composed of parenchyma and sclerenchyma alternately as irregular patches.
4. Presence of a bundle cap above each vascular bundle.

5. Vascular bundles are conjoint, collateral and open and arranged in a ring surrounding the pith.
6. The center of pith is made from thin walled loosely arranged parenchyma. Pith rays or medullary rays are found between two vascular bundles.

Activity

Observe the permanent slides of internal structure of dicot and monocot stem under microscope. Draw the diagrames along with the labelling and write comment on the corresponding permanent slide.

Practical Activity

To study the internal structure of monocot and dicot stem.

Cut the transverse sections of the given monocot stem (maize) and dicot stem (pea). Put it in a watch glass, wash it and transfer fine section or the thinnest cut into another watch glass and cover it with safranin for 1-2 minutes. Pick up the section, wash it gently with water and keep on the glass slide. Put one drop of glycerine. Cover it with coverslip and observe under microscope and draw a diagram observed in the microscope.

14.4 Root

It is the second vascular part of the plant. It is typically found under-ground but it can also be aerial (eg. money plant, maize). It can also defined as the non-leaf and non–node bearing part of the plant. Root can be studied into different parts like; root cap, meristematic region, region of elongation, region of root hair and region of maturation. According to the origin and structure of root, it is of two types: tap root and adventitious or fibrous root and are present in dicot and monocot plants respectively. Here we will learn about the internal structure of monocot and dicot roots.

Internal structure of Monocot root

The transverse section of root of a monocot plant contains following internal structures:

1. Epiblema
2. Cortex
3. Endodermis
4. Pericycle
5. Vascular bundles
6. Conjunctive tissue
7. Pith

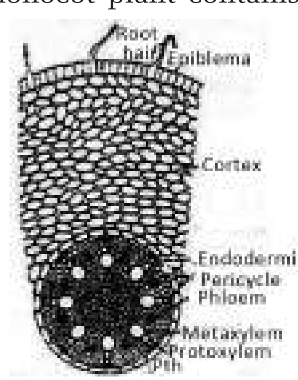


Fig: Internal structure of Monocot root

1. **Epiblema:** Epiblema is the outermost layer of the root without cuticle. It is made up of a single layer of compactly arranged parenchyma cells. The cells are thin-walled and are involved in absorption of water. Some of the epiblema cell consists of long unicellular projections called root hairs.
2. **Cortex:** Cortex is a major part of the ground tissue of root. It is made up of several layers of loosely arranged parenchyma cells. Intercellular spaces are prominent. It helps in the storage of water.
3. **Endodermis:** Endodermis is the innermost layer of the cortex. It is made up of compactly arranged parenchyma cell. Some of the cells in the endodermis are thin-walled and are known as passage cells. The passage cells passes water into the xylem vessels. The remaining cells in the endodermis are thick-walled. These thickenings are known as casparian thickenings. They are formed by the deposition of a waxy substance called suberin. The casparian thickenings help to create and maintain root pressure.
4. **Pericycle:** Pericycle is a layer below endodermis. It is made up of a single layer of parenchyma cells. It gives lateral roots.
5. **Vascular bundles:** Vascular bundles are found in radially, arranged in a ring.
   a) Xylem is composed of xylem vessels and tracheids.
   b) Phloem lies between xylem ridges. It is composed of sieve tubes, companion cells and phloem parenchyma.
6. **Conjunctive tissue:** Conjunctive tissue is made up of loosely arranged parenchyma cells. It is found in between the vascular bundles. The cell stores water.
7. **Pith:** Pith is large and well developed. It lies in the central region of the root. It is composed of few loosely arranged parenchyma cells.

Characteristics of a Monocot root

1. Presence of thin walled cells in the epiblema
2. Absence of cuticle and stomata
3. Presence of unicellular root hairs
4. Presence of passage cells and casparian thickenings in the endodermis
5. Presence of conjuctive tissue
6. Presence of a large pith
7. Presence of radial vascular bundles

Internal structure of Dicot root

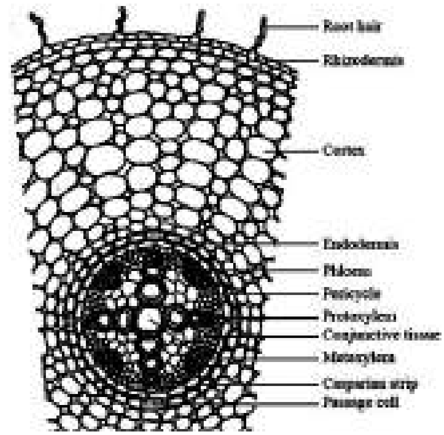


Fig: Internal structure of dicot root

The transverse section of root of a dicot plant contains following internal structures:

1. Epiblema
2. Cortex
3. Endodermis
4. Pericycle
5. Vascular bundles
6. Pith

1. Epiblema: Epiblema is the outermost layer of the root. It is a single layer made up of compactly arranged parenchymatous cells. The cells are thin walled and involved in absorption of water. It consists of long, unicellular projections called root hairs. It gives protection to the roots and also helps in absorption of water and minerals from soil.

2. Cortex: Cortex is thin walled, multilayered region made up of loosely arranged parenchyma cells. It has prominent intercellular spaces. The cortex transports water and minerals from the root hairs to the center of the root.

3. Endodermis: Endodermis is the innermost layer of the cortex formed by compactly arranged parenchyma cells. Intercellular space is absent. Some of the cells in the endodermis are thin-walled and are known as passage cells. The passage cells allow water to pass into the xylem vessels. The remaining cells consist of thickening on their radial walls. These thickenings are known as casparian thickenings. The casperian thickenings create and maintain root pressure.

4. Pericycle: Pericycle lies below the endodermis. It is made up of a single layer of parenchyma cells. Intercellular space is absent.

5. Vascular Bundles: Vascular bundles found radially and are arranged in ring.

a) Xylem consists of tracheids, vessels, xylem parenchyma and xylem fibres.

b) Phloem lies below the pericycle, alternating with xylem bundles. It consists of sieve tubes, companion cells and pholem parenchyma.

6. Conjunctive Tissue: Conjunctive tissue is made up of radially arranged parenchyma cells. It is found between the vascular bundles. The cells store water.

7. Pith: Pith is located in the central region with intercellular spaces. It is reduced in dicot root and absent in older roots. It helps to store food materials.

Characteristics of a Dicot Root

1. Presence of thin walled cells in the epiblema
2. Absence of cuticle, and stomata
3. Presence of unicellular root hairs
4. Absence of hypodermis
5. Presence of passage cells and casparian thickenings in the endodermis
6. Presence of conjuctive tissue
7. Absence of pith in older roots
8. Presence of radial vascular bundles

Activity

Observe the permanent slides of internal structure of dicot and monocot root under microscope. Draw the diagrammes along with the labelling and write comment on the corresponding permanent slide.

Practical Activity

To study the internal structure of monocot and dicot root.

Cut the transverse sections of the given monocot and dicot root. Put it in a watch glass, wash it and transfer fine section or the thinnest cut into another watch glass and cover it with safranin for 1-2 minutes. Pick up the section, wash it gently with water and keep on the glass slide. Put one drop of glycerine. Cover it with coverslip and observe under microscope and draw a diagram observed in the microscope.

14.5 Cell division

Introduction

The body of all the living organism is made up of cell. A cell is a small mass of cytoplasm which is bounded by a cell membrane and is capable of performing all the activities of the life. All organisms start their life as a single cell. As there are a lots of multi cellular organisms and unicellular organisms, it is an evident that a cell has an ability to multiply or reproduce. Similarly, as long as cell exists, it must perform all the metabolic activities. It is composed of nucleus and cytoplasm and there is a specific co-relation between cytoplasm and nucleus in a cell. With the growth and development of the cell, this relation get lost, which leads to the division of cell. The cell which is undergoing cell division is called mother cell and the cells formed after the cell division are called daughter cells. So, cell division can be defined as the process by which a mother cell divides to produce daughter cells. Cell division was first studied by **Prevost ad Dumas**

(1824 AD) in a zygote of frog.

Here in this chapter we will learn about the somatic cell, process and importance of mitosis cell division, germ cell and process and importance of meiosis cell division.

Cell cycle

The cycle of the changes that take place during cell growth and cell division is called cell cycle. According to the statement, “Omnis cellula & cellula” stated by Rudolf Virchow (1859AD), which means; every cell is derived from the pre-existing cell. Every organism is developed from a single cell. So cell divide to multiply. And the changes take place between two consecutive cell division are collectively called cell cycle. The division of nucleated cells is completed by two important activities such as Karyokinesisis (nuclear division) followed by cytokinesis (division of cytoplasm). Sometimes cytokinesis does not follow the karyokinesis, as a result multinucleated cells are formed.

It consists of three major stages: **Interphase, M-phase (mitotic or meiotic phase) and Cytokinesis.**

(A) Interphase

It is the stage between the end of one cell division to the beginning of another cell division. It is the longest phase of cell cycle. Any of the cellular structure does not divide in this phase that’s why this phase is also called resting phase. But it is actually metabolically active stage. All the necessary materials for the cell division are synthesized in this phase. Thus it is also called preparatory phase. The entire changes or the entire synthetic activities can be studied into three sub phases; **Gap one (G1/first growth phase), Synthetic (S-phase) and Gap two (G2/second growth phase).** The changes occurred in the cell in these respective sub phases are as follows:

1) G1 phase:

Cell size increases due to high rate of biosynthesis

RNAs, proteins, carbohydrates and lipids are synthesized

2) S-phase:

Replication of DNA

Synthesis of histone proteins take place

3) G2 phase:

Duplication of cell organelles like mitochondria, plastids, centrioles, etc take place

Proteins present in spindle fibres are synthesized

Energy required for M-phase are synthesized and stored in the form of ATP

(B) M-phase

Interphase is followed by M-phase. It is actually the phase at which nucleus of a cell divides (karyokinesis). Karyokinesis completes in four major stages; prophase, metaphase, anaphase and telophase. As a result of M-phase daughter nuclei are formed.

(C) Cytokinesis

M-phase is immediately followed by the cytokinesis. Once nucleus is divided then cytoplasm and other cell organelles are equally divided to form daughter cells.

The cell may divide by any of the following methods:

a) Direct cell division (Amitosis)

b) Indirect cell division (Mitosis and Meiosis)

Direct cell division (Amitosis)

In this method, the cell divides into two cells. A constriction occurs at the middle of the nucleus and finally nucleus divides into two nuclei and then again constriction occurs in the cytoplasmic membrane dividing cytoplasm. Finally two daughter cells are formed. During the nuclear division unequal distribution of chromosomes take place. Thus, the two daughter cells formed do not resemble with their parents as well as each other. Usually this type of cell division takes place in lower organisms like bacteria, some protozoans, yeast, etc.

Do you know?

Amitosis is frequently found in the fast growing cells of cancer.

Indirect cell division

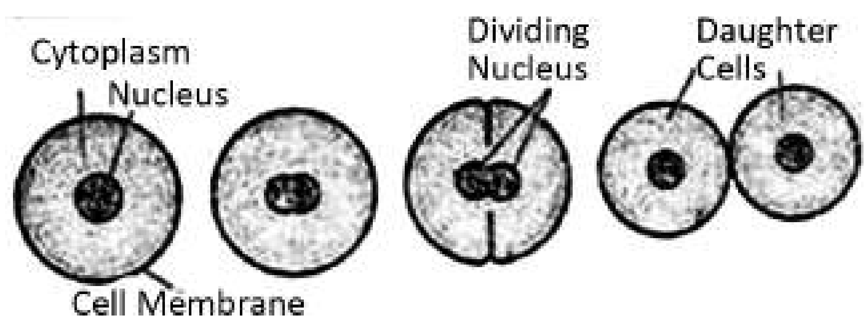


Fig. Amitosis

The process of indirect cell division includes two distinct types of cell division: Mitosis and Meiosis.

A) Mitosis

> *Do you know?*
> Uncontrolled and unregulated Cell Division can lead to Cancer

Mitosis is a form of eukaryotic cell division in which one mother cell divides into two daughter cells each having equal number of chromosomes to that of mother cell. Thus this type of cell division is also called equational cell division. It occurs in vegetative or somatic cells of an organism so it is also called somatic cell division. Somatic term is derived from the Greek word *soma* which means body. Hence, all body cells of an organism which help in the growth and development of the body apart from the sperm and egg cells are called somatic cells. In this cell division, mature somatic cell divides in such a way that daughter cells are quantitatively or qualitatively similar to the mother cell. In case of plants mitosis occurs in meristematic cells. All the multicellular organisms are developed from a single cell zygote by the process of mitosis. The term mitosis was coined by Walter Flemming in 1882 AD.

The process of mitotic cell division or mitosis can be studied in three major stages: Interphase, Karyokinesis and Cytokinesis.

I) Interphase

It is the stage between two consecutive cell division. Though this stage looks non-dividing or restive, this stage is very important and metabolically active stage in the cell division because all the necessary requirements or proteins are synthesized in this stage. The changes occur in the cell are as follows:

a) A cell has large nucleus with intact nuclear membrane and distinct nucleolus.

b) The chromosomes occur in the form of chromatic network.

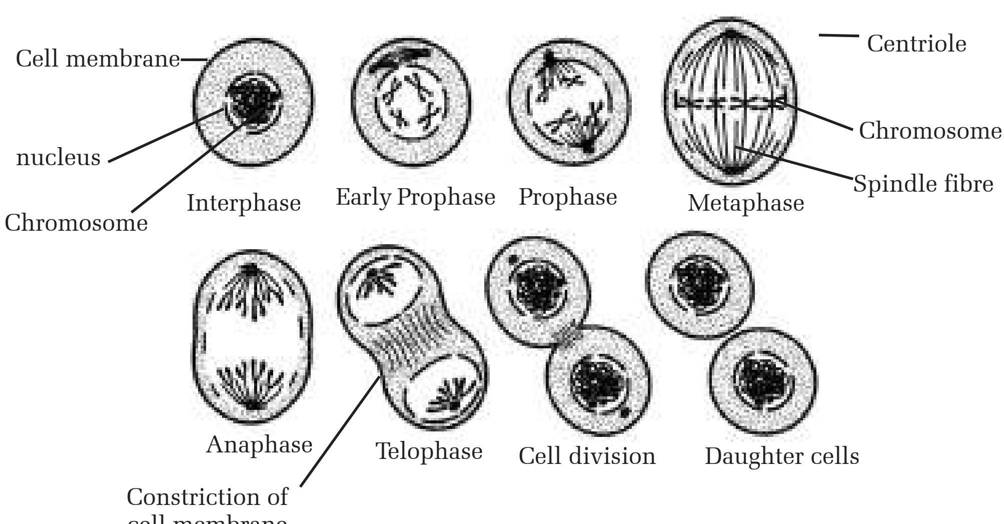


Fig: Mitosis cell division

c) DNA replication takes place and form exact two copies of DNA.
d) RNA and proteins (histone proteins and spindle proteins) are synthesized.
e) All cell organelles (mitochondria, plastid, centriole etc) are duplicated.

II) Karyokinesis

The process of division of nucleus is called karyokinesis. It can be studied in the following four phases; Prophase, Metaphase, Anaphase and Telophase.

Prophase: This is the first phase of nuclear division. The changes that take place in this phase are as follows:

a) The nuclear membrane and nucleolus start to disappear.
b) A structure known as the centrosome duplicates itself to form two daughter centrosomes called centriole that migrate to opposite ends of the cell.
c) The centrioles organise the production of microtubules that form the spindle fibres and also ray like structures around them called asters are developed.
d) The chromosomes condense into compact structures. Each replicated chromosome can now be seen to consist of two identical chromatids (or sister chromatids) held together by a structure known as the centromere.

Metaphase: This is the second phase of nuclear division. The changes that take place in this phase are as follows:

a) Formation of spindle fibre completed.
b) Nuclear membrane and nucleolus completely disappear.
c) Centrioles reach their respective poles.
d) All chromosomes are arranged in an equator in such a way that the centromeres lie on equatorial plane called metaphasic plate. This process of arrangement of all the chromosomes at the equator is called congression.

Anaphase: The changes that take place in this phase are as follows:

a) The centromere of each chromosome is splitted into two so that two sister chromatids have their own centromeres and daughter chromosomes are formed.
b) Repulsion between two sister chromatids and contraction of spindle fibre occur so that each daughter chromosomes with only one chromatid move towards opposite poles.
c) Daughter chromosomes appear as U, V, L and J during the migration of daughter chromosomes.

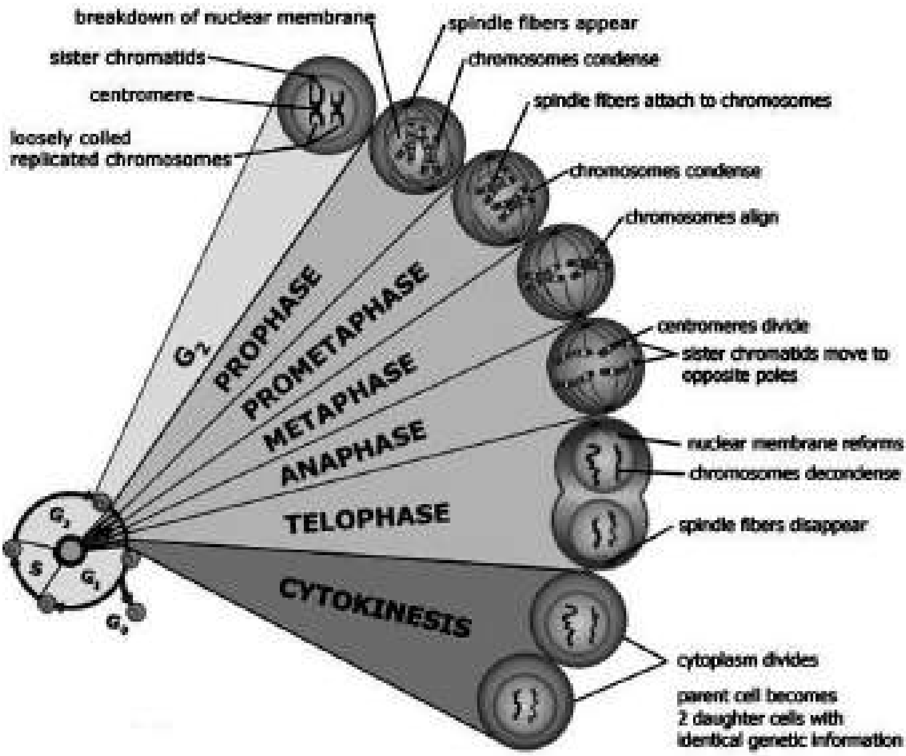
Telophase: This is the final stage of nuclear division. The changes that take in this phase are as follows:

a) Daughter chromosomes reach their respective poles.

b) The chromosomes elongate and overlap to form chromatin network.

c) Spindle fibres disappear, whereas nuclear membrane and nucleolus reappear.

d) At the end of telophase, two daughter nuclei are formed each having equal number chromosomes to mother cell.



III) Cytokinesis

Karyokinesis is followed by cytokinesis. After the division of nucleus into two identical daughter nuclei, remaining cell organelles and cytoplasm divides to form two identical daughter cells. And this process of division of cytoplasm and cell organelles is called cytokinesis. It takes place by two distinguished methods: cell plate formation method and furrowing/ cleavaging/constriction method.

a) **Cell plate formation:** In plant cells, cytokinesis occurs by cell plate formation method. In this process the granular bodies produced by golgi complex and microtubules are arranged in the equatorial region, form a cell plate and divide cytoplasm into two equal halves.

b) **Furrowing:** In this process, a peripheral constriction appears at the equatorial region which continues towards the centre and meets in the centre of the cell to divide cytoplasm into two equal halves.

Significance of mitosis

a) It helps in growth and development of zygote into adult in multicellular organisms.

b) It helps in reproduction in unicellular organisms.

c) It helps to produce new cells to replace the old cells, healing of wound, regenerating the lost part, etc.

d) It helps to maintain the genetic stability by maintaining equal number of chromosomes in daughter cells and mother cell.

> *Do you know?*
>
> Roughly 5×10^9 cells are lost from the surface of skin, alimentary canal, etc. daily in our body which are replaced by the new cells formed by mitosis cell division.

> **Activity**
>
> Observe the permanent slides showing different stages of mitosis under microscope and also draw the observed structures in your copy.

B) Meiosis

Meiosis is the form of eukaryotic cell division that produces haploid sex cells or gametes (which contain a single copy of each chromosome) from diploid cells (which contain two copies of each chromosome). It can also be defined as a type of cell division in which the number of chromosomes is reduced by half. It occurs only in certain special cells called reproduction cells or germ cells of the organisms. A germ cell is any biological cell that gives rise to the gametes of an organism that reproduces sexually. During meiosis, homologous chromosomes separate, and the haploid cells that have only one chromosome from each pair. The cell which undergo meiotic cell division is called meiocyte. Only half of the chromosomes are distributed from diploid mother cell to the daughter cells produced by meiotic cell division (sperm and ovum) and hence this cell division is also called reductional cell division. The term ‘meiosis’ was originated from greek word *meioum* which means to reduce and was given by **Farmer and Moore** (1905AD) whereas the process was first demonstrated by **Van Benden (1883AD)**. The process takes the form of one DNA replication followed by two successive nuclear and cellular divisions (Meiosis I and Meiosis II). As in mitosis, meiosis is preceded by a process of DNA replication that converts each chromosome into two sister chromatids.

Meiosis cell division takes place in reproductive cells at the time of gametes or spore formation and can be studied in the following three stages; Interphase, Karyokinesis and Cytokinesis.

I) Interphase

The changes take place in interphase of meiosis are similar to that which takes place in interphase of mitosis. But the replicating process in S-phase is longer than that occurs in mitosis.

II) Karyokinesis

In meiotic cell division, nucleus divides for twice. First, one diploid mother nucleus divides into two haploid daughter nuclei, called meiosis-I and then each haploid nucleus formed as a result of meiosis-I divides into two haploid daughter nuclei, called meiosis-II. As a result, four haploid nuclei are formed from one diploid nucleus at the end of meiosis.

Meiosis-I

It completes in four phases; Prophase-I, Metaphase-I, Anaphase-I and Telophase-I.

Prophase-I: It is a very complex and long phase. It is further divided into 5 sub-phases to make its study easier and they are Leptotene, Zygotene, Pachytene, Diplotene and Diakinesis.

1) Leptotene: The changes occur in this sub-phase are as follows:

a) Size of the nucleus increases.

b) Condensation of chromatin materials take place resulting thin, elongated, single stranded and thread like chromosome. Hence, called leptotene (slender ribbon).

c) Centrosome splits into two centrioles and each centriole start to move towards its respective pole.

d) Nuclear membrane and nucleolus remain intact.

2) Zygotene: The changes that take place in this sub-phase are as follows:

a) Shortening and thickening of chromosome continues.

b) Homologous chromosomes come near to each other and form a pair called bivalent. This process of pairing up of homologous chromosomes to form a bivalent is called synapsis.

3) Pachytene: The changes that take place in this sub-phase are as follows:

a) Each chromosome splits longitudinally into two sister chromatids. Hence from the structure of bivalent two chromosomes split to form 4 arms. This structure is now called tetrad.

b) Non sister chromatids overlap or coil with each other. During coiling non sister chromatids attach at some points. These points of contact between two non-sister chromatids are called chiasmata (singular: chiasma).

c) Crossing over begins. The process of exchange of genetic material from chiasmata in between two non-sister chromatids of homologous pair is called crossing over.

4) **Diplotene:** The change that take place in this sub-phase are as follows:

   a) Crossing over takes place.

   b) Nucleoprotein dissolves. Hence, non-sister chromatids start to separate from each other except in chiasmata.

   c) Nuclear membrane and nucleolus start to disappear.

5) **Diakinesis:** The changes that take place in this sub-phase are as follows:

   a) Chiasmata move towards the terminal end of the chromosome. This process of movement of chiasmata towards the end of the chromosome and disappear is called terminalization.

   b) Nuclear membrane and nucleolus completely disappear at the end of this phase.

   c) Spindle fibres start to appear.

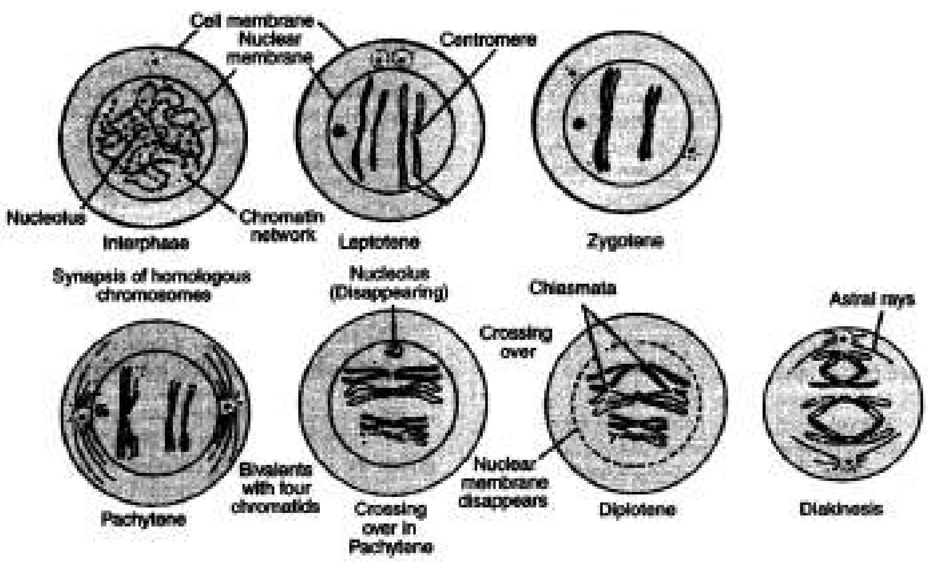


Fig: Prophase I of meiosis I

Metaphas-I: In this phase, the following changes take place.

a) The homologous pairs/homologous chromosomes arrange themselves at the equatorial region in such a way that the two members of a pair lie on opposite sides of the equatorial plane.

b) The chromosomes are attached to the spindle fibres by their centromeres.

