

Note: Please refer to the text book for any clarifications

NCERT Physics Class XI, Part I

CHAPTER - 1

PHYSICAL WORLD

SUMMARY

1. Physics deals with the study of the basic laws of nature and their manifestation in different phenomena. The basic laws of physics are universal and apply in widely different contexts and conditions.
2. The scope of physics is wide, covering a tremendous range of magnitude of physical quantities.
3. Physics and technology are related to each other. Sometimes technology gives rise to new physics; at other times physics generates new technology. Both have direct impact on society.
4. There are four fundamental forces in nature that govern the diverse phenomena of the macroscopic and the microscopic world. These are the 'gravitational force', the 'electromagnetic force', the 'strong nuclear force', and the 'weak nuclear force'. Unification of different forces/domains in nature is a basic quest in physics.
5. The physical quantities that remain unchanged in a process are called conserved quantities. Some of the general conservation laws in nature include the laws of conservation of mass, energy, linear momentum, angular momentum, charge, parity, etc. Some conservation laws are true for one fundamental force but not for the other.
6. Conservation laws have a deep connection with symmetries of nature. Symmetries of space and time, and other types of symmetries play a central role in modern theories of fundamental forces in nature.

CHAPTER – 2

UNITS AND MEASUREMENT

| Base quantity | SI Units | | |
|----------------------------|----------|--------|--|
| | Name | Symbol | Definitions |
| Length | metre | m | The metre is the length of the path travelled by light in vacuum during a time interval of $1/299,792,458$ of a second. (1983) |
| Mass | kilogram | kg | The kilogram is equal to the mass of the international prototype of the kilogram (a platinum-iridium alloy cylinder) kept at international Bureau of Weights and Measures, at Sevres, near Paris, France. (1889) |
| Time | second | s | The second is the duration of 9,192,631,770 periods of the radiation corresponding to the transition between the two hyperfine levels of the ground state of the cesium-133 atom. (1967) |
| Electric current | ampere | A | The ampere is that constant current which, if maintained in current two straight parallel conductors of infinite length, of negligible circular cross-section, and placed 1 metre apart in vacuum, would produce between these conductors a force equal to 2×10^{-7} Newton per metre of length. (1948) |
| Thermo dynamic Temperature | kelvin | K | The Kelvin, is the fraction $1/273.16$ of the thermodynamic dynamic temperature of the triple point of water. (1967) |
| Amount of substance | mole | mol | The mole is the amount of substance of a system, which contains substance as many elementary entities as there are atoms in 0.012 kilogram of carbon - 12. (1971) |
| Luminous intensity | candela | cd | The candela is the luminous intensity, in a given Intensity direction, of a source that emits monochromatic radiation of frequency 540×10^{12} hertz and that has a radiant intensity in that direction of $1/683$ watt per Steradian. (1979) |

| Name | Symbol | Value in SI Units |
|-------------------------------|--------|--|
| minute | min | 60 s |
| hour | h | 60 min = 3600 s |
| day | d | 24 h = 86400 s |
| year | y | 365.25 d = 3.156×10^7 s |
| degree | ° | $1^\circ = (\pi/180)rad$ |
| litre | L | $1 \text{ dm}^3 = 10^{-3} \text{ m}^3$ |
| tonne | t | 10^3 kg |
| carat | c | 200 mg |
| bar | bar | $0.1 \text{ MPa} = 10^5 \text{ Pa}$ |
| curie | Ci | $3.7 \times 10^{10} \text{ s}^{-1}$ |
| roentgen | R | $2.58 \times 10^{-4} \text{ C/kg}$ |
| quintal | q | 100 kg |
| barn | b | $100 \text{ fm}^2 = 10^{-28} \text{ m}^2$ |
| arc | a | $1 \text{ dam}^2 = 10^{-2} \text{ m}^2$ |
| hectare | ha | $1 \text{ hm}^2 = 10^4 \text{ m}^2$ |
| standard atmospheric pressure | atm | $101325 \text{ Pa} = 1.013 \times 10^5 \text{ Pa}$ |

SUMMARY

1. Physics is a quantitative science, based on measurement of physical quantities. Certain physical quantities have been chosen as fundamental or base quantities (such as length, mass, time, electric current, thermodynamic temperature, amount of substance, and luminous intensity).
2. Each base quantity is defined in terms of a certain basic, arbitrarily chosen but properly standardized reference standard called unit (such as meter, kilogram, second, ampere, Kelvin, mole and candela). The units for the fundamental or base quantities are called fundamental or base units.
3. Other physical quantities, derived from the base quantities, can be expressed as a combination of the base units and are called derived units. A complete set of units, both fundamental and derived, is called a system of units.
4. The International System of Units (SI) based on seven base units is at present internationally accepted unit system and is widely used throughout the world.
5. The SI units are used in all physical measurements, for both the base quantities and the derived quantities obtained from them. Certain derived units are expressed by means of SI units with special names (such as Joule, Newton, watt, etc).