Anaphase-I: The changes that take place in this phase are as follows:

a) Homologous chromosomes separate from each other and start to move towards their respective poles. This process of separation of homologous chromosome is called disjunction.

b) Centromeres of chromosomes do not break so each chromosome bears two chromatids. These chromosomes are called dyads.

c) Since centromeres do not divide, the number of chromosomes moving towards the poles is reduced into half of the mother nucleus. As a result, haploid nucleus is formed in each pole.

Telophase-I: The changes that take place in this phase are as follows:

a) The chromosomes become uncoil and form chromatin threads.

b) Spindle fibre and astral rays disappear.

c) Nuclear membrane and nucleolus reappear.

d) At the end of telophase-I, two haploid nuclei are formed. In some cases telophase –I may be totally absent.

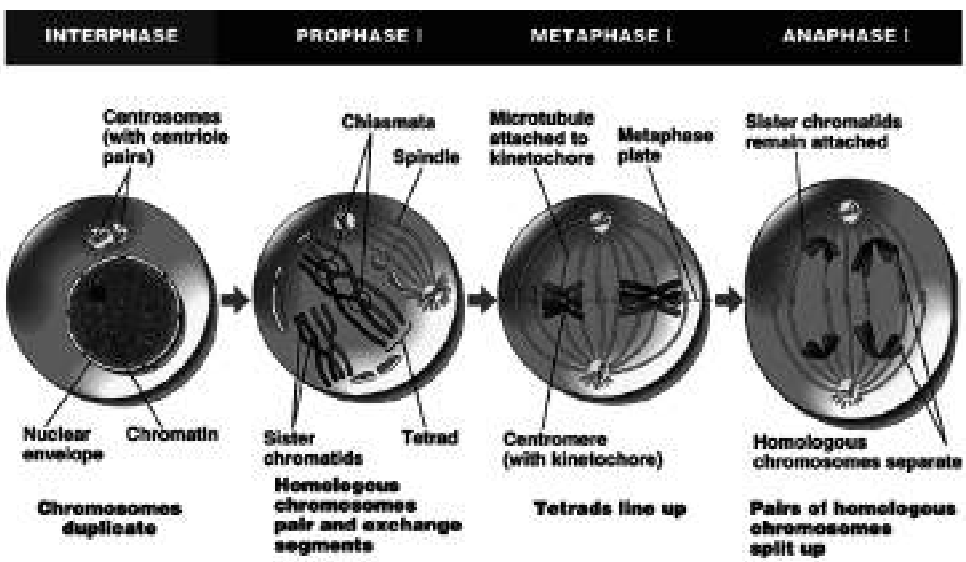


Fig: Different stages in Meiosis I

Cytokinesis-I

Telophase-I may or may not be followed by cytokinesis. If followed, it takes place through either cell plate formation or furrowing method which results in the formation of two haploid cells.

Interkinesis: It is the stage that occurs between telophase-I and prophase-II. It is similar to interphase but DNA replication process does not occur here.

Meiosis-II

All the changes occurs in this phase are similar to that which occurs in mitosis.

So this phase is also called **meiotic mitosis or homotypic division**. The difference is only haploid number of chromosomes are present. It is also divided into Prophase-II, Metaphase-II, Anaphase-II and Telophase-II.

Prophase-II: The changes that take place in this phase are as follows:

a) Dyad chromosomes become shorter and thicker.

b) The nuclear membrane and nucleolus start to disappear and spindle fibres and astral rays appear.

Metaphase-II: The changes that take place in this phase are as follows:

a) Nuclear membrane and nucleolus completely disappear.

b) Dyad chromosomes arrange themselves in an equatorial plane.

c) The centromere of each chromosome is attached with the spindle fibres from both the poles.

Anaphase-II: The changes that take place in this phase are as follows:

a) The centromere of each chromosome separates so two sister chromatids of dyad separates.

b) Separated sister Chromatids (now called daughter chromosome) move towards their respective poles.

Telophase-II: The changes that take place in this phase are as follows:

a) At each pole, chromosomes uncoil, elongate and form a chromatin network.

b) Spindle fibres disappear and nuclear membrane and nucleolus reappear.

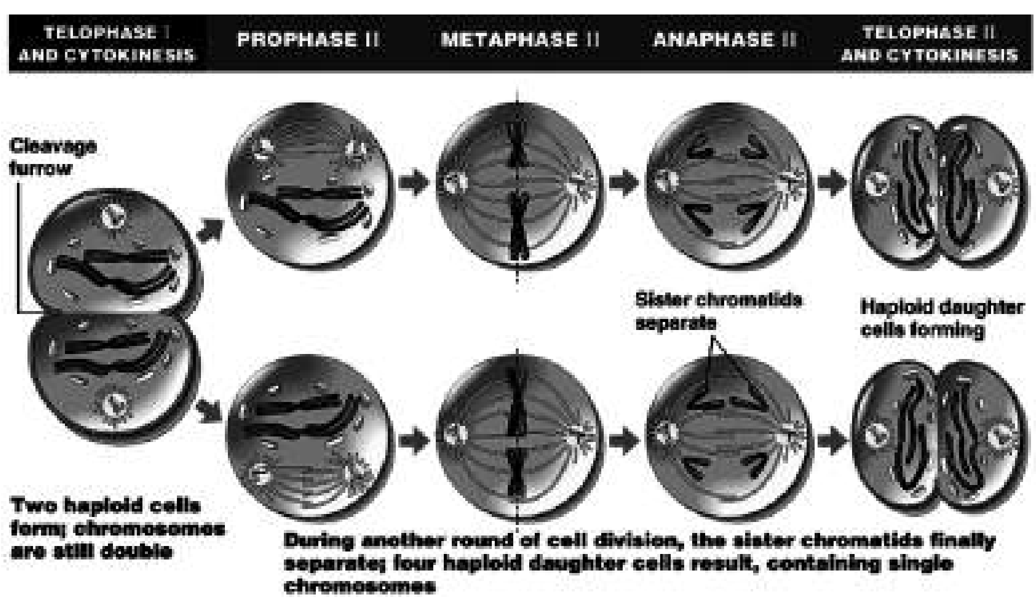


Fig: Different stages in Meiosis II

Cytokinesis-II: Telophase-II is followed by cytokinesis (cell plate formation or furrowing). As a result, four haploid nuclei are formed at the end of cytokinesis-II from a single diploid nucleus.

Significance of meiosis

1. Meiosis helps in the formation of haploid gametes or spores which are responsible for the sexual reproduction.
2. During the meiotic cell division, crossing over takes place. Which form a new combination of characteristics. These new characteristics lead to the variation and evolution.
3. It helps organisms to be better fitted in the environment.
4. It reduces the number of chromosomes during gamete formation (gametogenesis) so that the number of chromosomes after fertilization will be fixed/constant in the offspring. Thus, it helps to maintain the genetic stability in the offspring by keeping fixed number of chromosomes in the offspring.

Activity

Observe the permanent slides showing different stages of meiosis under microscope and also draw the observed structures in your copy.

Summary

1. Cellular biology is the branch of biology which deals with the study of cells.
2. A tissue is defined as a group or layer of cells of common origin that perform specific functions.
3. Epithelial tissues form the surface of the skin, lines many cavities of the body and cover the internal organs.
4. Compound epitheliumfunction as a protective covering, they cover dry surface of skin, moist surface of buccal cavity, pharynx and inner lining of salivary glands.
5. All connective tissues except blood secrete structural proteins called collagen or elastin.
6. Muscular tissues are composed of fine fibrils known as myofibrils.
7. Nervous tissue consists of neurons, neuroglia, ependymal cells and neuro secretory cells. They are responsible for sensing stimuli and transmitting signals to and from different parts of an organism.

8. The transverse section of stem of a monocot plant contains Epidermis, Hypodermis, Ground tissues and Vascular bundles.
9. The transverse section of stem of a dicot plant contains Epidermis, Hypodermis, Cortex, Endodermis, Pericycle, Vascular bundles, Medullary rays and Pith.
10. The transverse section of root of a monocot plant contains Epiblema, Cortex, Endodermis, Pericycle, Vascular bundles, Conjunctive tissue and Pith.
11. The transverse section of root of a dicot plant contains Epiblema, Cortex, Endodermis, Pericycle, Vascular bundles and Pith.
12. A cell is a small mass of cytoplasm which is bounded by a cell membrane and is capable of performing all the activities of the life.
13. The cell which is undergoing cell division is called mother cell and the cells formed after the cell division is called daughter cells. So, cell division can be defined as the process by which a mother cell divides to produce daughter cells.
14. The division of nucleated cells is completed by two important activities such as Karyokinesisis (nuclear division) followed by cytokinesis (division of cytoplasm).
15. Amitosis (direct cell division) takes place in lower organisms like bacteria, some protozoans, yeast, etc.
16. Mitosis is also called equational cell division because one mother cell divides to form two daughter cells each having equaled number of chromosomes to that of mother cell.
17. Meiosis can also be called as reductional cell division because the number of chromosomes is reduced by half.
18. A germ cell is any biological cell that gives rise to the gametes of an organism that reproduces sexually.

Exercise

A. Tick (√) the best alternatives from the followings:

1. The study of tissue is known as:

   a. Histology b. Biology

   c. Pathology d. Taxonomy

2. The stele consist of:

   a. Vascular bundles

   b. Vascular bundles, pericycle, pith and medullary rays

   c. Vascular bundles, pericycle, pith

   d. Vascular bundles and pith

3. Blood, bone, tendon, ligament, adipose and areolar tissues are the examples of:

   a. Epithelial tissue b. Cuboidal tissue

   c. Muscular tissue d. Connective tissues

4. The nuclei in cuboidal epithelium cell is located:

   a. Basal b. Central

   c. Apical d. Eccentric

5. Mitosis is Coined by:

   a. Farmer and Moore b.Lavoiser

   c. W. Flemming d. T.H. Morgan

6. The division of nucleus is called

   a.Karyokinesis b. Cytokinesis

   c. Both a and b d. None

7. How many meiotic divisions are necessary to produce 100 gametes?

   a. 100 b. 25 c. 200 d. 50

8. Which one is in the correct sequence of cell cycle?

   a. S, G1, G2, M b. S, M, G1,G2

   c. M, G1, G2, M d. G1, S, G2, M

9. Cytokinesis is a division of:

   a. nucleus b. cytoplasm

   c. protoplasm d. cell organelles

10. DNA replication occurs in:

a. G1 phase b. S phase

c. G2 phase d. M phase

B. Give short answer to the following questions:

1. What is tissue?
2. Write the location of simple squamous epithelium.
3. What are the components of nervous tissue?
4. Write the function of :

a. Epithelial tissue b. Connective tissue

c. Musculat tissue d. Nervous tissue

5. Name the phases of cell cycle.
6. What is crossing over?
7. Define cell division.
8. What is the basic difference between plants and animals mitosis cell division?
9. Write any two major changes that occur in anaphase of mitosis.
10. Mentionthe differences between the followings:

a) mitosis and meiosis

b) metaphase and anaphase

c) anaphase of mitosis and anaphase of meiosis

d) cytokinesis in animals and cytokinesis in plants

e) haploid cell and diploid cell

11. Explain the significance of meiotic cell division in animals.
12. Meiosis and mitosis both maintain the genetic stability. Explain how they maintain genetic stability.
13. Give reason:

a) Mitosis is also called equational cell division.

b) Meiosis brings out variation.

c) Cell division is a necessary phenomenon.

14. Draw a well labelled diagram showing metaphase and anaphase of mitosis.

C. Give long answer to the following questions.

1. Describe the type of epithelial tissue.
2. What are the types of connective tissue? Explain.
3. Give a short account about nervous tissue and muscular tissue.
4. Describe the internal structure of dicot stem.
5. Describe the internal structure of dicot root.
6. Describe the internal structure of monocot stem.
7. Describe the internal structure of monocot root.
8. Describe the somatic cell division and mention its significance.
9. Explain with diagram about the prophase-I.
10. Describe the changes that occur in the nucleus during anaphase of mitosis. How is it different from the anaphase-I of meiosis. Explain along with a diagram.

Project work

1. Bring a young stem of sugar cane and cut its thin transverse section. Then stain it with safranin for one minute. Then wash the section and put it in a clean slide. Observe it in the microscope.
2. Draw a neat and well labelled diagram of the roots and stems of monocot and dicot plant in a separate chart paper and colour it. Explain about the characteristics and structure of these parts in a classroom.
3. Take a white chart paper and draw all the stages of mitosis, colour them and label them. Then present a presentation on the different stages of the mitosis.
4. Take a white chart paper and draw all the stages of meiosis, colour them and label them. Then present a presentation on the different stages of the mitosis.

Glossary

Lignified: make rigid and woody by the deposition of lignin in cell walls

Shearing: to remove by cutting or clipping

Ultrafiltration: filtration using a medium fine enough to retain colloidal particles, viruses or large molecules.

Prominent: particularly noticeable

Dividing capacity: the tissue which can undergo mitotic cell division

Hydrophytes: the plants found in water (aquatic plants)

Fistular stem: the stem that consists of hollow cavity inside

Angular stem: the stem consisting of angular outlines

Exarch: the arrangement of vascular bundles where protoxylem is outside and metaxylem is inside

Endarch: the arrangement of vascular bundles where protoxylem is inside and metaxylem is outside

Sister chromatids: chromatids of same chromosome

Non-sister chromatids: chromatids of different chromosome

Homologous chromosome: the identical male and female parent chromosome that occurs in the form of pair

Dyad: a meiotic chromosome after separation of the two homologous

Unit 15

Life and Life Processes

Thomas Addison (1793-1860)

Addison was a renowned physician and scientist. Among other pathologies, he discovered Addison's disease (a degenerative disease of the adrenal glands) and Addisonian anemia (pernicious anemia), a hematological disorder later found to be caused by failure to absorb vitamin B_{12}.

15.1 Introduction

Life processes are the processes which occur throughout the life. For the survival of the organism, every organism show different metabolic activities like digestion, respiration, excretion, reproduction, etc. We will discuss about digestive system, respiratory system, urogenital system in rat, structure and life cycle of rice and endocrine system in human in this unit.

15.2 Rat

Introduction to rat

Rat is a mammal which has almost similar characteristics to the human. It has same type of circulatory system, similar muscles, and a similar skeletal structure. We will learn about the digestive system, respiratory system and urogenital system of rat in this unit.

Kingdom: Animalia
Phylum: Chordata
Sub-phylum: Vertebrata
Class: Mammalia
Type: Rat (*Rattus rattus*)

Digestive system of rat

All of the cells participate in a metabolic activities. During metabolic activities they release energy and to release energy they need food. The required food for all of the cells in an organism is obtained and processed by the digestive system. In digestive system the raw food is broken down into smaller molecules and those which cannot be broken down into simple particles, eliminated out through the anus.

The digestive system is a long tract found inside the animal, start from the mouth and ends with the anus, and helps in digestion.The process of enzymatic breakdown of complex food materials into the simplest form or absorbable form for cells is called digestion which occurs in alimentary canal. The simplest form of food then absorbed and transferred to all the other cells of the body through circulatory system and in each cell these simplest form of food release energy with the help of respiratory system. The pathway followed by the food for the digestion is called alimentary canal which includes the mouth, pharynx, esophagus, stomach, small intestine, rectum and anus. During this passage, there are several accessory glands like salivary gland, liver, pancreas, etc. which help in the process of digestion. The rat's digestive system is similar to the human's digestive system.

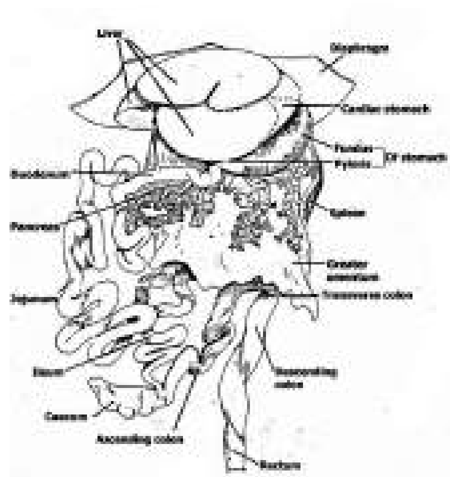


Fig: Digestive system in rat

Alimentary canal

Alimentary canal of rat is divided into two parts: Mouth and pharynx and Viscera.

A) Mouth and pharynx

Mouth: The mouth is the most anterior part of the digestive system. Within the mouth, oral cavity is present where crushing of food as well as mixing the food with saliva takes place. Tongue is also present which helps to move food and mixed food with saliva uniformly. Saliva contains various enzymes which help in digestion of carbohydrate. Sixteen numbers of teeth are also present in oral cavity of rat which are categorized as two types: incisors (4 in number) which help in cutting and molars (12 in number) which help in grinding.

Pharynx: A space at the back of the oral cavity is called pharynx which is divided into: Nasopharynx (receives air from the external nares and is not directly in the oral cavity), Oropharynx (space ventral to the soft palate) and Laryngopharynx (space posterior to the soft palate and anterior to the esophagus). From the oral cavity, only food diverges in the laryngopharynx and directed towards the esophagus.

B) Viscera

The other organs of the digestive system are located within the body cavities. All the organs of the body cavity, particularly those of the digestive system, are called the viscera. Viscera includes the following organs.

Stomach: The food passes from the esophagus into the stomach. It is bean-shaped sac like structure, which is partially covered by the liver. It acts as a storage organ so that fewer and larger meals can be consumed. Digestion of proteins begins in stomach.

The lining of stomach secrete pepsin which converts proteins into peptones. But the enzyme pepsin get activated only in an acidic environment and that acidic environment is created by the dilute hydrochloric acid secreted by cells lining of the stomach.

The muscular walls of the stomach churn the food and mix it with enzymes. During this time, sphincters (circular muscles) located at each end of the stomach prevent the food from escaping.

Small intestine: Food pass to the small intestine from the stomach. It is divided into duodenum, jejunum and ileum. The actual digestion and the absorption of the products occur here. The glands present near to it secrete enzymes which help to breakdown protein and carbohydrate. Liver secretes bile which emulsifies the fat and alkaline nature of bile inactivates the pepsin from the stomach and neutralicex the acid food. Pancreas secrete enzymes to breakdown fats, carbohydrates and proteins.

The walls of small intestine is made up of columnar epithelial tissues containing villi. These villi help in the absorption of digested substances and pass them on to the blood capillaries for distribution.

Caecum: Caecum is located at the junction of the small and large intestine. Ingested cellulose which cannot be digested in stomach and small intestine undergo microbial fermentation. Due to this special process to form caecum product rats produce a special feces. They again ingest this feces to digest for a second time. This behavior is called **coprophagy**.

All mammals do not have caecum. Humans have a short caecum provided with the appendix. The human caecum provides space for digestion, but does not have the microbes for cellulose fermentation.

Large intestine or colon: Running from the caecum, the colon ascends, crosses the abdominal cavity, and descends again. The large intestine of a rat is very similar in structure to a large intestine of a human. The large intestine is divided into three parts, the ascending, transverse, and descending colon. The colon connects posteriorly with the poorly differentiated rectum of the rat. The rectum connects the colon and the anus. The primary function of the large intestine is to absorb most of the water of the digestive secretions, conserving it for use within the body.

Accessory Glands

Salivary glands: There are three pairs of salivary glands (Parotid, sub-maxillary and sub-lingual gland). The saliva, as previously mentioned, contains enzymes (amylase or ptyaline), which begin the digestion of carbohydrates, and mucus, which moistens the food and sticks it together to facilitate swallowing.

Gastric glands: The glands which are present in the mucosal wall of stomach. They secrete gastric juice which contains dilute HCl, mucus, and the enzymes (pepsin, renin and lipase)

Liver: It is the large, reddish brown mass located at the posterior to the diaphragm. The liver has a great number of functions. However, its role in digestion is to produce bile, a substance that emulsifies fats (breaks them into minute droplets), making them easier to digest.

In humans, the bile is stored in the gall bladder before being released into the small intestine. However, the rat lacks a gall bladder. Therefore, the bile is released through a duct directly into the small intestine, where it acts.

Pancreas: The pancreas is an irregular mass of brownish glandular tissue in the mesentery dorsal to the stomach. It secretes pancreatic juice which contain enzymes (trypsin, chymotrypsin, peptidases, amylase, lipase, sucrase, maltase, lactase, etc.). Along with the enzymes it also secretes hormone, insulin, which passes directly into the circulatory system and is not involved with digestion.

Intestinal glands: These are the microscopic glands found in the mucosal wall of the small intestine. It secretes intestinal juice which contains the enzymes (erepsin, lipase, nuclease, etc.).

Digestion

We all know that rats have a varied diet. The digestive system of rat helps rats to digest anything they eat. Digestion begins in the mouth where food are torn into small pieces with the help of two types of teeth. In human, food is broken down mixing with saliva but mixing with saliva does not occur in rat's mouth. With the help of peristalsis (muscle contractions) food is transported to the esophagus and then to stomach. In stomach physical breakdown of the food completes along with mixing with the saliva and the hydrochlone (HCl) acid and the chemical breakdown of the food starts. Now food is moved towards the small intestine, which is divided into three parts duodenum, jejunum and ileum. In duodenum, the food get mixed with bile juice which inactivates the acidic effect of HCl and emulsifies fat. Similarly, the pancreatic juice digest protein, fat and carbohydrate in the duodenum which continues in the jejunum. Then the digested food or nutrients are absorbed in the small intestine (ileum). Villi and microvilli present in the walls of small intestine help in the absorption of nutrients and the absorbed nutrients are then transported to the rest of the body through the blood stream or circulatory system.

Do you know?

A rat digestive system has 2 major differences with that of a human.
1) Rats do not have a gallbladder.
2) Rats have an enlarged caecum where the bacterial fermentation of the cellulose (present in grains and seeds) can take place.

The undigested food (cellulose) and waste materials now come to the rat's caecum where fermentation of the plant cellulose takes place due to the bacteria present over there. The caecum again absorbs water and minerals. After absorption of most of the nutrients and minerals in small intestine and caecum, the remaining are transported to the large intestine (colon). The water again get reabsorbed in the large intestine and prepares the solid fecal material for removal from the body. Feces now moves to the rectum, stores there and eliminate out through the anus.

> **Activity:** Observe the way of dissection of rat done by your teacher. Then observe the digestive tract of the rat after the dissection done by your teacher. Draw the observed structure, label it and make a short note on digestion in rat.

Respiratory system of rat

The respiratory system of rat is similar to that of most other mammals, including humans. The respiratory system is used to inhale oxygen, oxidation of food and exhale carbon dioxide. The rat respiratory system consists of nostrils, lungs, a trachea and a larynx.

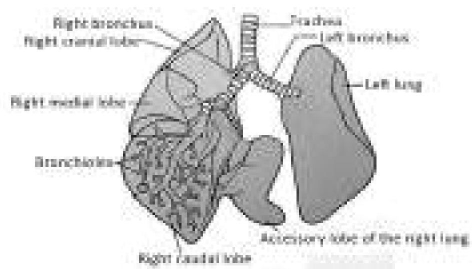


Fig: Respiratory system in rat

Nostrils, Glottis and Pharynx

Rats inhale and exhale oxygen gas through their nostrils. The nose is separated into two nostril cavities by a septum. The nostril cavities are two small openings on the nose, separated by a septum. Nostrils open into nasopharynx close to the glottis. As a rat is inhaling, the glottis, a small flap in the rodent's throat, closes to prevent food particles from entering the pharynx. This prevents the rat from choking on food particles.

Trachea and Bronchi

The trachea is a tube, in the upper portion of the rat's chest that connects the pharynx and nasal cavities to the lung. The trachea (also called windpipe) is provided with the rings of cartilage which prevent it from collapsing. Closer to the lungs, this pipe branches out into two tubes that are called bronchi. A right bronchous and a left bronchous connect the trachea to each corresponding lung cavity. Each bronchous connects to smaller bronchioles inside the lung cavity.

Larynx and Diaphragm

Larynx has no role in respiration but it is essential because it produces sound which rat uses to communicate. A passage of air is required to produce sound. The larynx get tightened or loosened to create the squeaks and other noises common to rats.

The diaphragm is a sheet of muscle present just below the lungs separating abdominal cavity and thoracic cavity. It helps to increase or decrease the volume of thoracic cavity and also to exert pressure on the lungs during breathing.

Lungs and Ribs

The lungs are the primary organ of the respiratory system. Rats have two lungs, one on either side of the rat's heart. Left lung of the rat contains only one lobe whereas right lung contains four lobes. Each lung is composed of numerous branches of bronchiole and each bronchiole ends with a microscopic unit called alveoli. Alveoli is the unit where the actual exchange of inhaled oxygen gas and released carbon dioxide gas takes place.

Lungs are protected with a membrane called pleural membrane or pleura. Pleura is filled with a pleural fluid which protects the lungs from mechanical injury and friction during the contraction and relaxation of lungs. Again the complete structure of respiratory tract are protected inside the rib cage.

Respiration in a Rat

Respiratory system of rat is as complex as humans. With the help of nostrils air is inhaled, which travels through the respiratory tract (trachea, bronchi, bronchioles) and come the alveoli. In alveoli carbon dioxide gas produced from the oxidation of food during releasing energy is already present. These two gases get exchanged in alveoli. Oxygen gas is then absorbed by the blood (red blood cells) and transported to different parts of the rat's body with the pumping of heart whereas carbon dioxide gas follow the same tract (bronchioles, bronchi, trachea, nostrils) and exhaled out.

Blood carries oxygen to the cellular level where simple food materials (converted in digestion) is oxidized to release carbon dioxide, water with the release of energy in the form of ATP. The energy released is utilized by rat for different activities.

Activity: Observe the way of dissection of rat done by your teacher. Then observe the respiratory tract of the rat after the dissection done by your teacher. Draw the observed structure, label it and make a short note on respiration in rat.

Urogenital system of rat

The excretory/urinary and reproductive systems of rat are closely integrated and are usually studied together as the urogenital system. But both the systems have their individual functions. Excretory system/urinary system helps in removal of the wastes while reproductive system helps in the formation of gametes (ovum and sperm), bringing ovum and sperm together, provide nourishment for the embryo or fetus until birth and regulation of the hormones for the development of sexual characteristics.

Urogenital organs of rat includes excretory/urinary organs and genital organs.

Excretory organs: Excretory organs in rats include a pair of kidneys, a pair of ureter, urinary bladder and urethra.

Kidneys: There are two kidneys in rat. Each kidney is a dark brown, bean shaped structures and made up of millions of nephrons (structural and functional unit of kidney). Kidneys help in the removal of nitrogenous waste products of protein like creatinine, uric acid, urea, ammonia and controlled amount of water and salts. The function of kidney occurs in the nephron. Nephron is made up of a cup shaped structure called Bowman's capsule, blood vessels (afferent arteriole and efferent arteriole) and glomerulus.

Glomerulus is a group of capillary loops that hangs into the Bowman's capsule. Blood driven with the high pressure in glomerulus due to which some nutrients like glucose, amino acid, etc., waste materials like creatinine, urea, uric acid, etc., some minerals and water get filtered in the Bowman's capsule. Which then travels through uriniferous tubules (proximal convoluted tubule, Henle's loop, distal convoluted tubule and collecting duct). During this passage, essential nutrients, water and minerals get reabsorbed by the blood, whereas waste materials now called urine come to the renal pelvis.

Ureters: There are two ureters, each originate from a kidney. It extends up to the urinary bladder. It transports the urine collected in the renal pelvis of the kidneys to urinary bladder.

Urinary bladder: Rat has one urinary bladder, where the collection of urine takes place.

Urethra: It is a canal like structure which extends from the urinary bladder and leads to the exterior. Its length differs in the female and male. It opens out through urethral orifice (vulva) which lies just above the vaginal orifice in female but in male it opens at the tip of the penis (urogenital orifice).

Genital organs/Reproductive organs

A) Genital organs of the male rat

The genital organs of the male rat includes the following parts:

1. **Testes (singular testis):** A pair of testes lies embedded in the scrotal sacs. The scrotal sacs are found in the form of pouches lying in front of the anus, between the hind limbs. Spermatogenesis occurs in epithelial lining within the testes. Testes also help in the secretion of steroidal hormone like testosterone and progesterone.

2. **Epididymis:** It is an irregular and convoluted tube lying along the inner edge of the testis. The epididymis helps in the storage of the sperm.

3. **Vas differentia:** Each vas deferens arises from the epididymis. It lies on the inner side of the

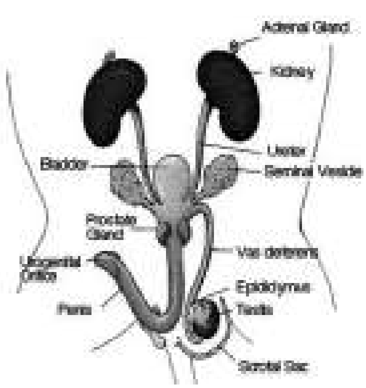


Fig: Genital organs of the male rat

testis and opens into the urethra. They carry sperm to the penis.

4. **Urethra:** It is the common duct for sperms and urine and opens into the penis.
5. **Spermatic cord:** It is an elongated cord like structure made up of connective tissues, blood vessels and nerves. It arises from the epididymis.
6. **Accessory glands:** Male genital system includes ampullary glands, seminal vesicles, prostate glands, Cowper's gland, etc. as accessory glands.

B) Genital organs of the female rat

The genital organs of the female rat includes the following organs:

1. **Ovaries:** The ovaries are small and compact yellowis pea sized bodies situated on the outer side of the kidneys.
2. **Fallopian tubes:** The fallopian tubes are narrow coiled and convoluted tubes which take matured ovaries to the uteri.
3. **Uteri:** The posterior ends of the fallopian tubes become thickened to form uteri. Both the uteri meet in the middle and form a common tube, vagina.
4. **Vagina:** It is a common tube formed by the fusion of the uteri form the both sides. It opens through the vaginal orifice called vulva.

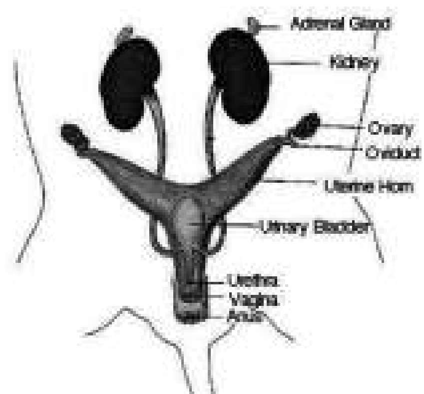


Fig: Genital organs of the female rat

5. **Clitoris:** It is a small rod like organ lying at the anterior end of the vulva.
6. **Accessory glands:** Female genital system includes Bartholin's gland, preputial glands as accessory glands.

The urogenital system in rat and human is approximately similar to each other. Only the structural differences are found.

> ***Do you know?***
> Duplex uterus (two uterine horns) is present in rat whereas only one uterus (simplex uterus) is present in human.

> **Activity:** Observe the way of dissection of rat done by your teacher. Then observe the urogenital tract of the rat after the dissection done by your teacher. Draw the observed structure, label it and make a short note on urine formation and reproduction in rat.

15.3 Life Cycle of a Rice Plant

The rice plant (*Oryza sativa*) is an annual grass that grows rapidly. It can mature from seed to a crop in three to six months depending on the variety and environmental condition. Rice is semi-aquatic plant and is commonly grown in flooded fields. It is a staple food for a large part of the world's human population, especially in East and Southeast Asia.

Kingdom: Plantae
Division: Tracheophyta
Sub-division: Angiosperm
Class: Monocot
Type: Rice (Oryza sativa)

Rice is a monocot plant. It has fibrous root, elongated leaves with parallel venation with hollow stem. The structure of root and stem of the rice is similar with the structure of monocot root and monocot stem as we discussed in previous unit.

About Rice Plants

Distribution: The rice plant is cosmopolitan in distribution. The plants are

commonly found in temperate regions and are also found in tropical and sub-tropical regions.

Habit: Annual or perennial shrubs, cultivated.

Root: Adventitious or fibrous

Stem: Erect, solid, cylindrical presence of distinct nodes and internodes, the stem is called culm.

Leaf: Alternate, simple, extipulate, sessile, leaf base forming tubular sheath, sheath open, surrounding the internodes completely, hairy or rough, linear, parallel venation.

Floral diagram:

Floral Diagram

Floral formula: Br%$P_{2or3}A_{3or6}G_{(1)}$

Fruit: A caryopsis with pericarp completely united with the seed coat.

Seed: Endospermic, with a single cotyledon

> ***Do you know?***
> Brown rice is a good source of magnesium, phosphorus, manganese, selenium, iron, folic acid, thiamine and niacin.

Rice varieties can be categorized into two groups as the short-duration varieties (matures in105–120 days) and the long-duration varieties (mature in 150 days). A 120-day variety, when planted in a tropical environment, remains around 60 days in the vegetative stage, 30 days in the reproductive stage, and 30 days in the ripening stage.

The life cycle of the rice plant is divided into three growth stages:

1. Vegetative growth stage
2. Reproductive growth stages
3. Ripening growth stages

Germination

The seed of a rice plant consists of an outer protective coat (hull); the embryo (which forms the actual plant); and the endosperm (starchy part that provides food for the seed). When the seed is planted in a flooded rice field, it will germinate; i.e., the first shoots and roots start to emerge from the seed and the rice plant begins to grow.

1. Vegetative growth stage

In the vegetative stage, tillers and leaves develop and the height of the plant increases gradually. The vegetative stage depends on the variety of rice, but is typically between 55 and 85 days.

a) Early Vegetative Stage

The early vegetative stage starts when the first root and shoot emerge and last just before the first tiller appears. During this stage, the roots grow down into the soil beneath the water. The roots hold the plant and absorb nutrients from the soil. Leaves also start to emerge and begin to grow toward the surface of the water. One leaf develops in every 3/4 days during early stage.

b) Late Vegetative Stage

During this stage, the plant is mature and reaches a height of 3 to 4 feet. It develops a reproductive stem called a tiller around 40 days after sowing. The stem stops growing in height at the end of this stage, eventually producing a flower head.

2. Reproductive phase

The flower head, commonly referred to as a spike, will then produce 100 to 150 tiny flowers. Each flower is small and forms along several long, thin stems that radiate upward from the tiller. These flowers, once pollinated, will then form seeds. This usually takes around 30 days to complete.

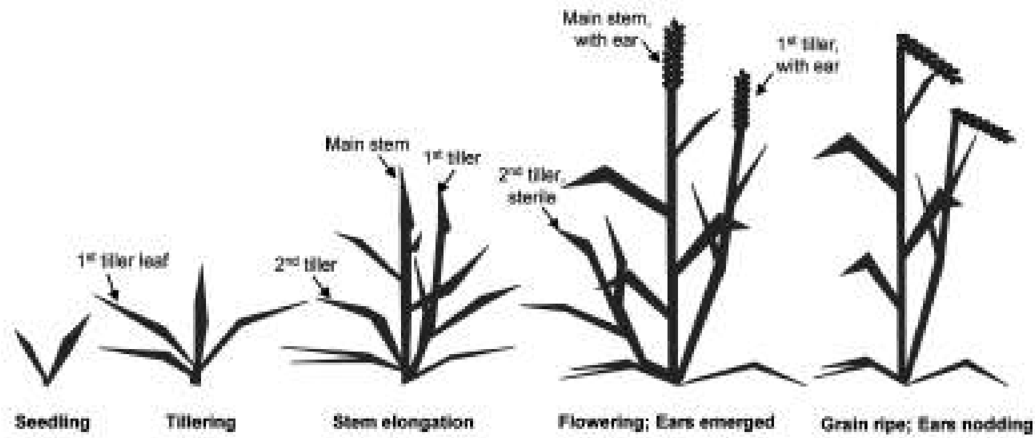


Fig: Different stages in rice plant

The first sign that the rice plant is getting ready to enter its reproductive phase is a bulging of the leaf stem that conceals the developing panicle, called the 'booting' stage. Then the tip of the developing panicle emerges from the stem and continues to grow. Rice is said to be at the 'heading' stage when the panicle is fully visible. Flowering begins a day after heading has completed. As the flowers open they shed their pollen on each other so that pollination can occur. Flowering can continue for about 7 days.

3. Ripening growth stage

The ripening phase starts at flowering and ends when the rain is mature and ready to be harvested. This stage usually takes 30 days. Rainy days or low temperatures may lengthen the ripening phase, while sunny and warm days may shorten it. The last three stages of growth make up the ripening phase.

Ripening follows fertilization and can be subdivided into milky, dough, yellow, ripe, and maturity stages. These terms are primarily based on the texture and color of the growing grains. The length of ripening varies among varieties from about 15 to 40 days. Ripening is also affected by temperature, with a range from about 30 days in the tropics to 65 days in cool temperate regions.

15.4 Endocrine system

The endocrine system is the collection of glands that produce hormones that regulate metabolism, growth and development, function of tissue, reproduction, sleep, and mood.

Endocrine glands are the glands of the endocrine system that secrete hormones directly into the blood. They are also called as ductless glands as they do not contain duct and each of these glands secrete chemical substances called hormones directly into the blood stream which are transported to every part of the body through blood circulatory system and help to regulate the activities of various organs. The secretion of one gland may influence the activity of the other glands. So, endocrine glands act with the dynamic balance. If one gland becomes overactive, the balance is upset and the other glands check it and then the other gland become overactive.

In our body exocrine glands are also present which are ducted gland, present near from the site of action and secrete enzymes.

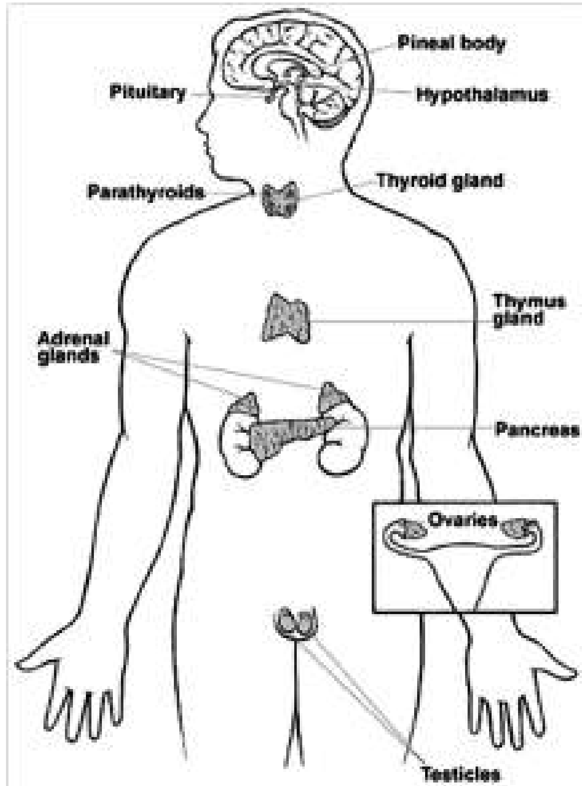


Fig: Endocrine glands

Hormones

Hormones are chemical messengers which are transported to the respective site by blood circulatory system and helps in the co-ordination and integration of different body activities. Hormones have the following characteristics:

1. They are produced in small quantities but their effects are long lasting.
2. They act as chemical messengers as they act on target cells to control and regulate metabolic activities, growth and development of a body.
3. They are transported to the target organs by blood.
4. They are composed of amino acids, steroids, etc.
5. They cannot be stored as they are destroyed easily.
6. They are soluble in water and blood and can easily diffuse through the cell membrane.

Hormones have following functions:

1. They regulate growth and development of the body and various metabolic activities.
2. They control reproductive activities like gametogenesis, development of the sexual characteristics, etc.
3. They conserve water and minerals in the body.

Endocrinology is the branch of science which deals with the structure, secretion and the functions of different endocrine glands. Thomas Addison is considered as the 'father of endocrinology'. He was the first who reported a disease due to malfunction of adrenal gland known as Addison disease.

The major glands of the endocrine system include: Pituitary gland, thyroid gland, parathyroid gland, pineal gland, pancreas, gonads and adrenal glands.

A) Pituitary gland

Pituitary gland is only about 1/3 of an inch in diameter (pea-sized) and located at the base of the brain. It is often called as the "master gland" as its hormones control other parts of the endocrine system. The pituitary gland is divided into three sections: the anterior, intermediate, and posterior lobes.

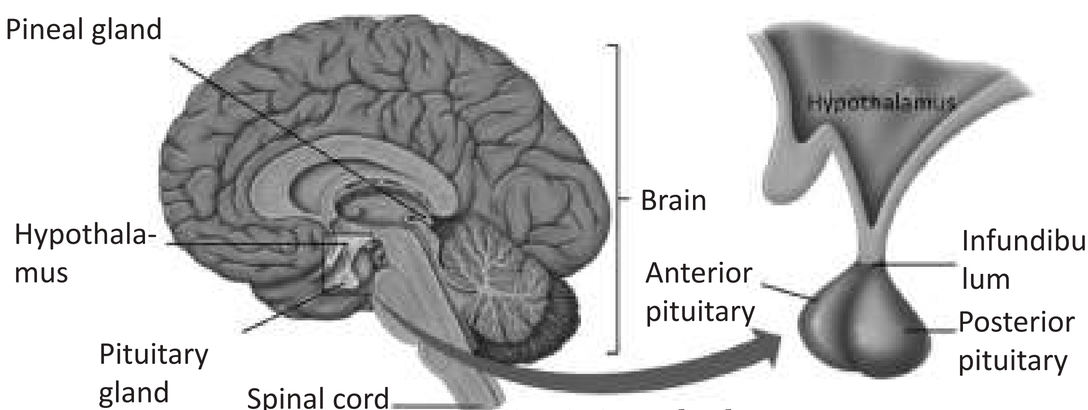


Fig: Pituitary glands

The anterior lobe is mainly involved in development of the body, sexual maturation, and reproduction. Hormones produced by the anterior lobe regulate growth and stimulate the adrenal and thyroid glands, as well as the ovaries and testes. It also generates prolactin, which enables new mothers to produce milk.

The intermediate lobe of the pituitary gland releases a hormone melanocyte stionulating normone (MSH) that stimulates the melanocytes, cells which control pigmentation — like skin color — through the production of melanin.

The posterior lobe produces antidiuretic hormone, which reclaims water from the kidneys and conserves it in the bloodstream to prevent dehydration. Oxytocin is also produced by the posterior lobe, aiding in uterine contractions during childbirth and stimulating the production and release of milk.

Here are the hormones secreted by the anterior and intermediate lobe of the pituitary gland.

1. **Growth hormone (GH) or Somatotropic hormone:** It stimulates the growth of all tissues of the body, especially skeletal muscle and bone. GH mobilizes fats, stimulates protein synthesis, and inhibits glucose uptake and metabolism. GH is vital for normal physical growth in children; its levels rise during childhood and peak during the puberty.

   Deficiency/Hyposecretion

   a) Dwarfism: small height and sexual immaturity

   b) Simmond's diseases: loss of body weight

   Hypersecretion

   a) Gigantism: extra ordinary height of body (more than 7.5 feet)

   b) Acromegaly: abnormal increase in size of bones of face, hand and feet giving gorilla like appearance.

2. **Thyroid stimulating hormones (TSH):** TSH stimulates the thyroid gland to produce thyroxine (T4) and triiodothyronine (T3). Thyroxine and triiodothyronine are essential to maintain the cellular metabolic rate, heart and digestive functions, muscle control, brain development and maintenance of bones.

   Thyrotropin-releasing hormone (TRH) stimulates its release.

3. **Adrenocorticotropic hormone (ACTH)** stimulates the adrenal cortex to release corticosteroids.

4. **Gonadotropin hormone (GTH):** The follicle stimulating hormone (FSH) and luteinizing hormone (LH) regulate the functions of the gonads in both sexes.

   a) FSH stimulates secretion of oestrogenhormone in female. In male, it stimulates formation of sperms (spermatogenesis).

   b) LH stimulates ovulation, secretion of progesterone hormone in females and androgen hormones in male.

5. **Prolactin hormone (PRL)** stimulates the growth of breast during pregnancy and promotes milk production in the mammary glands after child birth.

6. **Melanocyte stimulating hormone (MSH):** MSH stimulates the production and release of melanin pigment (melanocytes). It is secreted by the intermediate lobe of the pituitary gland.

Similarly, the hormones secreted by the posterior lobe of the pituitary gland are as follows:

1. **Oxytocin:** It promotes the contraction of uterine muscles during child birth in pregnant woman and hence this hormone is also called birth hormone. It also helps in stimulation of mammary glands to secrete milk. In males, it helps in transportation and ejection of sperm.
2. **Vasopressin or Anti-Diuretic Hormone (ADH):** It controls the permeability of distal convoluted tubule (DCT) and collecting duct (CD) to absorb water. It causes urine hypertonic. Hyposecretion: diabetes insipidus.

B) Thyroid gland

The thyroid gland is a butterfly shaped gland located in the anterior throat below larynx. It is richly supplied with blood vessels; hence it is red in colour. The thyroid regulates our metabolism through the action of thyroid hormone, utilising iodine from the blood and incorporating it into thyroid hormones. The thyroid gland secretes two hormones: thyroxine and calcitonin.

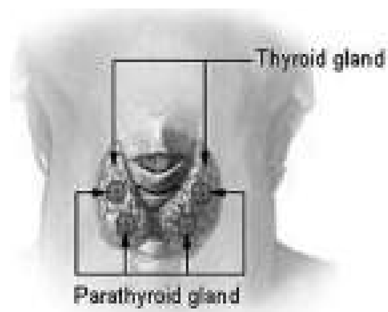


Fig: thyroid and parathyroid gland

Function of thyroxine

1. It regulates the production of body heat by regulating respiration.
2. It regulates metabolic rate.
3. It regulates mental and physical development.
4. It helps in absorption of glucose from intestine.

Hyposecretion: Hypothyroidism

1. Cretinism: retarded physical, mental and sexual growth, slow heart beat, low blood pressure, decrease body temperature
2. Myxoedema or Gull's disease: fatty and sluggishness in a person, slow heart beat, low body temperature and retarded sexual growth
3. Goitre (Simple goiter): enlargement of thyroid gland, thus swelling in neck
4. Hashimoto disease: an autoimmune disease in which the thyroid gland is destroyed by autoimmunity.

Hypersecretion: Hyperthyroidism

1. Grave's disease (Exopthalmic goiter or Parry's disease): bulging of eye balls, loss of body weight, increase heart beat, body temperature, restlessness

Function of Calcitonin

It maintains the calcium level in blood. It is secreted when calcium level in blood is high and helps in absorption of calcium by bone. Deficiency of calcitonin leads to the decrease of calcium level in bone by the deposition of calcium in blood. So, ultimately weakening of bones.

C) Parathyroid glands

The parathyroid glands are four tiny glands, located in the neck, two on each side. It controls the body's calcium levels. The parathyroids produce a hormone called parathyroid hormone or parathormone (PTH).

Function: Parathormone regulates calcium metabolism and controls the amount of calcium in blood and bone.

Hypo secretion:Hypoparathyroidism

Less secretion of parathormone due to accidental damage to parathyroid lowers concentration of calcium ions in the blood and tissues due to excretion of calcium in the urine.

Hyposecretion of parathormone also raises the level of phosphate ions in the plasma by reducing phosphate excretion.

Hyper secretion: Hyperparathyroidism

Over-activity of the glands results in more extraction of calcium from the bones, which in turn cause softening, bending and fracture. This condition is called osteoporosis.

The rise in blood calcium may cause deposition of calcium in the kidney tubules causing renal stones and kidney failure.

D) Adrenal glands (Suprarenal glands)

The adrenal glands or suprarenal glands are paired glands located above each kidneys in humans and in front of the kidneys in other animals. Each adrenal gland contains two parts: outer part called cortex and inner part called medulla. Adrenal cortex secretes mineralocorticoids, glucocorticoids and gonadocorticoids, whereas adrenal medulla secretes adrenaline (epinephrine) and noradrenaline.

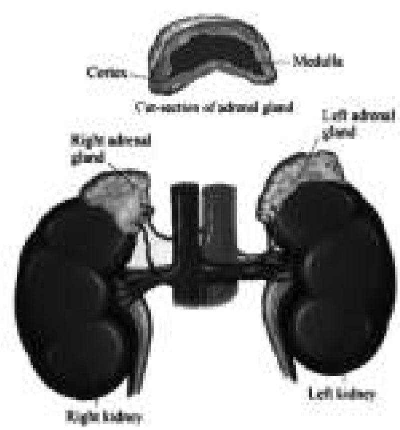


Fig: Adrenal glands

1) **Aldosterone (mineralocorticoids)** maintains water and electrolyte balance and volume of the blood in the body.

Function

- It increases the reabsorption of sodium from urine, saliva, bile and sweat to reduce its loss from the body.
- It also increases the excretion of potassium in exchange of the reabsorbed sodium.
- It also increases the reabsorption of water from the urine by raising the osmotic pressure of the blood through reabsorption of Na^+ in to it.

2) Glucocorticoids hormones regulate the metabolism of carbohydrates, proteins and fats.

Function

- These hormones convert proteins and fats into carbohydrates, which in turn is converted to glucose. This increases the glucose level in blood.
- In case of excessive bleeding, glucocorticoids hormone constrict blood vessels and balance the blood pressure due to blood loss.
- These hormones show anti-inflamatory and antiallergic effects.

3) **Sexcorticoids** or Gonadocorticoidssecrete androgen hormone in male while oestrogen and progesterone hormone in female. These hormones stimulate the development of secondary sexual characters like body hairs, menstrual cycle, enlargement of breast, change of voice in male child, etc.

Hypo secretion:

1) **Addisons disease:** This is characterised by ion imbalance, which lowers water retention. It is caused by deficiency of mineralocorticoids. It causes low blood sugar, low sodium ion concentration and high potassium ion concentration in the plasma which result in loss of weight, weakness, nausea, vomiting and diarrhea.

2) **Conn's disease:** It is a disease caused by hypo secretion of mineralcorticoids. In this disease nervous disorder occurs which lead to convulsions and death.

Hypersecretion

1) **Cushings syndrome:** It is caused by excess of glucocorticoides in blood. It is characterised by increase in blood sugar, blood pressure, excretion of sugar in urine, obesity, wasting of limb muscles, etc.

2) **Aldosteronism:** It is caused by excessive secretion of aldosterone. It is characterized by high sodium ion concentration and low potassium ion concentration in the plasma, increase blood volume and blood pressure.

3) **Adrenal virilism:** It is caused by excessive secretion of sex corticoids. In females, excess androgen production leads to develop male secondary sexual characters such as beard, moustache and hoarse voice.

Adrenal medulla secretes adrenaline and noradrenaline. Adrenaline and noradrenaline together prepare our body to cope with the emergency conditions like fear, anger, blood pressure, emotions, heartbeat, etc. Secretion of adrenaline is increased during cold, hot, drugs and emotional excitement. It also increases the rate of conversion of stored glycogen in liver into glucose. Adrenaline helps to dilate the blood vessels. Similarly the major function of noradrenaline is to control the blood circulation and to increase the oxygen consumption.