6. The SI units have well defined and internationally accepted unit symbols (such as m for meter, kg for kilogram, s for second, A for ampere, N for Newton etc.).
7. Physical measurements are usually expressed for small and large quantities in scientific notation, with powers of 10. Scientific notation and the prefixes are used to simplify measurement notation and numerical computation, giving indication to the precision of the numbers.
8. Certain general rules and guidelines must be followed for using notations for physical quantities and standard symbols for SI units, some other units and SI prefixes for expressing properly the physical quantities and measurements.
9. In computing any physical quantity, the units for derived quantities involved in the relationship(s) are treated as though they were algebraic quantities till the desired units are obtained.
10. Direct and indirect methods can be used for the measurement of physical quantities. In measured quantities, while expressing the result, the accuracy and precision of measuring instruments along with errors in measurements should be taken into account.
11. In measured and computed quantities proper significant figures only should be retained. Rules for determining the number of significant figures, carrying out arithmetic operations with them, and 'rounding off' the uncertain digits must be followed.
12. The dimensions of base quantities and combination of these dimensions describe the nature of physical quantities. Dimensional analysis can be used to check the dimensional consistency of equations, deducing relations among the physical quantities, etc. A dimensionally consistent equation need not be actually an exact (correct) equation, but a dimensionally wrong or inconsistent equation must be wrong.

Range of Lengths

The sizes of the objects we come across in the universe vary over a very wide range. These may vary from the size of the order of 10^{-14} m of the tiny nucleus of an atom to the size of the order of 10^{26} m of the extent of the observable universe. Table 2.3 gives the range and order of length and sizes of some of these objects.

We also use certain special length units for short and large lengths. These are

| | |
|---------------------|---|
| 1 Fermi | = 1 f = 10^{-15} m |
| 1 angstrom | = 1 Å = 10^{-10} m |
| 1 astronomical unit | = 1 AU (average distance of the Sun from the Earth) = 1.496×10^{11} m |
| 1 light year | = 1 ly = 9.46×10^{15} m (distance that light travels with velocity of 3×10^8 m s ⁻¹ in 1 year) |
| 1 parsec | = 3.08×10^{16} m (Parsec is the distance at which average radius of |

earth's orbit subtends an angle of 1 arc second)

While dealing with atoms and molecules, the kilogram is an inconvenient unit. In this case, there is an important standard unit of mass, called the unified atomic mass unit (u), which has been established for expressing the mass of atoms as,

1 unified atomic mass unit = $1u = (1/12)$ of the mass of an atom of carbon-12 isotope ($^{12}_6\text{C}$) including the mass of electrons = 1.66×10^{-27} kg.

Mass of commonly available objects can be determined by a common balance like the one used in a grocery shop. Large masses in the universe like planets, stars, etc., based on Newton's law of gravitation can be measured by using gravitational method (See Chapter 8). For measurement of small masses of atomic/subatomic particles etc., we make use of mass spectrograph in which radius of the trajectory is proportional to the mass of a charged particle moving in uniform electric and magnetic field.

ACCURACY, PRECISION OF INSTRUMENTS AND ERRORS IN MEASUREMENT

Measurement is the foundation of all experimental science and technology. The result of every measurement by any measuring instrument contains some uncertainty. This uncertainty is called error. Every calculated quantity which is based on measured values also has an error. We shall distinguish between two terms: accuracy and precision. The accuracy of a measurement is a measure of how close the measured value is to the true value of the quantity. Precision tells us to what resolution or limit the quantity is measured.

The accuracy in measurement may depend on several factors, including the limit or the resolution of the measuring instrument.