Adrenal gland is called emergency gland because it helps to cope the body to face emergency conditions resembled by 3F's where 3F's are fear, flight and fight respectively.

E) Pineal gland

The pineal gland is located near the center of the brain. It is very small, reddish-grey organ shaped like a pine cone. It secretes melatonin hormone.

1. The production of this hormone is increased in darkness and decreased in bright light. Melatonin hormone helps to control sleep patterns.
2. Melatonin blocks the secretion of gonadotropins (sex hormones) from the anterior pituitary gland.

F) Pancreas

The pancreas is located in the abdomen behind the stomach. It acts both as exocrine gland, that helps in digestion and endocrine gland that regulates blood sugar by secreting enzymes and hormones respectively. Hence, pancreas is also called mixed gland or dual gland or heterocrine gland.

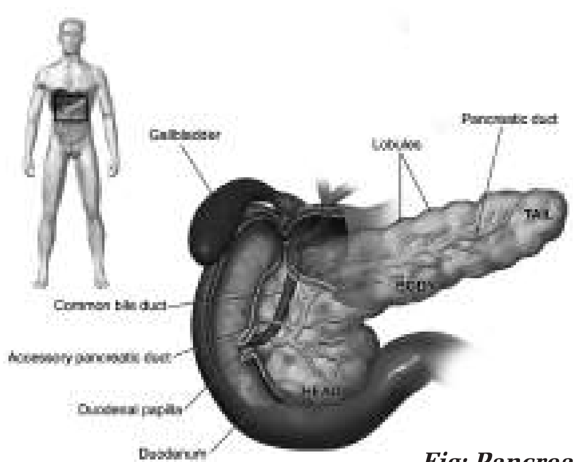


Fig: Pancreas

The endocrine system of the pancreas consists of islet cells (islets of Langerhans) that secrete and release insulin, glucagon and somatostatin hormones directly into the bloodstream. Insulin lowers the blood sugar level while glucagon raises the blood sugar.

Function of insulin: When the blood level of glucose rises, insulin is released by beta (β) cells and increases the rate of glucose uptake and metabolism by most body cells. It increases the protein metabolism. It also stimulates deposition of extra glucose in the form of glycogen in liver and muscle.

Do you know?

Somatostatin is released by Delta cells and act as an Inhibitor of Growth Hormone, Insulin and Glucagon.

Function of glucagon: When the blood glucose level is low, glucagon is released by alpha (α) cells and stimulate the liver to release glucose into the blood.

Function of somatostatin: It is secreted by the delta cells. It regulates the secretion of both insulin and glucagon.

Disorders

1) Hyposecretion of insulin results in diabetes mellitus; the symptoms include polyuria, polydipsia, and polyphagia.

2) Hypersecretion of insulin or deficiency of glucagon results in hypoglycemia(low blood glucose level). The symptoms include weakness, sweating, irritability and convulsions.

g) Gonads

A gonad produces the gametes (sex cells) of an organism. The male gonad, the testis, produces sperm and the female gonad, the ovary, produces egg cells.

The ovaries of the female, release two main hormones: Oestrogens and Progesterone and the testis of male secrets testosterone and androgen hormone.

Functions

1. Secretion of oestrogen hormone stimulates maturation of the female reproductive organ, oogenesis and development of the secondary sexual characteristics.
2. Progesterone hormone maintains the menstrual cycle. It also thicken the uterine wall for the maintenance of pregnancy.
3. Testosterone hormone secreted by testis in male promotes maturation of the male reproductive organs, stimulates spermatogenesis, development of secondary sex characteristics, and production of sperm.

Summary

1. Rat is a mammal which has almost similar characteristics to the human.
2. The digestive system is a system found inside the animal, start from the mouth and ends with the anus, and helps in digestion.
3. The process of enzymatic breakdown of complex food materials into the simplest form or absorbable form for cells is called digestion.
4. The respiratory system is the system which is used to inhale oxygen, oxidation of food and exhale carbon dioxide. The rat respiratory system consists of nostrils, lungs, a trachea and a larynx.
5. Excretory system/urinary system helps in removal of wastes while reproductive system helps in formation of gametes (ovum and sperm), bringing ovum and sperm together, provide nourishment for the embryo or fetus until birth and regulation of the hormones for the development of sexual characteristics.
6. The life cycle of the rice plant is divided into three growth stages: Vegetative growth stage, Reproductive growth stages and Ripening growth stages.
7. Rainy days or low temperatures may lengthen the ripening phase, while sunny and warm days may shorten it.
8. The endocrine system is the collection of glands that produce hormones that regulate metabolism, growth and development, function of tissue, reproduction, sleep, and mood.
9. Pituitary gland is often called as the "master gland" as its hormones control other parts of the endocrine system.
10. Growth hormone (GH) or Somatotropic hormone stimulates the growth of all tissues of the body, especially skeletal muscle and bone.
11. The thyroid gland secretes two hormones: thyroxine and calcitonin.
12. The parathyroids produce a hormone called parathyroid hormone (PTH).
13. The adrenal glands produce a variety of hormones including adrenaline and the steroids aldosterone and cortisol.
14. Sexcorticoids or Gonadocorticoids secrete androgen hormone in male while oestrogen and progesterone hormone in female.
15. The endocrine system of the pancreas consists of islet cells (islets of Langerhans) that secrete and release insulin and glucagon hormones directly into the bloodstream.
16. The ovaries of the female, release two main hormones: Oestrogens and Progesterone and the testis of male secrets testosterone hormone.

Exercise

A. Tick (√) the best alternatives from the followings:

1. The organs of the body cavity, particularly those of the digestive system is termed as:

   a. Viscera b. Vestigle

   c. Axon d. Colons

2. Protein is converted into peptones by:

   a. Pepsin b. Amylase

   c. Lipase d. Lactase

3. Panicle of rice develops in:

   a. Vegetative growth stage b. Late vegetative stage

   c. Reproductive growth stages d. Ripening growth stages

4. Long duration rice varities matures in .... days.

   a. 120 b. 150

   c. 140 d. 160

5. Endocrine glands are called as

   a. Duct glands b. Emergency glands

   c. Ductless gland d. Mixed glands

6. Excessive secretion of insulin results in

   a. Hypoglycemia b. Hyperglycemia

   c. Both d. None

B. Give short answers to the following questions.

1. Define digestive system.
2. What do you understand by coprophagy?
3. What is the role of gastric glands in digestion?
4. What is the role of pleural fluid?
5. What is spermatic cord?
6. List the growth stages in the life cycle of rice?
7. What do you mean by booting stage?

8. Define endocrine system.
9. Why is endocrine glands also called ductless gland?
10. List the major parts of pituitary gland with two hormones secreted by each.
11. Give reason: Pituitary gland is often called as the “master gland”.
12. List the hormones secreted by the anterior, intermediate, and posterior lobes of pituitary glands.
13. Write the function of:

    a. Thyroid gland   b. Parathyroid gland

    c. Adrenal glands   d. Gonads

    e. Pineal gland   f. Pancreas

C. Give long answers to the following questions.

1. Describe the digestive system of rat
2. Explain the mechanism of respiration in rat.
3. What are the roles of organs of excretory system of rats?
4. Give a short account in male and female reproductive organs. Describe the life cycle of rice plant.
5. Describe the structure of rice plant.
6. Where is pituitary gland located? Write down the hormones secreted by different lobe of this gland along with the function.
7. How pancreas helps to maintain the blood glucose level? Explain.

Project work

1. Draw a neat diagram of internal structures of rat (digestive tract, respiratory tract and urogenital tract). Use some colours or different waste materials to decorate it and label all the important parts. Present your work to your class.
2. Take a complete plant of mature rice and observe its structure. Also draw a neat and labeled diagram to show its complete life cycle.
3. Prepare a model of digestive tract, respiratory tract and urogenital tract of the mouse with the help of your seniors, guardians or teachers by using the materials found in your surroundings and present in your classroom.

Glossary

Soft palate: a continuation of the tissue lining the roof of the oral cavity

Epiglottis: a flap of tissue that blocks the larynx when food or fluid is in the laryngopharynx

Bowman's capsule: dilated end of a kidney tubule that surrounds a knot of capillaries

Glomerulus: ball-like network of capillaries that is surrounded by Bowman's capsule at the proximal end of a renal tubule

Loop of Henle/Henle's loop: that portion of the renal tubule of mammals and some birds that loops into the medulla of the kidney, and is essential for establishing the interstitial salt gradient needed for the production of concentrated urine

Staple food: a food that is eaten routinely in such quantities that it constitutes a dominant portion of a standard diet for a given people, supplying a large fraction of energy needs

Hypertonic: the solution having more amount of solute

Polyuria: a condition usually defined as excessive or abnormally large production or passage of urine (frequent urination)

Polydipsia: increased thirst

Polyphagia: excessive hunger or increased appetite

Melatonin: a neurotransmitter, which is mainly derived from the cells in the gastrointestinal tract, the retina, and the pineal gland and is responsible for maintaining sleep wake cycles, biological rhythms, and the modulation and inhibition of melanin synthesis

Melanin: one of the major pigments found in human skin, which determines the skin color

Spermatogenesis: process of formation of sperms

Oogenesis: process of formation of ovum

Unit 16

Heredity

Bateson introduced the term genetics to describe the study of heredity.Bateson co-discovered genetic linkage with Reginald Punnett and Edith Saunders, and he and Punnett founded the Journal of Genetics in 1910. Bateson also coined the term "epistasis" to describe the genetic interaction of two independent loci.

William Bateson (1861–1926)

16.1 Introduction

Genetics is the branch of science which deals with the study of heredity and variation. It deals with the similarities and differences between the different genus or different species of same genus of organisms. Offspring are produced either asexually or sexually from their parents. So, some of the characteristics are transmitted from parents to the offspring unchanged. This phenomenon of transmitting parental characteristics to the offspring is called **Heredity. Variation**, on other hand, makes some differences in the structure or characteristic of an organism among the same species or among the different species of a same genus. Which leads to the evolution. Variation is necessary to adapt in the changing environment. So, we can also say that the characteristics of the organisms depend on the surrounding or environment. That's why we can see different types of organisms in Himalayan region, Terai region and so on. The environmental factors and heredity, both, influence the development of an individual.

We have learnt some information about heredity in previous levels. Here, in this chapter we will learn about the Mendel's experiment (dihybrid cross), Mendel's law of inheritance, chromosome, sex linked inheritance and the process of formation of twins.

16.2 Mendel's Experiment

In sexual reproduction, an offspring is born by the fusion of male gamete and female gamete. A zygote is composed of a chromosomes from mother as well as from father. If mother and father have different characteristics then which one characteristic will be seen in an offspring? This question made curious to Gregor Johann Mendel (1822-1884), an Austrian monk. That's why he made many experiments to study about the transmission of hereditary characters and variation. On the basis of his experiments, he explained about the genetics and gave the natural law in hereditary characteristics. For this reason he is also called as the father of genetics.

Mendel selected 7 pairs of characteristics of pea plant (*Pisum sativum*) for his experiment. He chose pea plant for his experiments due to the following reasons:

a) They have very short life cycle. So, several generations can be obtained or experimented with in a single year.

b) They are bisexual plants. That's why they can be either self-pollinated or cross pollinated.

c) They produce large number of offspring after each generation.

d) They contain large number of pair of contrasting characteristics. Eg: tall-dwarf, axial-terminal, etc.

e) They are easy to cultivate.

And, the 7 pairs of contrasting characteristics are as follows:

a) size of the plant: tall and dwarf

b) position of flowers: axial and terminal

c) shape of the mature seeds: round and wrinkled

d) colour of the seed: green and yellow

e) colour of the flower: purple and white

f) shape of the mature pods: inflated and constricted

g) colour of the mature pods: yellow and green

We will learn about dihybrid cross and third law of inheritance (law of independent assortment) here in this unit.

Mendel's law of independent assortment (law of free recombination)

The law of independent assortment states that,"The different characters in a hybrid union are inherited independently and when two pairs of traits are followed in the same cross, they assort independently".

This law can be studied by the experiment on dihybrid cross experimented by Mendel, which is given below;

Dihybrid cross

To study how different characteristics would behave in relation to each other in their inheritance among several generations, Mendel crossed two varieties of pea plants which were differing in two pairs of contrasting characteristics. Such crosses which are made in between two pairs of contrasting characteristics are known as dihybrid cross.

Mendel selected two pairs of contrasting characteristics; seed surface and seed colour. For his dihybrid cross, Mendel chose two pure pea plants, one having round and yellow seed (RRYY) and the other having wrinkled and green seed (rryy). Then he cross pollinated the plants and got the F_1-generation. Where he found all the offspring plants of F_1-generation having round and yellow seeds but in a hybrid form (RrYy). On the basis of this result, he concluded that yellow colour is dominant colour over green in seed colour and round surface is dominant over wrinkled surface in seed surface.

For F_2-generation, he allowed the plants of F_1-generation to self-pollinate. According to the Mendel's law of purity of gametes each reproductive part (male or female) produce four types of gametes: RY, Ry, rY and ry for F_2-generation. When they were crossed with each other, sixteen combinations for offspring were possible but the union of these gametes form nine classes of genotypes and four classes of phenotypes which are shown below in a table (Punnett Square).

Observation

Mendel found in his experiment that the factors or genes of one pair in the dihybrid get separated and assort independently with the factors or genes of another pair. And, as a result of the above cross, 4 different types of pea plants were formed (round with yellow seed, round with green seed, wrinkled with yellow seed and wrinkled with green seed) in the ratio of 9:3:3:1. The above mentioned experiment show that each pair of alleles strictly segregate and is independent of each other which demonstrates the law of independent assortment.

The entire process of dihybrid cross is represented in the given phylogenetic chart:

Parents: Round and yellow seed × Wrinkled and green seed

Genotype: (RRYY) (rryy)

↓ ↓

Gametes: RY ry

F1-generation: RrYy (Hybrid plant with round and yellow seed)

Possible-gametes for F2-generation: RY Ry rY ry

F2-generation in Punnett Square (after self-pollination)

	RY	Ry	rY	ry
RY	RRYY	RRYy	RrYY	RrYy
Ry	RRYy	RRyy	RrYy	Rryy
rY	RrYY	RrYy	rrYY	rrYy
ry	RrYy	Rryy	rrYy	rryy

The phenotypic and genotypic ratio of the offspring obtained in the F2-generation are as follows:

Number of individuals	**Genotype class**	**Phenotype class**
1 2 2 4	RRYY RrYY RRYy RrYy	Homozygous Yellow Round Heterozygous Yellow Round (Hybrid)
		Yellow Round =9
1 2	rrYY rrYy	Homozygous Yellow Wrinkled Heterozygous Yellow Wrinkled
		Yellow Wrinkled = 3
	RRyy Rryy	Homozygous Green Round Heterozygous Green Round
		Green Round = 3
1	rryy	Homozygous Green Wrinkled
		Green Wrinkled = 1
16		9:3:3:1

Result

As a result of the dihybrid cross, Mendel found the phenotypic ratio 9:3:3:1 (9 round with yellow, 3 round with green, 3 wrinkled with yellow and 1 wrinkled with green). Similarly he found the genotypic ratio 1:2:2:4:1:2:1:2:1 shown in the above table.

Conclusion

In F_2-generation, besides the original two parental combinations of round with yellow and wrinkled with green, two new combinations, round with green and wrinkled with yellow, were also formed. Which shows that the characteristics can arrange themselves in new combination independently. And, this is the law of independent assortment.

This dihybrid experiment suggest that there is no rule that any two or three characteristics should always remain together. They can separate from each other and can form a new combination of characteristics. That's why hybridization is in practice, in agricultural field, so that different varieties of advantageous characteristics are formed in farm animals and agricultural crops.

Limitation

This law is not applicable everywhere. There are cases in which two or more characteristics are transmitted together rather than independently. For example: if the genes are present very near to each other in a same chromosome then they will move together as a unit. This is called linkage.

Activity

Draw a phylogenetic chart upto F_2 generation in a chart board, when a pure plant having round seeds with inflated pods first cross pollinated with wrinkled seeds with constricted pods and then self-pollinated in offspring. Also show the possible results. Explain among the classmates.

16.3 Sex linked inheritance in human

It is true that, when two pairs of contrasting characteristics are used in a cross, they combine independently, as stated in Mendel's law of independent assortment. It is also true that a single chromosome contains many genes and those genes are linked together hence inherited together. Those genes are called linked genes and the phenomenon is called linkage. If these genes are linked in a sex chromosome then it is called sex linkage. Therefore sex linkage can be defined as the continuous association of genes located on the sex chromosome and are inherited together with the sex chromosome from generation to generation. It was first introduced by Bateson and Punnet in 1906. When they were working (making cross) in sweet pea (*Latharus sativus*), they observed that two pairs of alleles were not combining independently and finally concluded that because Mendel chose characteristics controlled by genes located on different chromosomes, he did not notice the phenomenon of linkage.

Human cell is composed of 23 pairs of chromosomes. Among 23 pairs, 22 pairs are autosomes whereas 1 pair is sex chromosome. Sex chromosomes have genes for non-sexual characteristics along with the genes for sexual characteristics. These genes of non-sexual characteristics are transmitted from one generation to another generation along with the sex chromosome which are called sex linked characteristics, the genes are called **sex linked genes** and the inheritance of such non sexual characteristics along with sex chromosome is called **sex linked**

inheritance or sex linkage. Sex linked inheritance can also be defined as the inheritance of a trait (phenotype) that is determined by a gene located on one of the sex chromosomes.

Male and female are determined by the sex chromosome. Remaining autosomes are similar in male and female. Females are homogametic (sex chromosomes are similar), i.e. **XX** whereas males are heterogametic (sex chromosomes are different), i.e. **XY.** The genes for sex linked characteristics are either located in X or Y or on both. On the basis of its location, sex linked inheritance is classified into three types:

a) **X-linked characteristics/traits:** If the genes for non-sexual characteristics are located in X-chromosome. Example: Haemophilia. These type of characteristics are transmitted to the male whereas females are carrier.Many of the non-sex determining X-linked genes are responsible for abnormal conditions such as haemophilia, Duchenne muscular dystrophy, congenital night blindness, male pattern baldness, etc.

b) **Y-linked characteristics/traits:** If the genes for non-sexual characteristics are located in Y-chromosome. Example: Hairy ear (Hypertrichosis). These types of characteristics are transmitted from male to male.

c) **X-Y linked characteristics/traits:** If the genes for non-sexual characteristic are located in both homologous chromosomes X and Y. Example: Complete colour blindness, skin cancer, etc.

There are about 120 sex linked genetic disorders have been identified till now. Among those the most common sex linked traits are colour blindness and haemophilia. These sex linked traits which occur as a disorder are also called sex linked diseases. We will discuss about the colour blindness and haemophilia in this chapter.

Sex linked inheritance in Human

Meiosis is the process of making gametes, also known as eggs and sperm in most animals. During meiosis, the number of chromosomes is reduced by half, so that each gamete gets just one of each autosome and one sex chromosome.

Females make eggs, which always have an X chromosome. And males make sperm, which can have an X or a Y. Egg and sperm fuse to make a zygote, which develops into a new offspring. An egg plus an X-containing sperm will make a female offspring, and an egg plus a Y-containing sperm will make a male offspring. The process can be summarized as follow:

1. Female offspring get an X chromsome from each parent
2. Males get an X from their mother and a Y from their father
3. X chromosomes never pass from father to son

4. Y chromosomes always pass from father to son

The characteristic features of inheritance for a sex- linked trait can be summarized as follows:

1. The genes responsible for the sex linked traits are not transmitted from male parent directly to their male progeny because the progeny also receives X-chromosome from its mother whereas Y-chromosome from its father.
2. A male transmits his sex linked genes to all his daughters since daughters receive one X-chromosome from their father which get further transmitted to daughter's male progeny. i.e. All sex-linked genes therefore pass from male to female and then come back to a male of F2-generation.

   Suffer (Father) $\xrightarrow{\text{Pass trait from daughter}}$ Grand son
3. Carrier mother transmits her sex linked genes to her male progeny which get further transmitted to son's female progeny. i.e. All recessive sex linked genes pass from female to male and then come back to a female of F2-generation.

The cause of sex-linkage can therefore be shown with two reasons:

(i) The location of a gene in X chromosome

(ii) The absence of its allele in the Y chromosome.

For example, in human, the long arm of Y chromosome is homologous to the short arm of X chromosome. Genes located in that region of X chromosome does not show sex linked inheritance.

A few genes have been discovered in the non-homologous region of the Y-chromosome, which have no allele in X chromosome. Such genes follow an entirely different pattern of inheritance. For example, the incidence of hypertrichosis (long hair growth in the ears) is found due to Y-linked genes.

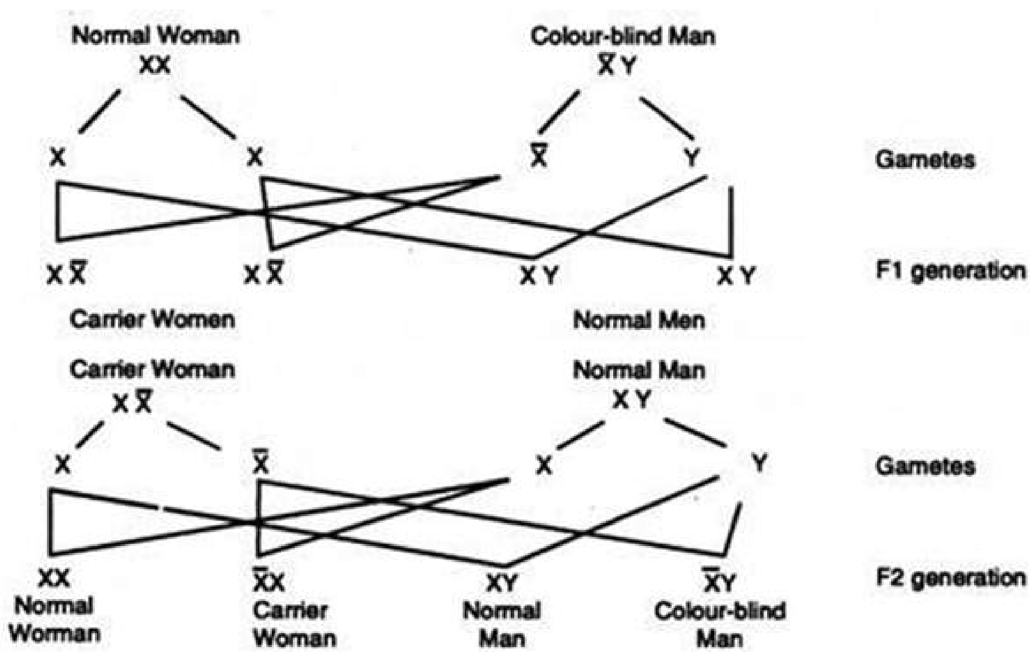
The situations can be explained in the following way: If a male with any of the above mentioned defects marry a normal homozygous female, all their sons will be normal. But their daughters will be the carriers (heterozygous) for the particular disease or defect, although they look almost normal.

This condition is produced due to a sex-linked recessive gene. When the daughters (carriers) get married to normal males, 50% of their sons will be perfectly normal while remaining 50% will get the defect. A few concrete instances will help in understanding the situation.

Red-Green colour blindness in man

There are several types of colour-blindness known, but the most common type is red-green blindness, which is an X-linked recessive trait. It is a defect in which person cannot differentiate red colour with green colour.

The genotype for normal vision may be symbolized by (XX), and colour blindness by ($\bar{X}$Y). $\bar{X}$ indicates the sex-linked recessive gene for colour blindness. If a colour blind man (Y) marries a normal woman (XX), in the F1 generation all male progeny (sons) will be normal (XY). The female progeny (daughters) though will show normal phenotype, but genetically they will be heterozygous ($\bar{X}$ X). Since these daughters bear the recessive gene of colour blindness, they are the carriers of the trait.



[$\bar{X}$ indicates the Sex-linked recessive gene for colour-blindness]

If such a carrier woman with normal vision (heterozygous for colour blindness) marries a normal man (XY), the following progeny may be expected in the F2 generation: Among the daughters, 50% are normal and 50% are carriers for the diseases; among sons, 50% are colour blind and 50% are with normal vision.

Activity

Draw a criss cross upto F1 generation and find out the result of that cross when a carrier woman marries a colour blind man. Similarly draw another chart to show the result of F1 generation when a colour blind woman marries a normal man.

Haemophilia in man

Haemophilia, a disorder in which blood fails to clot on exposure to air, cause prolonged bleeding even from the minor injuries. The trait was first observed in a European Royal Family; Queen Victoria (1819 – 1901) herself was a carrier of this recessive gene.

One of her sons, Leopold, Duke of Albany, died of haemophilia, when he was only 31 years. A woman may be carrying a gene for hemophilia and in the heterozygous condition ($\overline{X}X$), she does not exhibit any visible effect of the disease. But she is capable of transmitting the disease to 50% of her sons. Haemophilia is very rare in women, because, to have this disease the woman must be homozygous ($\overline{X}\ \overline{X}$) with the recessive genes. A haemophilic woman normally dies before adolescence due to severe bleeding.

This trait can also be explained in a criss-cross pattern of inheritance which is shown below. If a hemophilia carrier woman marries a normal man, 50% of her daughters will be normal; other 50% will be the carriers. Among the sons, 50% will be normal and other 50% will be hemophiliacs. Usual frequency of hemophilic male birth is about 1 in 10,000 while hemophilic female birth occurs once in 100,000,000 births.

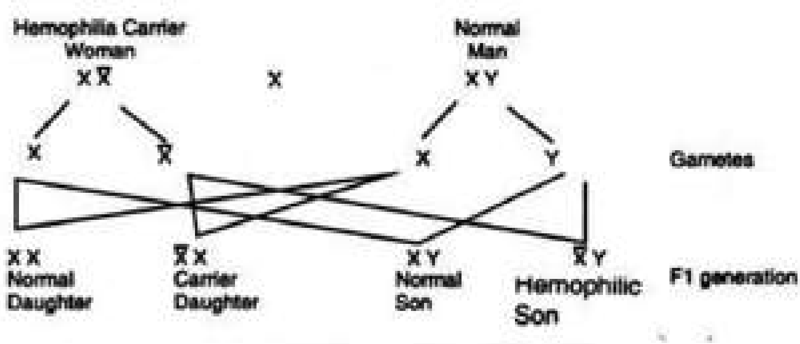


Fig: Inheritance of Haemophilia in man

16.4 Twins

Multiple birth is the birth of multiple offspring at a same time from the same pregnant woman. Example: twins, triplets, etc.

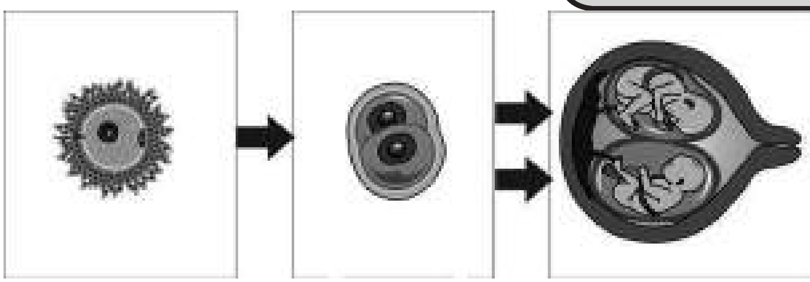
Twins are two offspring produced by the same pregnancy. Twins can be developed either from one zygote called identical (/monozygotic) twins or from two different zygotes called fraternal (/dizygotic which means two zygotes develop from two different eggs).

Formation of identical twins

A sperm fertilized an egg (ovum) to form a zygote. After fertilization, during the development of embryo, it splits into two and develop separately to form two babies with exactly the same genetic information. Identical twins look almost exactly alike and share the exact same genes. Most identical twins happen by chance.

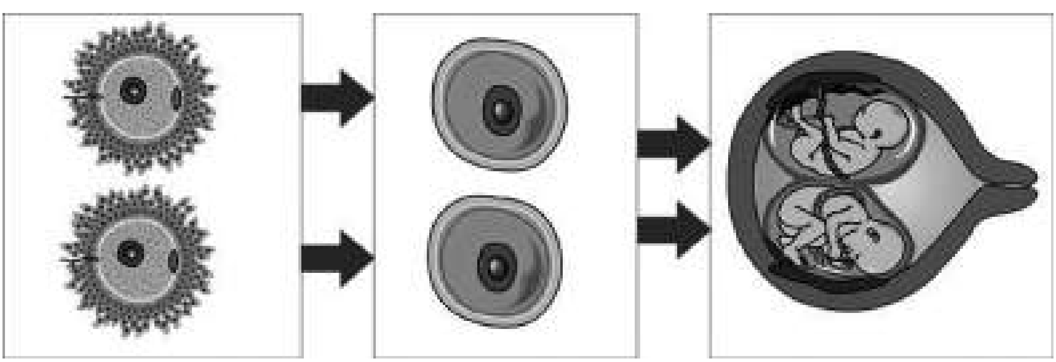
Do you know?

Medical opinion is still divided, but some scientists believe that there is a third type of twin which proposed that the egg splits in two, and each half is then fertilized by a different sperm. This theory is an attempt to explain why some fraternal twins look identical.



Formation of fraternal twins

Two different eggs (ova) developed at the same time are fertilized by two different sperm respectively to form two zygotes. Which develop to form fraternal twins. The fraternal twins are genetically unique which are no more alike than individual siblings born at different times.



- Twins may be male and male, male and female and female and female.
- Multiple births can be fraternal, identical, or a combination.

Multiple births are becoming more common nowadays. The possible reasons for it are advancing the average age of mothers and the associated rise in assisted reproductive techniques. Among multiple birth, about 90% account for birth is twins in human.

Factors that increase the probability of having twins

1) Advancing age of the mother – mother in their 30s and 40s have higher levels of the sex hormone oestrogen than younger women, which means that their ovaries are stimulated to produce more than one egg at a time.
2) Number of previous pregnancies – the greater the number of pregnancies a woman has already had, the higher her odds of conceiving twins.
3) Heredity – a woman is more likely to conceive fraternal twins if she is a fraternal twin, has already had fraternal twins, or has siblings who are fraternal twins.
4) Race – Black African women have the highest incidence of twins, while Asian women have the lowest.
5) Assisted reproductive techniques – many procedures rely on stimulating the ovaries with fertility drugs to produce eggs and, often, several eggs are released per ovulation.

Summary

1. The phenomenon of transmitting parental characteristics to the offspring is called Heredity.
2. The cross which are made in between two pairs of contrasting characteristics is known as dihybrid cross.
3. Chromosomes are thread like molecules that carry hereditary information for everything. The term 'chromosome' was given by Waldayer (1888AD).
4. The chromosomes which help to inherit hereditary characteristics from parents to offspring and have no role in sex determination and reproduction are called non-sex chromosomes (autosomes). There are 22pairs of autosomes in human.
5. The chromosome which play role in the reproduction and sex determination is called sex chromosome (allosome). There is 1 pair of sex chromosomes in human cell.
6. Sex linkage can be defined as the continuous association of genes located on the sex chromosome and are inherited together with the sex chromosome from generation to generation.
7. Sex linked inheritance (sex linkage) can also be defined as the inheritance of a trait (phenotype) that is determined by a gene located on one of the sex chromosomes.

8. The defect in which person cannot differentiate between two colours like red colour with green colour is called colour blindness.
9. Haemophilia is a disorder in which blood fails to clot on exposure to air, cause prolonged bleeding even from the minor injuries.
10. Twins are two offspring produced by the same pregnancy.
11. Twins are either identical (if they are developed from single zygote) or fraternal (if they are developed from two different zygotes).

Exercise

A. Tick (√) the best alternative from the followings:

1) Why did Mendel choose pea plant for his experiment?

   a) They have very short life cycle. b) They are bisexual plants.

   c) They are easy to cultivate. d) All of the above.

2) The phenotypic ratio of the dihybrid cross is:

   a) 9:3:2:1 b) 3:1

   c) 1:2:2:4:1:2:1:2:1 d) 9:3:3:1

3) The chromosome which play important role in sex determination is:

   a) autosome b) allosome

   c) none d) both

4) Which one is not a sex linked disease?

   a) Haemophilia b) Colour blindness

   c) Male pattern baldness d) Malaria

5) Fraternal twins are formed from:

   a) single zygote b) two different zygote

   c) two different mother d) none

B. Give short answers to the following questions.

1. Define dihybrid cross.
2. State law of independent assortment.
3. What is sex chromosome?
4. Differentiate between autosome and allosome.

5. When a woman with normal colour vision (but her father was colour blind) get married with a colour blind man then what are their chances of having colour blind children?
6. Differentiate between identical and fraternal twins.
7. Define sex linked inheritance.
8. What conclusion can you draw from the result obtained in the dihybrid cross of Mendel's experiment?
9. What is the genotypic and phenotypic ratio of Mendel's dihybrid cross?

C. Give long answers to the following questions.

1. If a man with normal vision marries a colour blind woman, show the phenotype and genotype of their children.
2. What is sex linked disease? How do they get transmitted to the offspring? Explain with suitable diagram.
3. A homozygous plant having round seed and inflated pods is first cross pollinate with a homozygous plant having wrinkled seed and constricted pods and then self pollinate. What result will be obtain in second generation? Show your result up to F2 generation with a proper chart.
4. Explain the process of formation of identical and non-identical twins in human.

Project work

Write down the phenotypic characteristics found among your classmates. Also collect the phenotypic characteristics of your friend's parents or grandparents. With the help of your data, conclude which one may be dominant and which one may be recessive. Also, explain the reason among your friends.

Glossary

Hybrids: the organisms produced when two genetically different organisms are crossed

Hybridization: the process of crossing two alternating characteristics or the process by which hybrids are formed

F1-generation(First filial generation): the first generation obtained by crossing two parents

F2-generation (Second filial generation): the generation obtained by crossing two parental stocks of F1 generation

Genotype: the genetic composition of an organism

Phenotype: the characteristic which appears externally

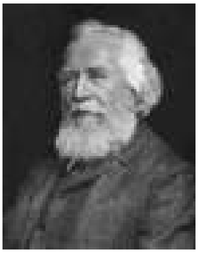
Dominant Characteristic: the appeared characteristics in F1 generation when two pure parents of contrasting characteristics are crossed

Recessive Characteristics: the unexpressed or suppressed characteristics in F1 generation when two pure parents of contrasting characteristics are crossed

Unit 17

Ecology

Haeckel identified many new species of living beings and gave names to thousands of them. He introduced new terms in biology such as phylogeny, anthropogeny, the kingdom Protista, phylum, stem cell, and ecology. He was also the first to regard psychology as a branch of physiology. Although he contributed much to the theory of evolution, he was best remembered for his extensive work on radiolarian.



Ernst Haeckel (1834-1919)

17.1 Introduction

Ecology can be defined as the branch of science which deals with the study of relationship between abiotic and biotic factors of the ecosystem. According to the Earnest Haeckel, "Ecology is the study of the reciprocal relationship between living organisms and their environment." Ecosystem is the structural and functional unit of ecology. It can also be defined as the inter-relationship between the abiotic and biotic factors of the environment.

Ecology can be divided into two types for study; **autecology** which means the study of individual animal or plant or their population throughout their life in the relation to the habitat to their environment and **synecology** which means the study of structure, nature and major fractions of communities (distribution of communitites).

In this unit we will learn about wetland ecosystem and the components of the wetland ecosystem and biogeochemical cycles (carbon cycle, oxygen cycle and nitrogen cycle).

17.2 Wetland ecosystem

Wetlands are one of the world's most important environmental assets. While covering only 6% of the Earth's surface, wetlands provide home for many important fishes and shellfish, including shrimp, blue crab, oysters, salmon, trout, and sea trout rely on, or are associated with, wetlands. Wetlands are also a habitat for migratory birds and waterfowl, including ducks, egrets, and geese. Wetlands also mitigate floods, protect coastal areas from storms, improve water quality, recharge groundwater aquifers, serve as sinks, sources, or transformers of materials, and produce food and goods for human use.

A **wetland** is a land area that is filled or soaked with water, either permanently or seasonally. Wetlands play a number of roles in the environment, mainly water purification, flood control, carbon sink and shoreline stability. The water found in wetlands can be freshwater, brackish, or saltwater.The main wetland types include swamps (marshes, bogs, and fens) and sub-types include mangrove, carr, pocosin, and varzea.

Do you know?

About 5% of the total surface area of Nepal is wetland area.

In Nepal, Koshitappu, Beeshazar lake and associated lakes, ghodaghodi lake area and Jagadishpur reservoir have been designated as wetland areas.

Note: *Constructed wetlands can be used to treat municipal and industrial wastewater as well as stormwater runoff. They may also play a role in water-sensitive urban design.*

Wetlands have many important functions that benefit ecosystem.

1. Wetlands provide habitat for a wide variety and number of wildlife and plants.
2. Wetlands filter, clean and store water.
3. Wetlands collect and hold flood waters.
4. Wetlands absorb wind and tidal forces.
5. Wetlands act like sponges by holding flood waters and keeping rivers at normal levels.
6. Wetlands filter and purify water as it flows through the wetland system. Plants found in wetlands help control land erosion.

Components of wetland ecosystem

Abiotic components of wetlands include the climate of the wetlands, water, light, rocks and minerals, soil, etc. Biotic components of wetlands include plants, animals, bacteria, and all other living organisms.

Examples of animals: Amphibians (especially found in wetlands), reptiles, birds, insects, and mammals.

Examples of plants: Mangrove, water lilies, cattails, sedges, tamarack, black spruce, cypress, and gum plants.

Abiotic components

Abiotic components means the non-living components of wetlands include the climate of the wetlands, water, light, rocks and minerals, soil, etc. These abiotic components provide the suitable environment for the proper growth, development and reproduction of the biotic components.

Water: Source of water in wetlands is mainly precipitation, surfaces water and ground water. The water found in wetlands can be freshwater, brackish, or saltwater.

Climate and Temperature: Temperatures vary greatly depending on the location of the wetland. Many of the world's wetlands are warm in summers and cold in winters, but temperatures are not extreme in temperate wetlands. However, wetlands found in the tropics, around the equator, are warm all year round.

Rainfall: The amount of rainfall a wetland receives varies widely according to its area. Wetlands in Wales, Scotland and Western Ireland typically receive about 1,500 mm (59 in) per year. In some places in Southeast Asia, where heavy rains occur, they can receive up to 10,000 mm (390 in). In the northern areas of North America, wetlands exist where as little as 180 mm (7.1 in) of rain falls each year.

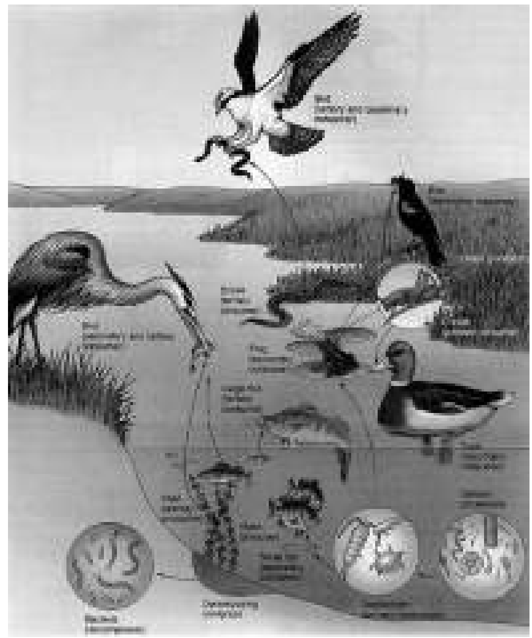


Fig: wetland ecosystem

Biotic components

The biota of a wetland system includes its vegetation (flora) and animals(fauna). The most important factor affecting the biota is the duration of flooding, fertility and salinity of soil.

Flora

The flora found in wetland can be studied into following types:

a) **Algae** are diverse water plants that can vary in size, color, and shape. Algae occur naturally in habitats such as inland lakes, inter-tidal zones, and damp soil and provide a dedicated food source for animals, fish, and invertebrates. There are three main groups of algae:

- Planktons are algae which are microscopic and free-floating. Planktons are primary producer in the ocean using photosynthesis to make food.
- Filamentous alga are long strands of algae cells that form floating mats.
- *Chara* and *Nitella* algae are upright algae that look like a submerged plant with roots.

b) Submerged wetland vegetation can grow in saline and fresh-water conditions. Some species have underwater flowers, while others have long stems to allow the flowers to reach the surface. Submerged species provide a food source for native fauna, habitat for invertebrates, and also possess filtration capabilities. Examples include sea grasses and eel grasses.

c) Floating water plants or floating vegetation is usually small, like arrow arum (*Peltandravirginica*). They donot need soil and are not rooted to the bottom of the pond.

d) Emergent wetland vegetation are rooted in soil but their shoots grow up out of the water. Example: cattails and rushes.

e) Scrub-shrub plants are shrubs or small trees which are rooted in the soil. They have a woody stem and grow up to 20 feet tall. Example: Buttonbush and alder.

Fauna

Many species of insects, frogs, turtles, fishes, shell fish, snakes, lizards, alligators and crocodiles live in wetlands, while others visit them each year to lay eggs. Mammals include numerous species of small mammals in addition to large herbivorous like beaver, swamp rabbit, Florida panther, etc.

Snapping turtles are one of the many kinds of turtles found in wetlands.

Frogs need both terrestrial and aquatic habitats to reproduce and feed. While tadpoles control algal populations, adult frogs feed on insects. Frog absorbs both nutrient and toxins from the surrounding environment due to their thin skin resulting in an above average extinction rate in unfavorable and polluted environmental condition.

Activity

Prepare a chart showing inter-relation between abiotic and biotic factors of your surroundings. And discuss it among the friends.

17.3 Biogeochemical cycle

The essential elements of the living matter are circulated by biogeochemical cycle in the atmosphere. The elements flow within the biogeochemical cycle in various forms from the abiotic (non-living) components to the biotic (living) components and then back into the atmosphere.The cycling of biological, geological and chemical elements through earth and its atmosphere is called biogeochemical cycle. These elements cycle through evaporation, absorption by plants and dispersion by wind. The cycling of the elements are essential to life processes. Therefore, repetition of this cycle is very important.

The biogeochemical cycle are classified into three types: Water cycle, Gaseous cycle and Sedimentary cycle (Mineral cycle)

Carbon Cycle

Carbon is the basic constituent of all living things. Carbon is a part of the ocean, air, and even rocks. The major source of carbon is:

1. Carbon dioxide in the atmosphere.
2. Carbonates of earth's crust derived from rocks.
3. Fossil fuels like peat, coal and petroleum products
4. In Oceans, carbon remains stored as carbonates in the form of limestone and marble rocks.

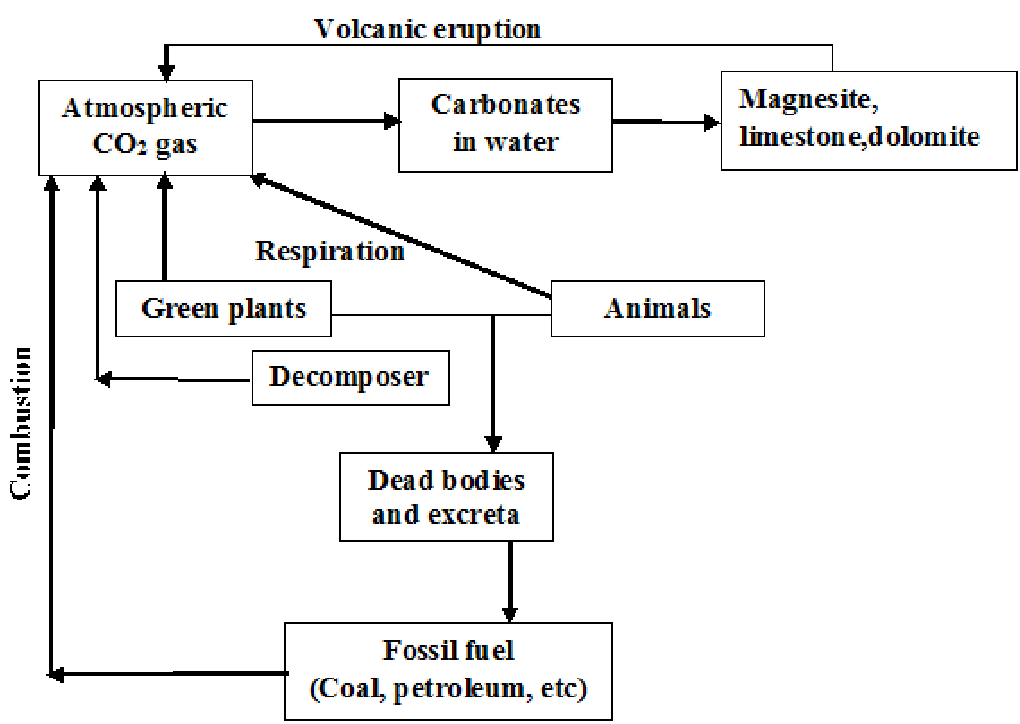


Fig: Carbon cycle

Carbon enters the atmosphere as carbon dioxide from respiration (breathing) and combustion (burning). Plants use **carbon dioxide** and sunlight to make their own food by the process of photosynthesis. Animals feed on the plants and however are exhaled as carbon dioxide through the process of breathing. When animals and plants die and get buried, they are eaten by decomposers and released to the atmosphere as carbon dioxide. However, if they are not decomposed, they turn into fossil fuels like coal and oil over millions of years. When humans burn fossil fuels, most of the carbon enters the atmosphere as carbon dioxide.

> Do you know?
> Carbondioxide and methane gas is the main source of carbon in the atmosphere. Both of these gases absorb and retain heat in the atmosphere and are partially responsible for the green house effect.

Carbon dioxide is released back in the atmosphere by the following process:

1. Carbon dioxide is released back to the environment by respiration of producers and consumers
2. Released by the decomposition of organic wastes and dead bodies by the action of bacteria and fungi.
3. Burning of wood and fossil fuels release carbon dioxide into the atmosphere
4. Volcanic eruptions and weathering of carbonate rocks by the action of acids

> ***Notes:*** *The carbon cycle is mainly balanced by photosynthesis process of green plants and respiration of plants and animals.*

Oxygen cycle

All living things need oxygen to create energy via the process called respiration. We breathe in oxygen and breathe out carbon dioxide. The atmosphere comprise only 0.35% of the Earth's total oxygen.

The oxygen cycle completes in the following steps:

1. The free oxygen is produced by the process of photolysis in the atmosphere. The sunlight energy breaks the oxygen bearing molecules to produce free oxygen.
2. In the biosphere, oxygen undergoes cycles of respiration and photosynthesis. Humans and animals breathe in oxygen and gives out. Plants utilize the carbondioxide in the process of photosynthesis to form carbohydrates and oxygen.
3. In the lithosphere, oxygen is fixed in minerals like silicates and oxides. The fixed oxygen is freed by chemical weathering.

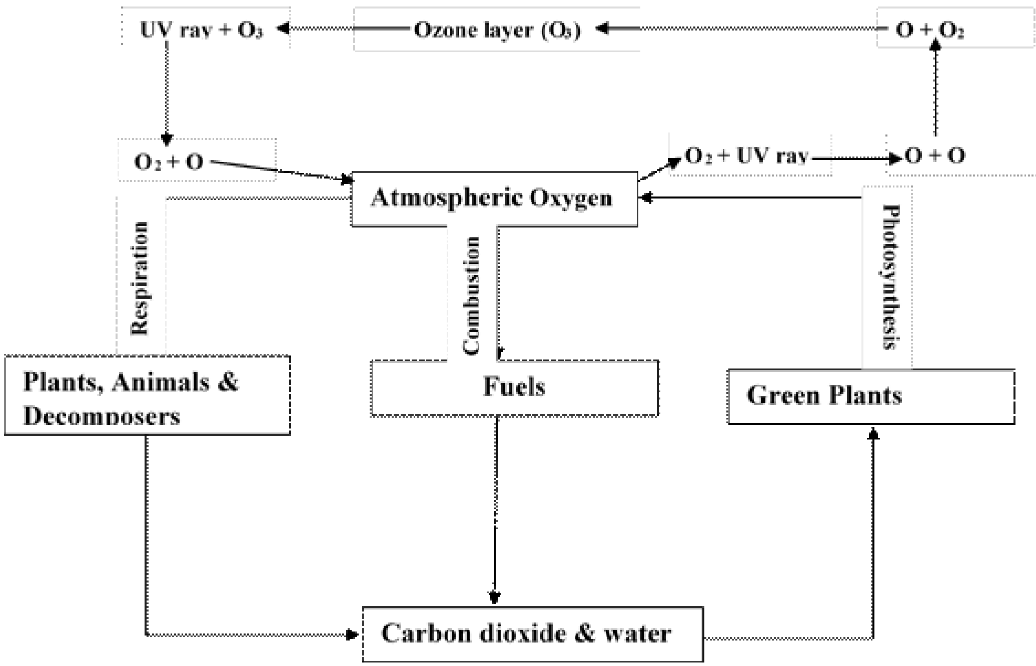


Fig: Oxygen cycle

In the atmosphere, when sunlight breaks the oxygen bearing molecules to produce free oxygen, this free oxygen then recombines with existing O_2 molecules to make O_3 or ozone. This cycle helps to protect the Earth from harmful ultra violet radiation.

In the biosphere, animals use oxygen to respire and give out in the air and water as carbon dioxide (CO_2). Carbon dioxide is then taken up by algae and terrestrial green plants and converted into carbohydrates and oxygen by the process of photosynthesis.

In the lithosphere, oxygen is found combined with the minerals such as silicates and oxides. A portion of oxygen is freed by chemical weathering. When oxygen bearing mineral is exposed to the extreme heat, chemical reaction occurs and produces free oxygen.

Nitrogen Cycle

Nitrogen is the most essential element for all living organisms. It is also an important component of protein, aminoacids, enzymes, nucleic acids, etc. In the earth's atmosphere, 78% of nitrogen is found. In nitrogen cycle, the atmospheric nitrogen is converted into various forms that can be utilized by the living organism. The various forms present in the environment are organic nitrogen, ammonium salts, nitrous oxide, nitrate, nitric oxide or nitrite.