SYSTEMATIC ERRORS

The systematic errors are those errors that tend to be in one direction, either positive or negative. Some of the sources of systematic errors are :

- (a) **Instrumental errors** that arise from the errors due to imperfect design or calibration of the measuring instrument, zero error in the instrument, etc. For example, the temperature graduations of a thermometer may be inadequately calibrated (it may read 104°C at the boiling point of water at STP whereas it should read 100°C); in a vernier calipers the zero mark of vernier scale may not coincide with the zero mark of the main scale, or simply an ordinary meter scale may be worn off at one end.
- (b) **Imperfection in experimental technique or procedure** To determine the temperature of a human body, a thermometer placed under the armpit will always give a temperature

lower than the actual value of the body temperature. Other external conditions (such as changes in temperature, humidity, wind velocity, etc.) during the experiment may systematically affect the measurement.

- (c) **Personal errors** that arise due to an individual's bias, lack of proper setting of the apparatus or individual's carelessness in taking observations without observing proper precautions, etc. For example, if you, by habit, always hold your head a bit too far to the right while reading the position of a needle on the scale, you will introduce an error due to parallax.

Systematic errors can be minimized by improving experimental techniques, selecting better instruments and removing personal bias as far as possible. For a given set-up, these errors may be estimated to a certain extent and the necessary corrections may be applied to the readings.

RANDOM ERRORS

The random errors are those errors, which occur irregularly and hence are random with respect to sign and size. These can arise due to random and unpredictable fluctuations in experimental conditions (e.g. unpredictable fluctuations in temperature, voltage supply, mechanical vibrations of experimental set-ups, etc.), personal (unbiased) errors by the observer taking readings, etc. For example, when the same person repeats the same observation, it is very likely that he may get different readings every time.

LEAST COUNT ERROR

The smallest value that can be measured by the measuring instrument is called its least count. All the readings or measured values are good only up to this value.

The **least count error** is the error associated with the resolution of the instrument. For example, a vernier calipers has the least count as 0.01 cm; a spherometer may have a least count of 0.001 cm. Least count error belongs to the category of random errors but within a limited size; it occurs with both systematic and random errors. If we use a metre scale for measurement of length, it may have graduations at 1 mm division scale spacing or interval.

Using instruments of higher precision, improving experimental techniques, etc., we can reduce the least count error.

ABSOLUTE ERROR, RELATIVE ERROR AND PERCENTAGE ERROR

The example gives the following rules:

- All the non-zero digits are significant.

- All the zero's between two non-zero digits are significant, no matter where the decimal point is, if at all.
- If the number is less than 1, the zero(s) on the right of decimal point but to the left of the first non-zero digit are not significant. [In 0.00 2308, the under lined zeroes are not significant].
- The terminal or trailing zero(s) in a number without a decimal point are not significant.

$$[V] = [M^0 L^3 T^0]$$

$$[v] = [M^0 L T^{-1}]$$

$$[F] = [M L T^{-2}]$$

$$[\rho] = [M L^{-3} T^0]$$

CHAPTER – 3

MOTION IN A STRAIGHT LINE

SUMMARY

1. An object is said to be in *motion* if its position changes with time. The position of the object can be specified with reference to a conveniently chosen origin. For motion in a straight line, position to the right of the origin is taken as positive and to the left as negative.
2. Path length is defined as the total length of the path traversed by an object.
3. Displacement is the change in position: $\Delta X = X_2 - X_1$. Path length is greater or equal to the magnitude of the displacement between the same points.
4. An object is said to be in uniform motion in a straight line if its displacement is equal in equal intervals of time. Otherwise, the motion is said to be non-uniform.
5. Average velocity is the displacement divided by the time interval in which the displacement occurs:

$$\bar{v} = \frac{\Delta x}{\Delta t}$$

On an $x-t$ graph, the average velocity over a time interval is the slope of the line connecting the initial and final positions corresponding to that interval.

6. Average Speed is the ratio of total path length traversed and the corresponding time interval.

The average speed of an object is greater or equal to the magnitude of the average velocity over a given time interval.

7. Instantaneous velocity or simply velocity is defined as the limit of the average velocity as the time interval Δt becomes infinitesimally small:

$$v = \lim_{\Delta t \rightarrow 0} \bar{v} = \lim_{\Delta t \rightarrow 0} \frac{\Delta x}{\Delta t} = \frac{dx}{dt}$$

The velocity at a particular instant is equal to the slope of the tangent drawn on position-time graph at that instant.