The process of nitrogen cycle completes in five steps:

1. Nitrogen fixation
2. Nitrification
3. Nitrogen assimilation
4. Ammonification
5. Denitrification

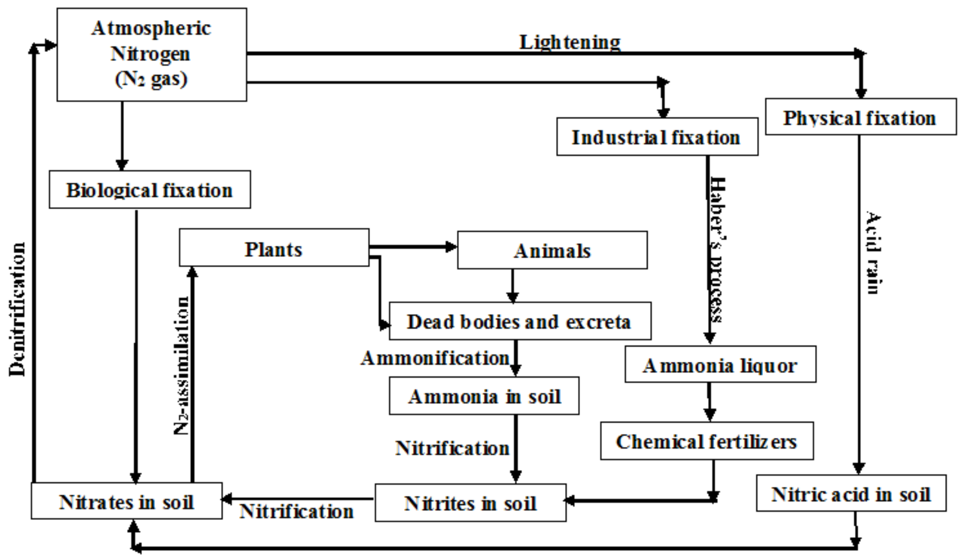


Fig: Nitrogen cycle

A. Nitrogen fixation

It is the process of conversion of atmospheric nitrogen into compounds of nitrogen. The nitrogen fixation occurs in three ways:

1. Biological fixation of nitrogen
2. Atmospheric fixation of nitrogen
3. Industrial fixation of nitrogen

1. Biological fixation of nitrogen

In biological fixation, some symbiotic bacteria, free living soil bacteria and blue green algae are able to change the atmospheric nitrogen into nitrates. This process is carried out by the bacteria present in the root nodules of leguminous plants and loose dug soil. The bacteria which convert atmosphere nitrogen into compounds of nitrogen are call nitrogen fixing bacteria. For example, clostridium Azotobacter, etc. Other organisms like symbiotic bacteria like Rhizobium and blue-green algae (cyanobacteria) and algae like Nostoc can also fix nitrogen.

2. Atmospheric fixation of nitrogen

This fixation occurs whenever lightening occurs in the air. During lightning, the atmospheric nitrogen reacts with oxygen in the air and forms nitric oxide and nitrogen dioxide. These oxides combine with water vapours of the air to form nitrous acids and nitric acids. These acids further combines with other salts and are converted into nitrate ions. This process is also called atmospheric fixation.

$$N_2 + O_2 \longrightarrow 2NO$$

$$2NO + O_2 \longrightarrow 2NO_2$$

$$NO_2 + O_2 \longrightarrow N_2O_5$$

$$N_2O_5 + H_2O \longrightarrow 2HNO_3$$

$$HNO_3 + CaCO_3 \longrightarrow Ca(NO_3)_2 + CO_2 + H_2O$$

3. Industrial fixation of nitrogen

In the industrial fixation of nitrogen, the nitrogen gas is made to combine with hydrogen gas to form ammonia gas artificially. This process is also called Haber's process. This ammonia can be used to prepare nitrates and ammonium salts which can be used as a fertilizer.

$$N_2 + 3H_2 \xrightarrow[\text{Fe/Mo}]{\text{(200-600)atm, 500°C}} 2NH_3$$

B. Nitrogen Assimilation

After the process of nitrification, plants absorb nitrates dissolved in soil for the synthesis of amino acids and proteins. When herbivores eat these plants, they get the nitrogen containing nutrients and when the carnivores eat these herbivores the nitrogen containing nutrients is transferred to them.

C. Ammonification

The simple inorganic compounds are utilized by the plants and then animals and converted into proteins and amino acids. When the remains of plants and animals and their waste products are decomposed by the microorganism in the soil like *Bacillus, Proteus, Clostridium, Pseudomonas, Streptomyces* etc. ammonia is liberated. This process is known as ammonification. This ammonia is dissolved in the soil in the form of ammonium ions.

D. Nitrification

Nitrification is the process of oxidation of ammonia to ammonium salts and then into nitrates or nitrites by the action of soil bacteria. *Nitrosomonas, Nitrobacter, Nitrosococcus*, etc perform oxidation of ammonia to their nitrates and are called as nitrifying bacteria.

$$2NH_3 + 3O_2 \xrightarrow{\text{Nitro somonas}} 2HNO_2 + 2H_2O$$

$$2\ HNO_2 + O_2 \xrightarrow{\text{Nitrobacter}} 2HNO_3$$

E. Denitrification

Denitrification is the process of reduction of nitrates into free nitrogen by the action of some bacteria in the soil which are called denitrifying bacteria. Some examples of denitrifying bacteria are *Agrobacterium, Micrococcus denitrificans, Pseudomonas, Bacillus*, etc. denitrification completes the nitrogen cycle.

Activity

Work in a group of 5 or 6 students. Discuss among your friends about the advantages of different biogeochemical cycles. What would happen if biogeochemical cycles are absent in the environment? Come up with the conclusion and present your report in front of the class.

Summary

1. A wetland is a land area that is filled or soaked with water, either permanently or seasonally. The water found in wetlands can be freshwater, brackish, or saltwater.
2. Abiotic components of wetlands include the climate of the wetlands, water, light, rocks and minerals, soil, etc.
3. Biotic components of wetlands include plants, animals, bacteria, and all other living organisms.
4. The circulation of essential elements in the various forms from the abiotic (non living) components to the biotic (living) components and then back into the atmosphere is called biogeochemical cycle.

5. The water found in wetlands can be freshwater, brackish, or saltwater.
6. Submerged and floating plants are the main flora of wetlands.
7. Many species of insects, frogs, turtles, fishes, shell fish, snakes, lizards, alligators and crocodiles live in wetlands.
8. Algae provide a dedicated food source for animals, fish, and invertebrates.
9. Carbon enters the atmosphere as carbon dioxide from respiration (breathing) and combustion (burning).
10. In the atmosphere, when sunlight breaks the oxygen bearing molecules to produce free oxygen, this free oxygen then recombines with existing O_2 molecules to make O_3 or ozone.
11. Nitrogen is an important component of protein, aminoacids, enzymes, nucleic acids, etc.
12. Nitrogen fixation is the process of conversion of atmospheric nitrogen into compounds of nitrogen in soil.
13. Assimilation is the formation of aminoacids and proteins from the inorganic nitrogen compounds obtained from nitrogen fixation and utilized by plants and animals.
14. Ammonification is the degradation of complex organic nitrogen compounds present in the dead remains of plants and animals and their waste materials into ammonia.
15. Nitrification is the process of oxidation of ammonia to ammonium salts and then into nitrates or nitrites by the action of nitrifying bacteria.
16. Denitrification is the process of conversion of nitrates into free nitrogen by the action of some soil bacteria

Exercise

A. Tick (√) the best alternative from the followings:

1. The flow of essential elements in a cycle is called:

   a. Elements cycle   b. Geochemical cycle

   c. Biogeochemical cycle   d. Chemical cycle

2. The atmosphere comprises only .... of the Earth's total oxygen.

   a. 0.35%   b. 0.33%   c. 0.32%   d. 0.34%

3. Rhizobuimis the example of:

   a. Nitrifying bacteria   b. Ammonifying bacteria

   c. Nitrogen fixing bacteria   d. Denitrifying bacteria

4. Ammonia is converted into nitrates by the action of:

   a. Nitrifying bacteria   b. Ammonifying bacteria

   c. Denitrifying bacteria   d. Nitrogen bacteria

5. The percentage of nitrogen in atmosphere is:

   a. 75.1%   b. 78.2%   c. 78%   d. 7.3%

B. Give short answers to the following questions.

1. What do you understand by wetlands?
2. List the biotic and abiotic components of wetland ecosystem.
3. Write the importance of wetlands.
4. Write the importance of algae in wetland ecosystem,
5. Define biogeochemical cycle.
6. What are the types of biogeochemical cycle found in atmosphere?
7. What is the main source of carbon?
8. Why is carbon cycle important?

9. Write the main steps of oxygen cycle.
10. Write any two main source of oxygen.
11. What is the difference between nitrification and dinitrification?
12. Name any two ammonifying bacteria.
13. Give the name of two nitrogen fixing bacteria.
14. What are denitrifying bacteria?

C. Give long answers to the following questions.

1. Describe the components of wetland ecosystem.
2. Describe carbon cycle in nature.
3. Write about the oxygen cycle.
4. Give a short account of nitrogen cycle.

Project work

1. Draw a neat and labelled diagram of different biogeochemical cycles in a chart paper, and paste it on the wall of your classroom.
2. Visit a wetland area near to your place. Observe the wetland and list some of the flora and fauna found in that wetland area. Submit your report to your teacher.

Glossary

Brackish:	slightly salty, saline
Swamps:	an area of low lying and uncultivated ground where water
Mangrove:	a shrub or small tree that grows in coastal saline or brackish water
Carr:	North European wetland
Pocosin:	a type of wetland with deep, acidic, sandy, peat soils
Decomposer:	an organism that decomposes organic matter in the soil
Weathering:	change the appearance by long exposure to air

Unit 18

Applied Biology

Sciopione was an internist, pathologist and pediatrician. He is known as the inventor of the mercury sphygmomanometer for the measurement of blood pressure.

Scipione Riva Rocci (1863–1937)

18.1 Introduction

With the knowledge of biology, we can apply that in the development of new techniques to make the study of biology easier which is called applied biology. Applied biology is really the study and development of tools and methods specific to biological research needs.

In this unit we will learn about blood group and its grouping, hypertension, antibiotic that we use in our practical life and organ transplantation.

18.2 Organ transplantation

Organ transplantation is a process in which a damaged or lost organ or part of recipient is replaced with functional ones from the body of donor. The tissue or organ which gets transplanted is called graft. There are many examples of tissue or organ transplantation on human beings such as skin grafting, heart transplantation, kidney transplantation etc.

There is less chance of success of organ transplantation due to production of antibodies by patient's immune system against the transplanted organ. Organs that have been successfully transplanted include the heart, kidneys, liver, lungs, pancreas, intestine, and thymus. Some organs, like the brain, cannot be transplanted.

Types of Transplantation

Depending on genetic relationship between donor and receiver, transplantation is of four types:

A) **Autograft:** Autografts are the transplant of tissue from one area to another area in same person. Skin grafting for plastic surgery is the common example.

There is high success rate in this type of transplantation.

B) **Isograft:** Isograft is the transplantation of tissue or organ between genetically similar individuals of a species. Example: Kidney transplantation.

C) **Homograft (Allograft):** Allograft is the transplantation of tissue or organ between genetically different persons of the same species. Most human tissue and organ transplants are allografts. Example: Skin transplantation and blood transplantation from others.

D) **Heterograft (xenograft):** Xenograft is the transplantation of tissue or organ between animals of different species. These types of transplantation are specially carried for research and experiment. Example: porcine heart valve transplant, which is quite common and successful.

Advantage of organ transplantation

1. The damaged tissue or organ can be replaced by another healthy tissue or organ in order to perform normal function.
2. Organ transplantation gives a new life to a person dying of organ failure.

Disadvantage of organ transplantation

1. As during the transplantation, due to disturbance of immune system, the recipient body may reject the transplanted organ and may die.
2. Organ transplantation is very expensive and cannot be affordable to all persons.

18.3 Blood pressure

Blood pressure is the pressure exerted by blood on the wall of blood vessels. It is of two types: systolic blood pressure and diastolic blood pressure. Systolic pressure is the pressure which the blood exerts on the wall of arteries at the end of contraction of ventricles and diastolic pressure is the pressure which exerts on the wall of the arteries when the ventricles are totally relaxed. The arterial blood pressure is always written in systolic pressure by diastolic pressure in mm of Hg form.

The device which is used to measure arterial blood pressure is

> ***Do you know?***
> The blood pressure of a normal adult is 120/80 mm Hg.

> ***Do you know?***
> In earlier time blood pressure was measured with arterial puncture. The sphygmograph was the first noninvasive instrument used to measure blood pressure which wasinvented in 1860 by Étienne Jules Marey.

sphygmomanometer.

High blood pressure (hypertension) can quietly damage our body for years before any symptoms develop. If it is uncontrolled, it may lead to a disability, a poor quality of life or even a fatal heart attack.

Causes of High blood pressure

1. Family history of high blood pressure (Genetics)
2. Too much smoking and drinking habit
3. Being overweight or obese
4. Lack of physical exercise
5. Consumption of too much salt in the diet
6. Stress
7. Older age
8. Chronic kidney disease
9. Adrenal and thyroid disorders
10. Inadequate sleep

Effects of high blood pressure

1. High blood pressure forces our heart to work harder than necessary in order to pump blood to the rest of your body causing our heart muscle to weaken and work less efficiently and damage our heart.
2. Uncontrolled high blood pressure can lead to stroke by damaging and weakening our brain's blood vessels, causing them to narrow, rupture or leak.
3. High blood pressure can injure both the blood vessels in and leading to your kidneys, causing several types of kidney disease (nephropathy).
4. High blood pressure can damage the vessels supplying blood to our retina, causing retinopathy.
5. High blood pressure damages the lining of your blood vessels and causes our arteries to harden and narrow (atherosclerosis), limiting blood flow.
6. High blood pressure can also cause Trouble sleeping, Memory loss, personality changes, trouble concentrating, irritability or progressive loss of consciousness, Chest pain, Heart attack, Complications in pregnant women, etc.

Control measures for controlling High blood pressure

1. **Weight loss:** Blood pressure often increases as weight increases. Being overweight can disrupt breathing while we sleep, which raises our blood pressure. Therefore, weight loss is most effective for controlling blood pressure.

2. **Regular exercise:** Regular physical activity like walking, jogging, cycling, swimming or dancing at least 30 minutes a day can lower our blood pressure by 4 to 9 millimeters of mercury (mm Hg).

3. **Healthy diet:** Eating a diet that is rich in whole grains, fruits, vegetables and low-fat dairy products can lower your blood pressure by up to 14 mm Hg.

4. **Reduce sodium consumption in diet:** Even a small reduction in the sodium in your diet can reduce blood pressure by 2 to 8 mm Hg.

5. **Limit the amount of alcohol while drinking:** Drinking more than moderate amounts of alcohol can actually raise blood pressure by several points.

6. **Quit smoking:** Each cigarette we smoke increases our blood pressure for many minutes after we finish. Quitting smoking helps your blood pressure return to normal.

7. **Cut back on caffeine:** Caffeine can raise blood pressure by as much as 10 mm Hg in people who rarely consume it, but there is little to no strong effect on blood pressure in habitual coffee drinkers.

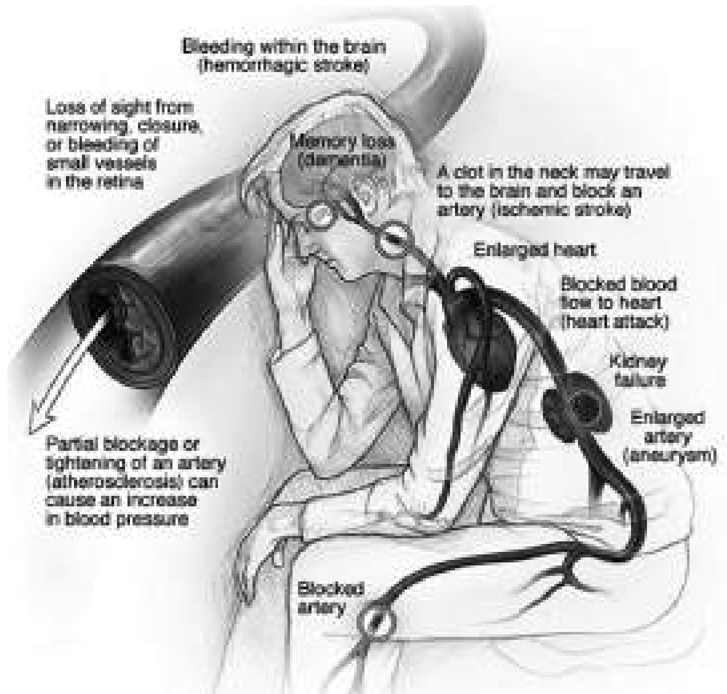
8. **Reduce your stress:** Chronic stress is an important contributor to high blood pressure. Occasional stress also can contribute to high blood pressure if we react to stress by eating unhealthy food, drinking alcohol or smoking.

9. **Monitoring blood pressure at home and visiting doctor regularly:** Home monitoring can also help to keep control on blood pressure

10. **Get support:** Family and Friends can encourage us to take care of ourselves, drive to the doctor's office or encourage joining exercise program that can keep the blood pressure normal.

Activity

Measure the blood pressure of all the students present in your classroom. Make a table and fill all the data on that table. Compare the level of blood pressure among boys and girls. Find out how many numbers of students have high blood pressure or low blood pressure or normal blood pressure. Prepare a report on it.



18.4 High Cholesterol

Cholesterol is a type of fat (lipid) in our blood made by the liver and distributed throughout the body. Our cells need healthy levels of cholesterol to function. It allows our bodies to make vitamin D and hormones, and makes up bile acids. Our body makes all we need, but we also get less than 25 percent of our body's cholesterol from the foods we eat, especially animal fats.

If we have too much cholesterol, it starts to build up in our arteries. This is called hardening of the arteries (artherosclerosis). This can narrow the arteries and make it harder for blood to flow through them, leading to dangerous blood clots and inflammation that can cause heart attacks and strokes.

There are different types of cholesterol.

1. Low-density lipoproteins (LDL) transport cholesterol around to where it's needed. If there's too much cholesterol, it may be deposited into the arteries and raise the risk of heart disease, heart attack and stroke. LDL is commonly referred to as "bad cholesterol."
2. High-density lipoproteins (HDL) take the extra cholesterol from your tissues and cells and return it to your liver for repurposing. That's why HDL is

called "good cholesterol." It has a lower risk to our heart.

Causes of high cholesterol level

Many things can affect cholesterol level including:

1. Eating too much saturated fat, Trans fat and cholesterol can raise our cholesterol level.
2. Being overweight may lower HDL ("good") cholesterol.
3. Being inactive or not doing exercise may lower HDL ("good") cholesterol.
4. Cholesterol starts to rise after the age 20.
5. If family members have or had high cholesterol, you may also have it.

Effects of High Cholesterol

1. High amount of LDL will make the arteries develop plaque, a hardened mixture of cholesterol, fat, and other elements which can block blood flow with a great risk of having a heart attack, stroke and Coronary Heart Disease (CHD)
2. High cholesterol can create a bile imbalance, leading to gallstones.
3. High cholesterol can interfere with blood flow to the legs, causing numbness, pain, or even gangrene.
4. Blocked arteries due to high cholesterol can block blood flow to our brain and the brain cell begins to die.
5. Pain and discomfort in the chest.

Control measures

Heart-healthy lifestyle changes can help lower risk for everyone. They include:

1. Eating a heart-healthy diet that is rich in fruits, vegetables, whole grains, fish, and low-fat or nonfat dairy foods.
2. Being active on most, if not all, days of the week.
3. Losing weight if you are overweight, and staying at a healthy weight
4. Not smoking.

18.5 Uric Acid

Uric acid is produced from the natural breakdown of our body's cells and from the foods we eat.

Most of the uric acid is filtered out by the kidneys and passes out of the body in urine. A small amount passes out of the body in stool. But if too much uric acid is being produced or if the kidneys are not able to remove it from the blood normally, the level of uric acid in the blood increases.

High levels of uric acid in the blood can cause uric acid to form crystals within joints or tissues of bones. This causes a painful condition called gout. In gout, crystals of uric acid are deposited in the joints, where they cause a type of arthritis called gouty arthritis. These same crystals can also deposit in the kidneys, where they can cause kidney stones.

Normal Uric acid levels are 2.4-6.0 mg/dL (female) and 3.4-7.0 mg/dL (male).

Effects of high level Uric acid (Hyperuricemia)

Gout causes inflammatory arthritis, swelling, redness, heat, stiffness in the joints and pain. A low fever of 99°F to 100°F is common. Chronic gout may reduce the function of the kidney and result in the formation of kidney stones. The deposits of hard lumps of uric acid in the tissues and the joints destroy the joints.

There are four main causes of the high levels of uric acid that lead to gout:

1. A diet rich in chemicals called purines, because purines are broken down by the body into uric acid. Foods that contain high levels of purines include anchovies; nuts; and organ foods such as liver, kidney and sweetbreads.
2. High production of uric acid by the body. This can happen for unknown reasons. It can also occur in certain inherited genetic metabolic disorders, leukaemia and during chemotherapy for cancer.
3. The kidneys do not excrete enough uric acid. This can be caused by kidney disease, starvation and alcohol use, especially binge drinking. This also can occur in people taking medications called diuretics (such as hydrochlorothiazide or furosemide).
4. Obesity or sudden weight gain can cause high uric acid levels because the body's tissues break down more purines.

Control measures of gout

- Eating a healthy and balanced diet and avoiding foods that contain purines.
- Drinking plenty of fluids every day (8-10 glasses of water) to help flush the uric acid out of the body system.
- Eating fresh fruits and vegetables.
- Increasing dairy products in diet.

- Doing regular exercise.
- Reducing the amount of alcohol consumption, especially beer.
- loosing if overweight or obese

18.6 Antibiotic

Antibiotics are medicines produced by microorganisms that are used in the treatment and prevention of bacterial infections.

The word antibiotic means 'against life.' This antimicrobial drug also called antibacterial may either kill or inhibit the growth of these tiny infection-causing organisms, known as pathogens.

Antibiotics are not effective against the viral diseases such as common cold or influenza.

> ***Do you know?***
>
> In 1928, Alexander Fleming identified penicillin, the first chemical compound with antibiotic properties. Fleming was working on a culture of disease-causing bacteria when he noticed the spores of a little green mold (*Penicillium chrysogenum*), in one of his culture plates. He observed that the presence of the mold killed or prevented the growth of the bacteria

Types of Antibiotics

I. On the basis of action towards pathogens

a) **Bactericidal:** This is the type of antibiotic that is used to kill bacteria. Example: Metronidazole, Cotrimoxazole, Fluoroquinolones, etc.

b) **Bacteriostatic:** This is the type of antibiotic that is used to slow the growth rate of bacteria but not to kill them. Example: Tertacyclines, Trimethoprim, Macrolides, etc.

II. On the basis of effect on pathogens

a) **Broad spectrum Antibiotic:** This antibiotic can be effective against several different types of harmful bacteria and may be used to treat a variety of infectious diseases. A broad-spectrum antibiotic acts against both Gram-positive and Gram-negative bacteria.

Example: Tertacyclines, Chloramphenicol.

b) **Narrow spectrum Antibiotic:** Narrow-spectrum antibiotics are effective against a selected group of bacterial types. It can be used only if the causative organism is identified.

Example: Penicillin, Streptomycin, Azithromycin, Clarithromycin, Clindamycin, Erythromycin, Vancomycin, etc

Advantages of Antibiotic

1. Antibiotics are used to treat or prevent bacterial infections and sometimes protozoan infections.
2. It is used as preservatives in canned or packed foods.
3. It may also be used as supplement in animal feed.

Disadvantages of Antibiotic

As a side-effect, antibiotics can also kill the body's normal and beneficial microbes along with the pathogen. The destruction of the body's normal bacterial flora provides an opportunity for drug-resistant microorganisms to grow vigorously and can lead to a secondary infection.

Other disadvantages of antibiotics include ineffectiveness against viruses, drug interactions and also frequent misuse, overuse and potential side effects lead to development of antibiotic-resistant strains.

18.7 Blood Group

A blood group is a classification of blood based on the presence or absence of antigens present in RBCs and antibodies in blood plasma. These antigens are designated by the English alphabets A and B and antibodies as a and b. Depending upon the presence and absence of these antigen and antibodies, four blood groups have been designated as A, B, AB and O. This system of blood grouping is called as ABO blood Grouping system.

> *Do you know?*
>
> The blood groups A, B and O were discovered by Karl Landsteiner in 1900 for which he got Nobel Prize in 1931.

In ABO blood grouping system,

1. If a person's blood have antigen A in the RBC and antibody b in the plasma, the person have blood group A.
2. If a person's blood have antigen B in the RBC and antibody a in the plasma, the person have blood group B.
3. If a person's blood have antigen A and B in the RBC and no antibody in the plasma, the person have blood group AB.
4. If a person's blood have no antigen in the RBC but both antibodies and b in

the plasma, the person have blood group O.

Blood Group	Antigen in RBC	Antibody in Plasma
A	A	b (anti-B)
B	B	a (anti-A)
AB	A AND B	none
O	NONE	a and b (anti A and B)

Rh-Factor

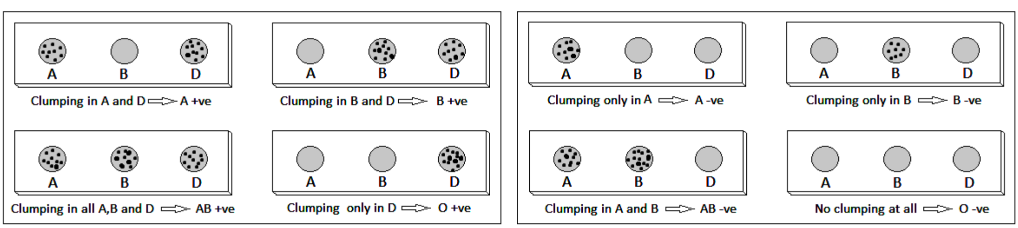
Rh-factor (Rhesus factor) is a protein (antigen) that is present on the surface of RBC. If Rh factor is present in a person's blood, then he is said to be Rh positive (Rh^+) and if Rh factor is absent, then he is said to be Rh negative (Rh^-).

Do you know?

Rh-Factor was first isolated from RBC's of Rhesus Monkey by Karl Landsteiner and Wiener in 1940.