8. Average acceleration is the change in velocity divided by the time interval during which the change occurs:

$$\bar{a} = \frac{\Delta v}{\Delta t}$$

9. Instantaneous acceleration is defined as the limit of the average acceleration as the time interval Δt goes to zero:

$$a = \lim_{\Delta t \rightarrow 0} \bar{a} = \lim_{\Delta t \rightarrow 0} \frac{\Delta v}{\Delta t} = \frac{dv}{dt}$$

The acceleration of an object at a particular time is the slope of the velocity-time graph at that instant of time. For uniform motion, acceleration is zero and the x - t graph is a straight line inclined to the time axis and the v - t graph is a straight line parallel to the time axis. For motion with uniform acceleration, x - t graph is a parabola while the v - t graph is a straight line inclined to the time axis.

10. The area under the velocity-time curve between times t_1 and t_2 is equal to the displacement of the object during that interval of time.
11. For objects in uniformly accelerated rectilinear motion, the five quantities, displacement x , time taken t , initial velocity v_0 , final velocity v and acceleration a are related by a set of simple equations called kinematic equations of motion :

$$\begin{aligned} v &= v_0 + at \\ x &= v_0 t + \frac{1}{2} at^2 \\ v^2 &= v_0^2 + 2ax \end{aligned}$$

If the position of the object at time $t = 0$ is 0. If the particle starts at $x = x_0$, x in above equations is replaced by $(x - x_0)$.

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CHAPTER - 4

MOTION IN A PLANE

SUMMARY

- Scalar quantities are quantities with magnitudes only. Examples are distance, speed, mass and temperature.
- Vector quantities are quantities with magnitude and direction both. Examples are is placement, velocity and acceleration. They obey special rules of vector algebra.
- A vector A multiplied by a real number λ is also a vector, whose magnitude is λ times the magnitude of the vector A and whose direction is the same or opposite depending upon whether λ is positive or negative.
- Two vectors A and B may be added graphically using head-to-tail method or parallelogram method.
- Vector addition is commutative:

$$A + B = B + A$$
 It also obeys the associative law:

$$(A + B) + C = A + (B + C)$$
- A null or zero vector is a vector with zero magnitude. Since the magnitude is zero, we don't have to specify its direction. It has the properties:

$$A + 0 = A$$

$$\lambda 0 = 0$$

$$0 A = 0$$
- The subtraction of vector B from A is defined as the sum of A and $-B$:

$$A - B = A + (-B)$$
- A vector A can be resolved into component along two given vectors 'a' and 'b' lying in the same plane:

$$A = \lambda a + \mu b$$
 where λ and μ are real numbers.
- A unit vector associated with a vector A has magnitude one and is along the vector A :

$$\hat{n} = \frac{A}{|A|}$$
 The unit vectors $\hat{i}, \hat{j}, \hat{k}$ are vectors of unit magnitude and point in the direction of the x- y and z-axes, respectively in a right-handed coordinate system.
- A vector A can be expressed as

$$A = A_x \hat{i} + A_y \hat{j}$$

Where A_x, A_y are its components along x-, and y -axes. If vector A makes an angle θ with the x-axis, then $A_x = A \cos \theta$, $A_y = A \sin \theta$ and $A = |A| = \sqrt{A_x^2 + A_y^2}$, $\tan \theta = \frac{A_y}{A_x}$.

11. Vectors can be conveniently added using analytical method. If sum of two vectors A and B , that lie in x-y plane, is R , then:

$$R = R_x \hat{i} + R_y \hat{j}, \text{ where, } R_x = A_x + B_x, \text{ and } R_y = A_y + B_y$$

12. The position vector of an object in x-y plane is given by $r = x \hat{i} + y \hat{j}$ and the displacement from position r to position r' is given by

$$\begin{aligned} \Delta r &= r' - r \\ &= (x' - x)\hat{i} + (y' - y)\hat{j} \\ &= \Delta x \hat{i} + \Delta y \hat{j} \end{aligned}$$

13. If an object undergoes a displacement Δr in time Δt , its average velocity is given by $v = \frac{\Delta r}{\Delta t}$. The velocity of an object at time t is the limiting value of the average velocity as Δt tends to zero:

$$v = \lim_{\Delta t \rightarrow 0} \frac{\Delta r}{\Delta t} = \frac{dr}{dt}$$

It can be written in unit vector notation as :

$$v = v_x \hat{i} + v_y \hat{j} + v_z \hat{k} \quad \text{where } v_x = \frac{dx}{dt}, v_y = \frac{dy}{dt}, v_z = \frac{dz}{dt}$$

When position of an object is plotted on a coordinate system, v is always tangent to the curve representing the path of the object.