Activity

Determination of Blood group



Clumping of blood with 'Anti D serum' indicates the presence of Rh-Factor in blood

No clumping of blood with 'Anti D serum' indicates the absence of Rh-Factor in blood

First of all, a finger is sterilized with a sterile cotton ball and pricked with a sterile blood lancet. Then, three drops of blood are taken in different places of a glass slides for A, B and Rh-Factor (D). Then, anti A serum is added in the drop of blood labeled A, anti B serum added in the drop of blood labeled B and Anti D serum is added in the drop of blood labeled D, respectively. Then the mixture of blood and antiserum is mixed properly with the help of a bamboo stick and allow resting for few minutes to see agglutination (clumpings).

If we see clumping in blood labeled A after putting anti serum-A, then its blood group is A. Similarly, If we see clumping in blood labeled B after putting anti serum-B, then its blood group is B, and if we see clumping in both A and B, then the blood group is AB.

Similarly, agglutination of sample D indicates the Rh^+.

Summary

1. Organ transplantation is a process in which a damaged or lost organ or part of recipient is replaced with functional one from the body of donor.
2. Organs that have been successfully transplanted include the heart, kidneys, liver, lungs, pancreas, intestine, and thymus. Some organs, like the brain, cannot be transplanted.
3. Autografts are the transplant of tissue from one area to another area in same person.
4. Isograft is the transplantation of tissue or organ between genetically similar individuals of a species.
5. Allograft is the transplantation of tissue or organ between genetically different persons of the same species
6. Xenograft is the transplantation of tissue or organ between animals of different species.
7. Blood pressure is the pressure exerted by blood on the wall of blood vessels.
8. Systolic pressure is the pressure which the blood exerts on the wall of arteries at the end of contraction of ventricles and diastolic pressure is the pressure which exerts on the wall of the arteries when the ventricles are totally relaxed.
9. Uncontrolled high blood pressure can lead to stroke by damaging and weakening our brain's blood vessels, causing them to narrow, rupture or leak.
10. Cholesterol is a type of fat (lipid) in our blood made by the liver and distributed throughout the body.
11. High amount of LDL will make the arteries develop plaque, a hardened mixture of cholesterol, fat, and other elements which can block blood flow with a great risk of having a heart attack, stroke and Coronary Heart Disease (CHD).
12. Gout causes inflammatory arthritis, swelling, redness, heat, stiffness in the joints and pain.
13. Antibiotics are medicines produced by microorganisms that are used in the treatment and prevention of bacterial infections.
14. The destruction of the body's normal bacterial flora provides an opportunity for drug-resistant microorganisms to grow vigorously and can lead to a secondary infection.
15. A blood group is a classification of blood based on the presence and absence of antigens present in RBCs and antibodies present in blood plasma.
16. In ABO blood Grouping system, there are four blood groups which have been designated as A, B, AB and O with Rh factor either positive or negative.

Exercise

A. Tick (√) the best alternative from the followings:

1. The transplantation of tissue or organ between animals of different species is called:

   a. Autograft　　b. Isograft

   c. Allograft　　d. Xenograft

2. The normal value of blood pressure in adult person is ....

   a. 130/80 mm Hg　　b. 120/90 mm Hg

   c. 120/80 mm Hg　　d. 80/120 mm Hg

3. The pressure which the blood exerts on the wall of arteries at the end of contraction of ventricles is ....

   a. Systolic pressure　　b. Diastolic pressure

   c. Blood pressure　　d. High blood pressure

4. Artherosclerosis is :

   a. Hardening of veins　　b. Hardening of arteries

   c. Asthama　　d. Arthritis

5. Hyperuricemia is:

   a. Urine infection　　b. High level Uric acid

   c. Increased frequency of urination　　d. Low level of Uric acid

6. Gout is caused due to:

   a. High cholesterol　　b. High uric acid

   c. Low uric acid　　d. Low cholesterol

7. Antibiotics are used in the treatment of:

   a. Virus　　b.Fungi

   c. Bacteria　　d. Protozoa

8. If a person's blood have antigen A in the RBC and antibody b in the plasma, the person have:

   a. Blood group A　　b. Blood group AB

   c. Blood group B　　d. Blood group O

9. Rh factor is present in the surface of:

   a. RBCs b. WBCs

   c. Blood d. Plasma

B. Give short answers to the following questions.

1. What is organ transplantation?
2. Name the organ that cannot be transplanted till now.
3. Give one example each of the following:

   a) Autograft b) Allograft

   c) Heterograft d) Isograft
4. What do you mean by allograft and isograft?
5. Write the differences between Xenograft and Allograft.
6. What is blood pressure?
7. Name the instruments used to measure blood pressure.
8. Write the differences between systolic and diastolic pressure.
9. Write the differences between Low-density lipoproteins (LDL) and High-density lipoproteins (HDL).
10. What is gout?
11. Define antibiotics.
12. List the advantages and disadvantage of antibiotics.
13. What is ABO system of blood grouping?
14. What is Rh factor?

C. Give the long answers to the following questions.

1. What are the different types of organ transplantation? Describe the procedure of organ transplantation.
2. Write the cause, effect and control measures of high blood pressure.
3. Write the cause, effect and control measures of high cholesterol level in blood.
4. Write the cause, effect and control measures of high uric acid in blood.
5. What are the types of antibiotics? Explain.

6. Describe the process of determining blood group in detail.

Project work

1. Ask the blood pressure, cholesterol level and uric acid level of your family members, relatives and neighbours. Separate them into different column, high, low or normal. If they say high or low then also note down the complications they are facing due to their problems. Prepare a report and present it to the class.
2. Collect the data of the blood group of your family members, relatives and if possible then of your neighbours too. Prepare a report and submit it to your teacher. It was found in other studies that most of the family members have similar type of blood groups. Is your result similar with it? If yes then discuss, why do they have similar type of blood groups?
3. Visit a pathology lab nearby your home. Collect some reports about different patients and also discuss what types of complications are facing by patients visiting there. Prepare a report and present it to the class.

Glossary

Grafting:	process of transplantation of the tissue
Systolic blood pressure:	the pressure exerted by the blood on the walls of blood vessels when ventrical contracts
Diastolic blood pressure:	the pressure exerted by the blood on the walls of blood vessels when ventricals relax
Retinopathy:	disease related with retina
Low-density lipoproteins:	cholesterol having low density
High-density lipoproteins:	cholesterol having high density
Gout:	disorder when uric acid level is high
Pathogen:	disease causing organism

Unit 19

The Earth

Geological diversion of Nepal

Nepal is a land locked mountainous country which is located at the central part of the Himalayan belt (2500km long). About 83% of this country is mountainous. Nepal is an underdeveloped country but very rich in different types of natural resources like water, minerals, forest, medical herbs, culture, etc. Exploration and proper use of these natural resources is the crucial thing which can bring up the sustainable development of this country. Various types of metallic, non-metallic and fuel mineral are found in Nepal and we will learn about those resources in this chapter.

Continuous geodynamic process in the Himalayan region occurs which results in the faulting, folding magmatic activities in Nepal. Nepal can be divided into five morpho-geotectonic zones from south to north on the basis of availability of mineral resources and they are:

a) **The southernmost Terai plain:** It is rich in gravel, sand, ground water, petroleum and natural gas.

b) **Sub-himalaya (Churia/Siwalik foot hills):** It is rich in construction materials, radioactive materials, petroleum, natural gas and coal.

c) **Lesser Himalaya (The mahabharat range including mid hills):** It is rich in metallic minerals like iron, copper, lead, zinc, cobalt, nickel, tin, tungsten, molybdenum, gold, uranium rare metals etc., industrial minerals like magnesite, phosphorite, limestone, dolomite, talc, clay, kaoline etc., gemstones like tourmaline, aquamarine/ beryl, garnet, kyanite, etc, fuel minerals like coal, lignite, methane gas, petroleum and natural gas, radioactive mineralsetc.

d) **Higher Himalaya:** It is rich in precious and semiprecious stones, marble and metallic minerals like lead, zinc, uranium, gold etc.

e) **Tibetan Tethys zone:** It is rich in limestone, gypsum, brine water (salt) and natural gas.

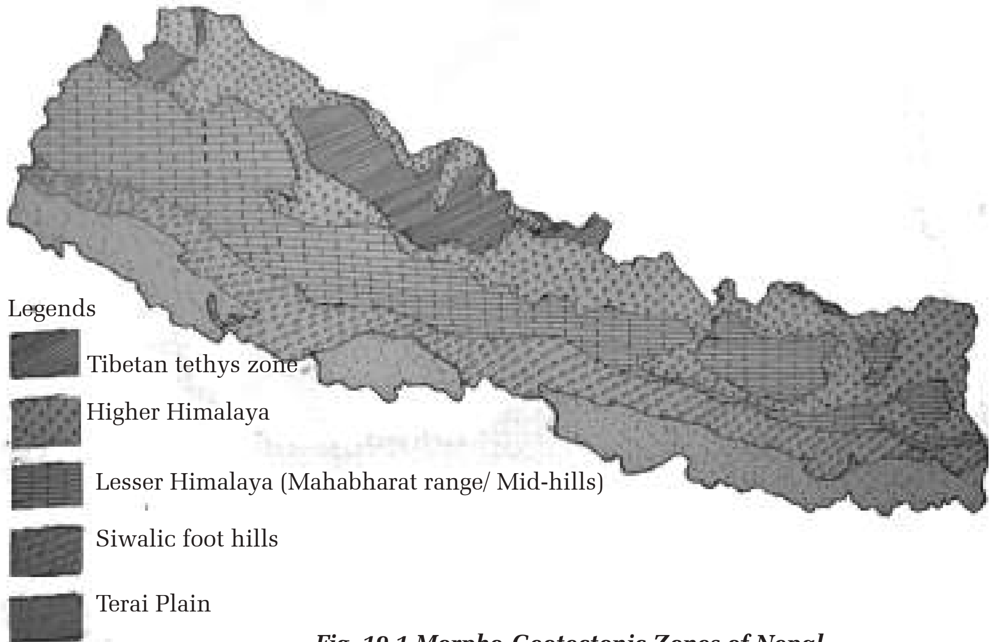


Fig. 19.1 Morpho-Geotectonic Zones of Nepal

Rocks, Minerals and Ores found in Nepal

Mineral deposits, Mines and Their Status

The projects carried out mainly by DMG, UNDP concluded with the sixty six minerals (metallic, non-metallic, gemstone and fuel deposits) in Nepal. Some of the important are described below.

A) Metallic Minerals

Metals are being used in our daily life. Metals are extracted from their respective ores. But of the least reactive metals are available in their native form. Some of the important metallic ore minerals and their status in Nepal are discussed below.

Iron (Fe) It is used extensively in various developmental works. From a small fort to a framework of huge buildings, all are made from iron. The main iron ores magnetite, haematite, limonite are reported from different part of Nepal including Phulchoki (Lalitpur), Thoshe (Ramechhap), Labdi Khola (Tanahun), Jirbang (Chitwan), Dhuwakot (Parbat), Purchaundi/ Lamunigad (Bitadi), Dahabagar, Kachali, and Ekghar/ Khanigaon (Bajhang). Iron prospects and old workings are also known from different parts of Baitadi, Bajhang, Jajarkot, Rolpa, Surkhet, Myagdi, Baglung, Parbat, Chitwan, Ramechhap, Okhaldhunga, Taplejung etc.

Copper (Cu): It is mainly used in electrical industries to make copper wires, crafts, making alloys, different sculptures etc. Nepal used to mine copper a long time ago but nowadays none of the copper mine is running. The copper ore found in Nepal are mainly chalcopyrite and few malachite, azurite, covellite, cuprite, bornite, chalcocite, etc. These copper ore are found in Gyazi (Gorkha), Okharbot (Myagdi), Wapsa (Solukhumbu), Kalitar (Makwanpur), Dhusa (Dhadhing), Bamangaon (Dadeldhura), Khandeshori/ Marma (Darchula), Kurule (Udayapur), Bhut Khola (Tanahun), Pandav Khani (Baglung), Baise Khani (Myagdi), Minamkot (Syangja), Chhirling Khola (Bhojpur), Jantare Khani (Okhaldhunga).

Zinc (Zn) and Lead (Pb): The ore of zinc and lead are found in Ganesh Himal area (Rasuwa), Phakuwa (Sankhuwasabha), Labang- Khairang, Baraghare and Damar (Makwanpur), Pangum (Solukhumbu), Salimar valley (Mugu/ Humla), Phulchoki (Lalitpur), Sisha Khani and Kandebas (Baglung), Dhuwakot (Parbat), Bhaludanda (Dhadhing, Khola Khani (Taplejung) etc.

The use of zinc are galvanizing iron, dry battery, pigments, soldering, dyeing, glue making etc., manufacturing various alloys e.g. brass, bronze etc. whereas the use of lead are making lead sheets, pipes, alloys, pigments, dyeing and printing process, insecticide, medicine etc.

Gold (Au): It is very precious metal and widely used in making coins, ornaments, jewelry, electroplating, metal coating and many other purposes. In Nepal, alluvial/ placer gold are obtained from the river gravel/ sediments deposited by the rivers like Mahakali, Chamliya, Jamari Gad, Seti, Karnali, Bheri, Rapti, Lungri Khola & Phagum Khola (Rolpa), Kaligandaki, Myagdi Khola, Modi, Madi, Marshyangdi, Trishuli, Budhigandaki, and Sunkoshi. There is no gold mine yet recorded in Nepal although local people extract small scale of gold from above mentioned rivers.

Silver (Ag): In Nepal, small amount of silver in Ganesh Himal (Rasuwa), Barghare (Makwanpur), and polymetal sulphide of Bering Khola (Ilam), Netadarling (Arghakhanchi) and Samarbhamar (Gulmi). Silver ore has not yet been reported in Nepal, however, it is present in association with zinc-lead ores and gold. It is mainly used to make jewelry, coins, handicrafts, utensils dental fillings, etc.

B) Nonmetallic Minerals

Nepal is rich in different types of rocks or non-metallic minerals including magnesite, phosphorite, talc, limestone, dolomite, quartz, mica, clay, silica sand, gemstones, decorative and dimension stones, construction materials etc. Some important rocks are discussed below.

Limestone ($CaCO_3$): Limestone deposits are found in Lesser Himalayan region only. It is majorly used in manufacturing cement and lime. The need of cement is increasing day by day as rapid increase in developmental activities. Limestone deposits are found in Kathmandu, Khotang, Udayapur, Syangja, Palpa, Arghakhanchi, Dang, Pyuthan, Sallyan, Rolpa, Bajhang, Baitadi and Darchula districts.

Dolomite [$CaMg(CO_3)_2$]: Generally dolomite and limestone occur together. Dolomite occur mainly in Dhankuta, Khotang, Udayapur, Sindhuli, Dolakha, Kavre, Kathmandu, Makwanpur, Dhadhing, Syangja, Palpa, Baglung, Gulmi, Arghakhanchi, Dang, Pyuthan, Sallyan, Rolpa, Rukum, Jajarkot, Surkhet, Dailekh, Jumla, Achham, Doti, Bajhang, Bajura, Baitadi and Darchula districts in the Lesser Himalayan and in some parts of Higher Himalayan region. It is used in paints, soap, detergent and agriculture purpose.

Phosphorite [$Ca_3(PO_4)_2$]: It is one of the main raw materials to manufacture chemical fertilizers like fused magnesium phosphate, triple super phosphate etc. This rock is found Junkuna, Dhaubisaune (Baitadi), tarugad, juilgad, goichan-kandechaur (in Bajhang), sewar khola (Dang) and Mari Khola (Pyuthan)

Magnesite ($MgCO_3$): It is found in Kharidhunga (Dolakha), Kampughat (Udayapur), palpa, Baitadi, etc. Kharidhunga a type of magnesite deposit found in Dolakha is one of the biggest and best quality magnesite in south Asia. It is used to manufacture high temperature refractory bricks that can be used for lining in the furnaces, steel industries and crucibles in chemical industries.

Talc [$H_2Mg_3(SiO_3)_4$] or $Mg_3Si_4O_{10}(OH)_2$: It is found in Lalitpur, Dolakha, Sindhupalchok, Dhadhing, Chitwan, Tanahun, Kaski, Syangja, Surkhet, Bajhang, Bajura, Baitadi and Darchula districts. It is used in cosmetics, ashtray, ornaments (highly pure talc), in paper, plastic, ceramic, paints, soaps, plasters, crayons, leather(medium grade talc) and as dry fire extinguisher powder (low pure talc).

Ceramic clay/ Red clay: Ceramic clay or kaolin are found in Daman and Kharka/ Tistung (Makwanpur), Panchmane (Kathmandu), Dalchhap and few other places. They are mainly used in ceramics.

Similarly red clay is found in Panchkhal (Kavre), Lamosure (Hetaunda), Trijuga/ Beltar (Udayapur), Chidika (Arghakhanchi), Guttu (Surkhet) which is used in cement factory.

Graphite: It can be found from the places of Lesser Himalayan regions like Ilam, Dhankuta, Sankhusabha, Nuwakot, Sindhupalchok, Dadeldhura etc. It is mainly used in lead pencils, foundry facings, crucibles, paints, lubricants etc.

Salt: Brine water that occurs in Narsing Khola (Mustang), Chhiding Khola and Chharkabhot (Dolpa) are tapped and dried for common salt production.

C) Gemstones

Semiprecious gem stones found in Nepal are tourmaline, aquamarine/ beryl, garnet, kyanite, amethyst, citrine, smoky quartz (quartz crystals); and precious stones like ruby, sapphire etc. are known only from few districts.

(a) **Tourmaline:** Five distinct types of tourmaline are recognized from Nepal. It is available in pink, bright green, light orange sometimes with repeated colour banding, olive green with amber coloured core. It is found in Langtang valley (Rasuwas), Naje (Manang), Daha (Jajarkot), etc. Tourmaline crystals are cut and polished for making gem.

(b) **Beryl/ Aquamarine:** Beryl from Taplejung (Ikabu, Lodantar) area is expensive. It is found in Taplejung, Phakuwa (sankhuwasabha), Manang, Lekhpatan and Tikachaur in Jajarkot; Jagat, Panchmane, Baguwa, Tarkeghyang, Nibuwagaon in Sindhupalchok are the other known places for Beryl.

(c) **Garnets:** Garnets are metamorphic rock found in higher Himalayan region. Deep red or red coloured almandine, hessonite and pyrope garnet are found mainly in Sankhuwasabha and Taplejung districts.

(e) **Rubies and Sapphire:** Ruby and Sapphire are found in Chumar, Ruyil (Dhadhing) and Lari/ Ganesh Himal (Rasuwa). They are available in highly tectonized and intensely folded metamorphic rocks.

D) Decorative and Dimension Stones

Marble: It is a coloured decorative stones which can be obtained from Godavari (Lalitpur), Anekot (Kavre),Chhatre Deurali (Dhadhing), Bhainse and Sukaura (Makwanpur).

Granites: It can be obtained from Makwanpur (Palung and Ipa), Sindhuli, Udaypur, Dadeldhura in the Lesser Himalaya.

Quartzite is a hard metamorphic rock which consists of mainly quartz. It is found in Taplejung, Ilam, Dhankuta, Ramechhap, Sindhupalchok, Makwanpur, Dhadhing, Tanahun, Kaski, Syangja, Parbat, Baglung, Achham, Doti, Bajhang, Bajura, Dadeldhura, Baitadi, Darchula, etc.

Slate: It is the common material for roofing. It is found in Dhankuta, Sindhupalchok, Ramechhap, Nuwakot, Dhadhing, Tanahun, Baglung, Syangja, Palpa,Parbat, Jajarkot, Achham, Doti, Dadeldhura, Baitadi, Bajhang, Bajura, etc.

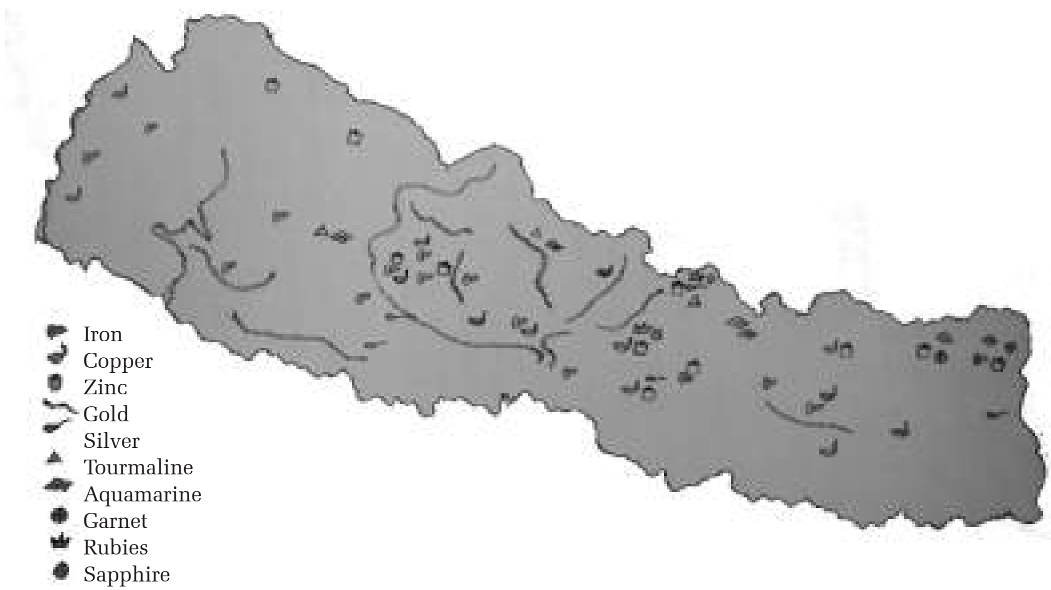


Fig. 19.2 Distribution of some metallic minerals

Fossils

The preserved remainings, or traces of remainings, of ancient organisms are called fossils. The preserved remaining or their traces are preserved under the sedimentary rocks. A fossil can preserve an entire organism or just part of the body like bones, shells, feathers, and leaves can all become fossils. Fossils can be very large or very small.

Fossils found in Nepal

Nepal has a great biodiversity. Due to the different natural phenomenon different types of organisms got killed in ancient time which are preserved under the rocks and form fossils. Some of the fossils of ancient animals which were present in the territory of Nepal are preserved in the Natural History Museum (NHM).

The museum was established in 1975. Since then it has provided with more than 50,000 specimens of Nepal's modern and prehistoric plants and animals. Some of the important fossils preserved there are fossilized skull of the Archidiskodon (elephant found in Silwalik Hills of Nepal), molar teeth of Sivapithecus, a hominoid. Both of these fossils, skull and teeth are believed to be around 3 million years and 8-10 million years old respectively. Similarly in the Himalayan region, Mustang, cephalod fossils like saligram (ammonites) are also found.

A) **Saligrams:** Saligrams, found in the Kali Gandaki river, are black fossils. These fossils are also called ammonites. The term ammonite is derived from a term ammonis cornua which means the horn of Ammon; an Egyptian god who wore curled horns on his head. According to geologists, saligrams are 140 to 165 million years old and found in two colors, black and white. They are the remains of extinct aquatic animals, mollusks of the cephalopod family. They look like black balls and due to its specific colour the name of that river was kept Kali Gandaki. Some ammonites have a spiral shape and a few have a tube like structure. In life, they resemble a shell-less octopus, squid, or cuttlefish.Similarly, spiti shale with its Saligram Member, composed of shales with concretions, which contain cephalopods, bivalves, and other fossils, as illustrated here. Large gemstones (e.g., ruby) is also be considered as saligram.

Fig: 19.3 Saligram

B) **Giraffa punjabiensis (Extinct Giraffe):** The genus Giraffa were evolved almost 25 million years ago. The extinct Giraffa punjabiensis inhabited in large parts of Eurasia including the ancient land of Nepal. The molar teeth of this extinct Giraffe were found on the banks of Surai khola in Nepal.

C) **Archidiskidon planifrons (Extinct Elephant):** Many fossilized body parts of the elephant, Archidiskidon were recorded from different parts of Nepal. There is a fossilized skull of the Archidiskidon (3 million years) found on the banks of Rato Khola in the collection of Natural History Museum.

D) **Ramapithecus sivalensis (Extinct Primate):** The age of these fossil remains is estimated about 12.5 million to 8.5 million years old whichwas recorded from the Siwalik Hills of Nepal and other adjoining areas. The molar teeth of the ancient hominoid, Ramapithecus (around 9.0 – 9.5 million years old), was found in Tinau Khola on way to Tansen from Butawal. During that time it was claimed that the jaw was more like a human jaw. Ramapithecus was understood to have direct ancestral roots to modern humans.

E) **Hexaprotodon sivalensis (Extinct Hippopotamus):** The estimated age of this fossil is 5.35 to 0.11 million years old. It was recorded from Indonesia, India and Nepal. In Nepal the fossils of Hexaprotodon sivalensis have been discovered at Gidniya Village and Rato Khola in Nepal.

The skull and lower jaw of this extinct hippopotamus is one of the appealing items in the collection of Natural History Museum.

Fig 19.4: Stamps on different fossils found in Nepal

Age of the rocks

Do you know?

Post Authority of Nepal issued a set of 4 stamps (all of Rs.50) of some fossils recorded from the territory of the Nepal on December 31, 2013.

Age of the rocks can be determined with the help of different methods; relative dating method and absolute or numeric dating method. In relative dating method, age of given samples are determined as 'older' or 'younger' than a standard one, but in absolute or numeric dating method, age is obtained in number of years. Here we will learn only about the absolute or numeric dating method which depends upon the duration of "half-life" of common radioactive elements.

Radioactive elements use to email invisible particles/rays from their nucleus continuously which causes the change in their atomic number, atomic mass, atomic volume and properties that ultimately transforms one type of element to another stable element. This is called decaying of radioactive elements. **The fifty percentage of total radioactive atoms present in a sample takes a fixed duration to be changed into stable ones which is called the half-life**.