14. If the velocity of an object changes from v to v' in time Δt , then its average acceleration is given by:

$$\bar{a} = \frac{v - v'}{\Delta t} = \frac{\Delta v}{\Delta t}$$

The acceleration a at any time t is the limiting value of \bar{a} as $\Delta t \rightarrow 0$:

$$a = \lim_{\Delta t \rightarrow 0} \frac{\Delta v}{\Delta t} = \frac{dv}{dt}$$

In component form, we have: $a = a_x \hat{i} + a_y \hat{j} + a_z \hat{k}$

$$\text{where, } a_x = \frac{dv_x}{dt}, a_y = \frac{dv_y}{dt}, a_z = \frac{dv_z}{dt}$$

15. If an object is moving in a plane with constant acceleration $a = |a| = \sqrt{a_x^2 + a_y^2}$ and its position vector at time $t = 0$ is r_o , then at any other time t , it will be at a point given by:

$$r = r_o + v_o t + \frac{1}{2} a t^2$$

and its velocity is given by :

$$v = v_o + a t$$

where v_o is the velocity at time $t = 0$

In component form:

$$x = x_o + v_{ox} t + \frac{1}{2} a_x t^2$$

$$y = y_o + v_{oy} t + \frac{1}{2} a_y t^2$$

$$v_x = v_{ox} + a_x t$$

$$v_y = v_{oy} + a_y t$$

Motion in a plane can be treated as superposition of two separate simultaneous one-dimensional motions along two perpendicular directions.

16. An object that is in flight after being projected is called a projectile. If an object is projected with initial velocity v_o making an angle θ_o with x-axis and if we assume its initial position to coincide with the origin of the coordinate system, then the position and velocity of the projectile at time t are given by :

$$x = (v_o \cos \theta_o) t$$

$$y = (v_o \sin \theta_o) t - (1/2) g t^2$$

$$v_x = v_{ox} = v_o \cos \theta_o$$

$$v_y = v_o \sin \theta_o - g t$$

The path of a projectile is parabolic and is given by:

$$y = (\tan \theta_o) x - \frac{g x^2}{2(v_o \cos \theta_o)^2}$$

The maximum height that a projectile attains is:

$$h_m = \frac{(v_o \sin \theta_o)^2}{2g}$$

The time taken to reach this height is:

$$t_m = \frac{v_o \sin \theta_o}{g}$$

The horizontal distance travelled by a projectile from its initial position to the position it passes $y = 0$ during its fall is called the range, R of the projectile. It is:

$$R = \frac{v_0^2}{g} \sin 2\theta_n$$

17. When an object follows a circular path at constant speed, the motion of the object is called uniform circular motion. The magnitude of its acceleration is $a_c = v^2/R$. The direction of a_c is always towards the centre of the circle.

The angular speed ω , is the rate of change of angular distance. It is related to velocity v by $v = \omega R$. The acceleration is $a_c = \omega^2 R$.

If T is the time period of revolution of the object in circular motion and ν is its frequency, we have $\omega = 2\pi\nu$, $v = 2\pi\nu R$, $a_c = 4\pi^2\nu^2 R$.

| Physical Quantity | Symbol | Dimensions | Unit | Remark |
|-------------------------|------------|---------------------|-------------------|--|
| Position vector | r | [L] | m | Vector. It may be denoted by any other symbol as well. -do- |
| Displacement | Δr | [L] | m | |
| Velocity | | | | |
| (a) Average | \bar{v} | [LT ⁻¹] | m s ⁻¹ | $= \frac{\Delta r}{\Delta t}$, <i>vector</i> |
| (b) Instantaneous | v | | | $= \frac{dr}{dt}$, <i>vector</i> |
| Acceleration | | | | |
| (a) Average | \bar{a} | [LT ⁻²] | m s ⁻² | $= \frac{\Delta v}{\Delta t}$, <i>vector</i> |
| (b) Instantaneous | a | | | $= \frac{dv}{dt}$, <i>vector</i> |
| Projectile motion | | | | |
| (a) Time of max. Height | t_m | [T] | s | $= \frac{v_0 \sin \theta_0}{g}$ |
| (b) Max. Height | h_m | [L] | m | $= \frac{(v_0 \sin \theta_0)^2}{2g}$ |

| | | | | |
|------------------------------|----------|-------------|-------------|---|
| (c) Horizontal range | R | [L] | m | $= \frac{v_0^2 \sin 2\theta_0}{g}$ |
| Circular motion | | | | |
| (a) Angular speed | ω | $[T^{-1}]$ | rad/s | $= \frac{\Delta\theta}{\Delta t} = \frac{v}{r}$ |
| (b) Centripetal acceleration | a_c | $[LT^{-2}]$ | $m\ s^{-2}$ | $= \frac{v^2}{r}$ |