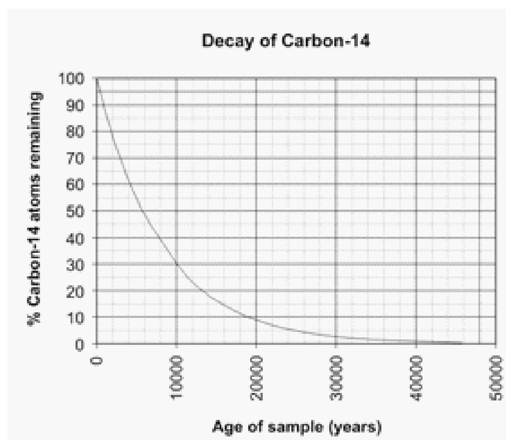
In absolute dating method, a radioactive isotope with known half-life are used to establish the absolute age of the fossil or rock. Some of the commonly used techniques are: Radiocarbon dating (use carbon-14 isotope), Uranium-lead dating (use lead-206 and lead 207 isotopes), Uranium-thorium dating (use to date fossil bones), Potassium-argon dating, argon-argon dating, etc.

Radio-Carbon dating method

Carbon element has 3 isotopes: ${}_6C^{12}$, ${}_6C^{13}$ and ${}_6C^{14}$. Among them ${}_6C^{12}$ and ${}_6C^{13}$ are stable in nature and they occupy almost 99.99% of total carbon existence in nature, but ${}_6C^{14}$ is highly radioactive and continuously undergo decaying into stable ${}_7N^{14}$ element by releasing β-particles with the half-life of 5730±40 years.

${}_6C^{14}$ isotopes are formed in atmosphere (between 9 to 15Km) when cosmic radiation hits nitrogen atoms. The result of this cosmic reaction will be $6C^{14}$ and a proton.

$${}_7N^{14} + {}_0n^1 \longrightarrow {}_6C^{14} + {}_1p^1,$$



${}_6C^{14}$ carbon atoms also have same chemical property as that of ${}_6C^{12}$ or ${}_6C^{13}$. Therefore, they also make CO_2 and get into the plants body through photosynthesis and in animals body through consumption of food. They also normally get out from the body of organisms through the process of excretion. Entering and exiting of ${}_6C^{14}$ to and from the body of organisms continue till they survive at more or less uniform rate. But, after dying of the organisms, the atoms of ${}_6C^{14}$ left in their body start to decay into ${}_7N^{14}$ continuously at uniform rate of their half-life. By counting exact number of ${}_6C^{14}$ left in the sample very precisely, one can state almost exact "death-date" of the organism.

Since the duration of half-life of ${}_6C^{14}$ is only 5730±40 years, radiocarbon dating method is appropriate to find the age of the rock or any carbon containing sample which is not older than 62,000 years only.

The figure alongside shows the order of the counting of ${}_6C^{14}$ left in a sample along with the duration of death-date)or the age of the samples.

Uranium-lead radioactive dating

This is one of the oldest method in the history of radioactive dating of rocks and minerals that were formed about 1 million years to over 4.5 billion years ago. An isotope of Uranium (${}_{92}U^{238}$) is decayed into lead (${}_{82}Pb^{206}$) and other isotope ${}_{92}U^{235}$

into lead $_{82}Pb^{207}$ with their half-lives 4.47 billions year and 710 million years respectively.

This method is usually performed on the mineral zircon which has very high affinity with uranium and thorium but it strongly rejects lead. If any trace of lead is detected in the zircon then one can easily assume that the lead is absolutely radiogenic (product of radioactive decay). By finding the current ratio of uranium-lead atoms in the minerals, its age can be calculated easily.

> Do you know?
> Uranium-lead radioactive dating can also be used on baddeleyite, monazite minerals.

Uranium-thorium dating method

In this method, an isotope of uranium ($_{92}U^{234}$) is decayed into thorium ($_{90}Th^{230}$) with half-life of about 80,000 years. In between uranium and thorium, uranium is water soluble whereas thorium is not. Due to this hydrophilic and hydrophobic natures of uranium and thorium isotopes respectively, they are used selectively into the ocean floor sediments. The age of the rocks is calculated by finding their ratios (uranium-234/thorium-230).

Summary

1. Nepal is an underdeveloped country but very rich in different types of natural resources like water, minerals, forest, medical herbs, culture, etc.
2. Nepal can be divided into five morpho-geotectonic zones from south to north on the basis of availability of mineral resources and they are: Terai plain, Sub-himalaya (Churia/Siwalik foot hills), Lesser Himalaya (The mahabharat range including mid hills), Higher Himalaya and Tibetan Tethys zone.
3. Some of the important mineral found in Nepal are: Metallic Minerals (Iron, Copper, Zinc, Lead, Gold, Silver, etc), nonmetallic Minerals (Limestone, Dolomite, Phosphorite, Magnesite, Talc, Ceramic clay/ Red clay, Graphite, Salt etc), Gemstones (Tourmaline, Beryl/ Aquamarine, Garnets, Rubies and Sapphire etc) and Decorative and Dimension Stones(Marble, Granites, Quartzite, Slate etc)
4. Age of the rocks can be determined with the help of different methods; relative dating method and absolute or numeric dating method.
5. Some of the commonly used techniques for the determination of age of rocks are: Radiocarbon dating (use carbon-14 isotope), Uranium-lead dating (use lead-206 and lead 207 isotopes), Uranium-thorium dating (use to date fossil bones), Potassium-argon dating, argon-argon dating, etc
6. Radiocarbon dating is simply called carbon dating.

7. In Uranium –lead radiometric dating, Uranium-235 or Uranium-238 isotopes are used to date the absolute age of the rocks.
8. In Uranium-thorium radiometric dating, Uranium-234 and Thorium-230 are used with a half-life of about 80,000 years.
9. Half-life is the fixed duration in which 50% of a radioactive element is changed into normal and stable element.

Exercise

A. Tick (√) the best alternative from the followings:

1. Which one of the following is not found in Nepal?

   a) Iron b) Garnet

   c) Limestone d) None

2. Which one of the following is not a radioactive element?

   a) Radium b) Plutonium

   c) Uranium d) Potassium

3. Which one of the following is cephalod fossil found in Nepal?

   a) Molar teeth of this extinct Giraffe

   b) Molar teeth of the ancient hominoid

   c) Skull of this extinct hippopotamus

   d) Saligram

B. Give short answers to the following questions.

1. Name any three minerals found in Nepal.
2. What do you mean by age of the rocks?
3. Write any three methods of determination of age of rocks.
4. Write any two differences between radio-carbon dating and uranium-lead dating techniques.
5. Explain Uranium-thorium dating technique in short.

C. Give long answers to the following questions.

1. Compare radiocarbon dating with Uranium-lead dating and Uranium-thorium dating individually.

Glossary

Geodynamic process: the dynamic process of Earth (movement of tectonic plates)

Half-life of the element: the time period to decay half of the sample element

Hydrophilic: soluble in water

Hydrophobic: insoluble in water

Unit 20

The Universe


Michel de Nostredame (1503–1566)

Michel de Nostredame (1503–1566) was a 16th century physician who wrote under the Latin pen name Nostradamus. He is most well-known for his book, The Prophecies, which was first published in 1566. The book contains hundreds of predictions about the future, written in the form of short, cryptic four line poems known as quatrains.

During his lifetime Nostradamus also published a popular almanac, and in the last year of his life he introduced an innovation in the almanac format to include a prediction for each day of the year.

History of space exploration

Introduction

Humans have dreamed about spaceflight since antiquity. The Chinese used rockets for ceremonial and military purposes centuries ago, but only in the latter half of the 20th century were rockets developed that were powerful enough to overcome the force of gravity to reach orbital velocities that could open space to human exploration.

As often happens in science, the earliest practical work on rocket engines designed for spaceflight occurred simultaneously during the early 20th century in three countries by three key scentists: in Russia, by Konstantin Tsiolkovski; in the United States, by Robert Goddard; and in Germany, by Hermann Oberth.

In the 1930s and 1940s Nazi Germany saw the possibilities of using long-distance rockets as weapons. Late in World War II, London was attacked by 200-mile-range V-2 missiles, which arched 60 miles high over the English Channel at more than 3,500 miles per hour.

After World War II, the United States and the Soviet Union created their own missile programs. On October 4, 1957, the Soviets launched the first artificial satellite, Sputnik 1, into space. Four years later on April 12, 1961, Russian Lt. Yuri Gagarin became the first human to orbit Earth in Vostok 1. His flight lasted 108 minutes, and Gagarin reached an altitude of 327 kilometers (about 202 miles).

The first U.S. satellite, Explorer 1, went into orbit on January 31, 1958. In 1961 Alan Shepard became the first American to fly into space. On February 20, 1962, John Glenn's historic flight made him the first American to orbit Earth.

"Landing a man on the moon and returning him safely back to Earth within a decade" was a national goal set by President John F. Kennedy in 1961. On July 20, 1969, Astronaut Neil Armstrong took "a giant step for mankind" as he stepped on the moon. Six Apollo missions were carried out to explore the moon between 1969 and 1972.

During 1960s unmanned spacecraft photographed and probed the moon before astronauts ever landed. By the early 1970s orbiting communications and navigation satellites were in everyday use, and the Mariner spacecraft was orbiting and mapping the surface of Mars. By the end of the decade, the Voyager spacecraft had sent back detailed images of Jupiter and Saturn, their rings, and their moons.

Skylab, America's first space station was a human-spaceflight highlight of the 1970s, was the Apollo Soyuz Test Project, the world's first internationally crewed (American and Russian) space mission.

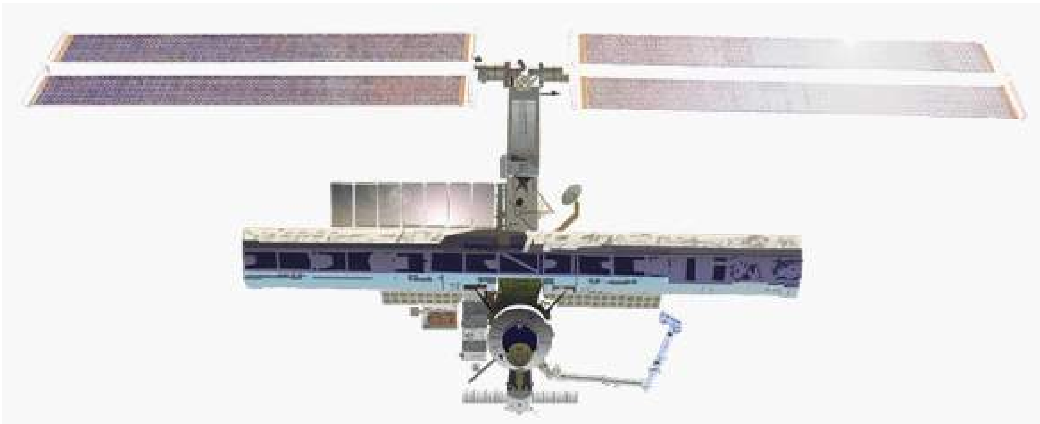
In 1980s satellite communications expanded to carry television programs, and people were able to pick up the satellite signals on their home dish antennas. Satellites discovered an ozone hole over Antarctica, pinpointed forest fires, and sent photographs of the nuclear power-plant disaster at Chernobyl in 1986. Astronomical satellites found new stars and gave us a new view of the center of our galaxy.

Space Shuttle

In April 1981 launch of the space shuttle Columbia ushered in a period of reliance on the reusable shuttle for most civilian and military space missions. Twenty-four successful shuttle launches fulfilled many scientific and military requirements until January 1986, when the shuttle Challenger exploded after the launch, killing its seven crews.

In the past, satellites were used to provide information on enemy troop formations and movements, early warning of enemy missile attacks, and precise navigation in the featureless desert terrain. The advantages of satellites allowed the coalition forces to quickly bring the war to a conclusion, saving many lives.

Space systems will continue to become more and more integral to homeland defense, weather surveillance, communication, navigation, imaging, and remote sensing for chemicals, fires and other disasters.



International Space Station

The International Space Station is a research laboratory in low Earth orbit. With many different partners contributing to its design and construction, this high-flying laboratory has become a symbol of cooperation in space exploration, with former competitors now working together.

And while the space shuttle will likely continue to carry out important space missions, particularly supporting the International Space Station, the Columbia disaster in 2003 signaled the need to step up the development of its replacement. Future space launch systems will be designed to reduce costs and improve dependability, safety, and reliability. The NASA (governmental space agency of the United States) and other nations have their own launch systems, and there is strong competition in the commercial launch market to develop the next generation of launch systems.

Astronomical Telescope

The virtual image forming optical instrument which is used to view distant objects clearly is called an astronomical telescope.

Types of Telescopes

There are two basic types of telescopes. They are refractors telescope and reflectors telescope.

Refractor telescopes

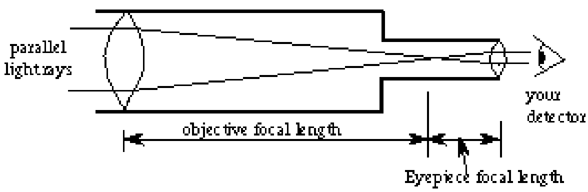
A refractor telescope uses a glass lens as its objective. The glass lens is kept in front of the telescope and light is refracted as it passes through the lens.

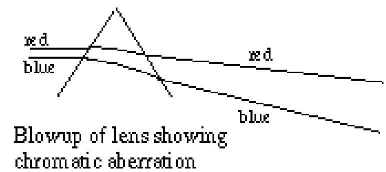
Advantages of refractor telescope

a. Refractor telescopes are rugged.

b. The glass surface inside the tube is sealed from the atmosphere so it rarely needs cleaning.

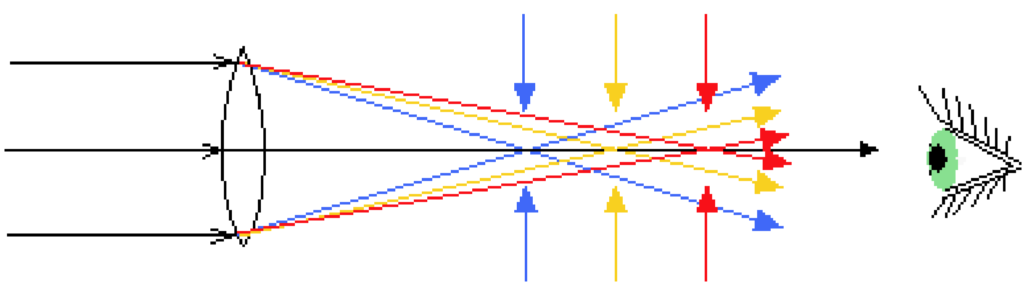
c. Since the tube is closed off from the outside, air currents and effects due to changing temperatures are eliminated.





Disadvantages of refractors telescope

a. All refractors suffer from an effect called chromatic aberration that produces a rainbow of colors around the image. Because of the wave nature of light, the longer wavelength light is bent less than the shorter wavelength light as it passes through the lens.



b. Ultraviolet light does not pass through the lens at all.

c. The intensity of light passesd through it decreases as the thickness of the lens increases.

d. It is difficult to make a glass lens with no imperfections inside the lens and with a perfect curvature on both sides of the lens.

e. The objective lens can be supported only at the ends. The glass lens will sag under its own weight.

Reflector telescopes

A reflector telescope consists of a mirror as its objective. The mirror is close to the rear of the telescope and light is reflected as it strikes the mirror. All celestial objects (including those in our solar system) are so far away that all of the light rays coming from them reach the Earth as parallel rays. Since the light rays are parallel to each other, the reflector telescope's mirror has a parabolic shape. The parabolic-shaped mirror focuses the parallel light rays to a single point. All modern research telescopes and large amateur ones are of the reflector type.

Advantages of reflector telescope

a. Since all wavelengths will reflect off the mirror in the same way, the reflector telescopes do not suffer from chromatic aberration.

b. The objective mirror is supported along the back side so they can be made very big too.

c. Reflector telescopes are cheaper in comparison to refractors telescope of the same size.

Disadvantages of the reflector telescope

a. A reflector telescope's tube is open to the outside and the optics need frequent cleaning.

b. The secondary mirror is used to redirect the light into a more convenient viewing spot. The secondary mirror and its supports can produce diffraction effects.

Uses of astronomical telescope

a. To gather as much light as possible: This is done by using a large aperture lens or mirror. The amount of light gathered in it depends on the area of the lens used.

b. To resolve fine detail: This is done by using a large aperture lens or mirror. The larger the aperture the finer will be the detail.

c. To magnify the image of a distant object: This is done by using a lens or mirror with a long focal length. The actual magnification of a telescope can be worked out from the formula:

$$\text{Magnification} = \frac{\text{Focal length of the objective (F)}}{\text{Focal length of eyepiece lens (f)}}$$

Numerical Illustration

Calculate the magnification of a telescope with an objective of focal length 1200 mm using two eyepieces of focla lenght 25 mm and 10 mm.

Magnification for the first case = 1200/25 =48x

Magnification for the second case = 1200/10 =120x

20.3 Constellation

The small group of stars seen in the sky especially at night with fixed shape is called constellation. Constellation available in the sky is of different shapes and sizes. There are 88 constellations known to us till this time. We can see them in the northern and southern sky at night. The ancient people of many civilizations named the constellations of northern and southern hemispheres. Some examples of constellations in the northern hemispheres are Leo, Pisces, Ursa Major, Cassiopeia, Andromeda, etc. and some examples of southern hemispheres constellations are telescopian, Musca, Tucana, etc.

Do you know?

The biggest constellation is Hydra which extends over more than 3% of the night sky, while the smallest is Crux covering a mere 0.165%. Centaurus contains the largest number of visible stars at 101.

The Earth with its inclined axis revolves round the sun from west to east. This is why some of the constellations of one season cannot be seen in the other season at the same watch time from the same place. Some constellations can be seen from both the hemispheres too but at difference watch time.

The position of the pole star (Dhruva Tara) can be located with the help of Ursa Major. All constellations appear to revolve around the pole star which remains stationary throughout the year. This is because it is straight above the North Pole on the axis of rotation of earth.

Preparation of a model of Constellation

When we construct a model, we have to think of the positions of the zodiacal constellations compared with the positions of the earth in its orbit. We have to take into account the summer and the winter, i.e. the positions where the Earth is at the nearest or farthest away, in perihelion or aphelion.

We will use the overhead projector for drawing the constellations (only the brightest stars) on the cardboard. For making this model more understandable we could paint the mystic figures onto the constellations freehand or using transparencies. The model could work as a two-dimensional version on the wall or it could be more illustrative as a three-dimensional version in the middle of the class. Also we can demonstrate these things by letting the pupils be the parts of this model, each carrying a cardboard of one constellation. So for this, we need twelve "constellations", one "Earth" and one "Sun".

Activity

Construction of the zodiacal model

- Let us prepare groups of 2-3 persons.
- Each group will take one constellation to paint.
- First draw the zodiacal constellations to the cardboards.
- You will use the overhead projectors for reflecting the pictures from the transparencies to the wall in such a way that the distance between the overhead projector and the wall should be the same all the time.
- After drawing the constellation (only the brightest stars) we can change the transparency with the mystic figure.
- Then draw it onto the constellation in the right position and size.
- You can use some colours too.
- After painting all the constellations you are going to collect all parts together to form a compact system of zodiac-sun-earth.
- In the middle of the classroom you will make an elliptic circle, in the middle of it and there will be the sun and the earth.
- The orbit of the earth can be marked on the floor. Try to form the right order for the constellations and to take the seasons into account.

Planetarium

Planetarium is an educational device for showing the locations and movements of the planets and other objects in the universe. A modern planetarium is a complex optical instrument. It projects images of the planets, moon, and stars onto a domed ceiling, creating an accurate representation of the night time sky. It can be also defined as a place where we can go to see what the night sky looks like. Planetarium has a large room with a dome-shaped ceiling and many seats. A special projector in this room can shine images on the domed ceiling and show us the stars and other objects in the night sky. Many planetariums also have telescopes which we can look through and exhibits about space.

A typical planetarium forms images of the stars by focusing light from one or more bright lamps through thousands of tiny holes drilled through metal plates. The plates are positioned around two spherical structures, one for Northern Hemisphere stars and the other for Southern Hemisphere stars. The images of the moon and planets are produced by separate projection devices mounted on a frame between the two star spheres. By means of various sets of gears and electric motors, the planetarium can show the rising and setting of the stars and the motion of the moon and planets along the ecliptic. The planetarium can also show the appearance of the heavens from any given place on Earth at any given time far into the past or future. Additional projection devices are used for depicting such phenomena as eclipses, auroras, or meteors, and for showing such aids to instruction as the system of celestial coordinates or outlines of the constellations.

Do you know?

The term “planetarium” was initially used to represent a small model of the solar system. The first planetarium was designed in the year 1682 by the famous astronomer Christiaan Huygens. Then, throughout the next 150 years, thousands of model planetariums were built that consisted a series of balls which showcased the sun and the planets of the solar system. These balls were fixed to rods connecting the gears that would further move the rods in order to make the planets circle the sun.

Uses of Planetarium

- The large dome-shaped projection screen on to which scenes of stars, planets and other celestial objects can be made to appear.
- It can be made to move realistically to stimulate the complex motions of the heavens.
- The celestial scenes can be created using a wide variety of technologies such as slide projector, video and full dome projector systems, lasers etc.
- Typical systems can be set to display the sky at any point of time.

Summary

1. The International Space Station is a research laboratory in low Earth orbit. With many different partners contributing to its design and construction, this high-flying laboratory has become a symbol of cooperation in space exploration.
2. In 1942 the German V2 was the first rocket to reach 100 km from the Earth's surface (the boundary of space). The rocket was designed by Wernher Von Braun, who later worked with NASA as the creator of the rockets that went to the moon.
3. In 1947, the first animals were launched into space. Fruit flies were used to study the effects of space travel on animals, and were chosen because they are more similar to humans than you might imagine.
4. On 12th April 1961, Russian Cosmonaut Yuri Gagarin became the first man in space. Gagarin's spacecraft, Vostok 1, completed one orbit of the earth, and landed about two hours after launch.
5. The first woman in space was Russian cosmonaut Valentina Tereshkova. After her 1963 mission, Valentina became an important member of the Russian Government, and has been awarded many honours and prizes for her achievements. A crater on the far side of the Moon is named after her.
6. The U.S. National Space Policy of 2010 set out goals for space exploration; to send humans to an asteroid by 2025 and to the planet Mars in the 2030s. Many of the astronauts that will be involved in these exciting missions are only children right now.
7. A telescope is defined as the virtual image forming optical instrument which is used to view distant objects clearly.
8. The refractor telescope uses a lens to gather and focus light. The first telescopes built were refractors. The small telescopes sold in department stores are refractors, as well as, those used for rifle scopes.

9. The reflector telescope uses a mirror to gather and focus light. It is more beneficial over refractor telescope.
10. The small group of stars seen in the sky especially at night with fixed shape is called constellation. Constellation available in the sky is of different shapes and sizes. There are 88 constellations known to us till this time.
11. The Earth revolves round the sun from west to east. This is why the constellations of one season cannot be seen in the other season. Some constellations can be seen from both hemispheres throughout the year at different watch time.
12. Planetarium is an educational device for showing the locations and movements of the planets and other objects in the universe. A modern planetarium is a complex optical instrument.

Exercise

A. Tick (√) the best alternative from the following:

1. When was Soviets launched the first artificial satellite?

   i. On October 4, 1957   ii. On October 4, 1958

   iii.On October 8, 1957   iv. On October 8, 1958

2. On 12th April 1961, who became the first man to reach in the space?

   i. Russian Cosmonaut Yuri Gagarin

   ii.American Cosmonaut Yuri Gagarin

   iii. Russian Cosmonaut Neil Armonstrong

   iv. American Cosmonaut Neil Armonstrong

3. Which of the following formulae is correct in the case of telescope?

   i. $\text{Magnification} = \dfrac{\text{Focal length of the objective (F)}}{\text{Focal length of eyepiece lens (f)}}$

   ii. $\text{Magnification} = \dfrac{\text{Focal length of the eyepiece lens (F)}}{\text{Focal length of objective lens (f)}}$

   iii. $\text{Magnification} = \dfrac{\text{Focal length of the objective (F)}}{\text{Focal length of objective lens (f)}}$

   iv. Both (i) and (ii)

4. Ursa Major can be seen clearly in the month of:
   i. April in the northern part of the sky
   ii. May in the northern part of the sky
   iii. April in the southern part of the sky
   iv. May in the southern part of the sky
5. The star at the end of Ursa Minor is:
   i. Pole star ii. Ursa Minor iii. Orion iv. Vesta

B. Give short answers to the following questions:

1. What is a telescope?
2. Define an astronomical telescope.
3. What is the principle under which astronomical telescope constructed?
4. Define constellation.
5. How many constellations are known to us?
6. Name any four examples of constellation.
7. What is planetarium?

C. Give long answers to the following questions:

1. How does an astronomical telescope work? Explain.
2. Mention the uses of an astronomical telescope.
3. How do you prepare a model of constellation? Explain.
4. Mention the importance of constellation in our daily life.
5. Point out the uses of Planetarium.

Project work

1. Take an astronomical telescope from the market or prepare yourself using essential materials, and observe the sky at clear night, and find the constellations, galaxy and other stars. On the basis of this information, prepare a report and present it to the class.
2. Prepare a model of constellation (sign of zodiac), and prepare a report on the basis of it and present it to the class.

Glossary

Rugged: uneven surface

Chromatic aberration: color deviation or distortion