CHAPTER - 5

LAWS OF MOTION

SUMMARY

1. Aristotle's view that a force is necessary to keep a body in uniform motion is wrong. A force is necessary in practice to counter the opposing force of friction.
2. Galileo extrapolated simple observations on motion of bodies on inclined planes, and arrived at the law of inertia.
- 2.1 Newton's first law of motion is the same law rephrased thus: "Everybody continues to be in its state of rest or of uniform motion in a straight line, unless compelled by some external force to act otherwise". In simple terms, the First Law is "If external force on a body is zero, its acceleration is zero".
3. Momentum (p) of a body is the product of its mass (m) and velocity (v):
$$p = m v$$
4. Newton's second law of motion

The rate of change of momentum of a body is proportional to the applied force and takes place in the direction in which the force acts. Thus

$$F = k \frac{dp}{dt} = k m a$$

Where F is the net external force on the body and a its acceleration. We set the constant of proportionality $k = 1$ in SI units. Then

$$F = \frac{dp}{dt} = m a$$

The SI unit of force is Newton: $1 \text{ N} = 1 \text{ kg m s}^{-2}$.

- (a) The second law is consistent with the First Law ($F = 0$ implies $a = 0$)
- (b) It is a vector equation.
- (c) It is applicable to a particle, and also to a body or a system of particles, provided F is the total external force on the system and ' a ' is the acceleration of the system as a whole.
- (d) F at a point at a certain instant determines a at the same point at that instant. That is the

Second Law is a local law; ' a ' at an instant does not depend on the history of motion.

5. Impulse is the product of force and time which equals change in momentum. The notion of impulse is useful when a large force acts for a short time to produce a measurable change in momentum. Since the time of action of the force is very short, one can assume that there is no appreciable change in the position of the body during the action of the impulsive force.

6. **Newton's third law of motion**

To every action, there is always an equal and opposite reaction

In simple terms, the law can be stated thus:

Forces in nature always occur between pairs of bodies. Force on a body A by body B is equal and opposite to the force on the body B by A.

Action and reaction forces are simultaneous forces. There is no cause-effect relation between action and reaction. Any of the two mutual forces can be called action and the other reaction. Action and reaction act on different bodies and so they cannot be cancelled out. The internal action and reaction forces between different parts of a body do, however, sum to zero.

7. **Law of Conservation of Momentum**

The total momentum of an isolated system of particles is conserved. The law follows from the second and third law of motion.

8. **Friction**

Frictional force opposes (impending or actual) relative motion between two surfaces in contact. It is the component of the contact force along the common tangent to the surface in contact. Static friction f_s opposes impending relative motion; kinetic friction f_k opposes actual relative motion. They are independent of the area of contact and satisfy the following approximate laws:

$$f_s \leq (f_s)_{\max} = \mu_s R$$

μ_s (co-efficient of static friction) and μ_k (co-efficient of kinetic friction) are constants characteristic of the pair of surfaces in contact. It is found experimentally that μ_k is less than μ_s .

CHAPTER – 6

WORK, ENERGY AND POWER

SUMMARY

- The work-energy theorem states that the change in kinetic energy of a body is the work done by the net force on the body.

$$K_f - K_i = W_{\text{net}}$$

- A force is conservative if (i) work done by it on an object is path independent and depends only on the end points $\{x_i, x_j\}$, or (ii) the work done by the force is zero for an arbitrary closed path taken by the object such that it returns to its initial position.
- For a conservative force in one dimension, we may define a potential energy function $V(x)$ such that

$$F(x) = -\frac{dV(x)}{dx} \text{ or}$$

$$V_i - V_j = \int_{x_i}^{x_j} F(x) dx$$

- The principle of conservation of mechanical energy states that the total mechanical energy of a body remains constant if the only forces that act on the body are conservative.
- The gravitational potential energy of a particle of mass m at a height x about the earth's surface is

$$V(x) = mgx$$

where the variation of g with height is ignored.

- The elastic potential energy of a spring of force constant k and extension x is

$$V(x) = \frac{1}{2} kx^2$$

7. The scalar or dot product of two vectors **A** and **B** is written as **A.B** and is a scalar quantity given by: $\mathbf{A} \cdot \mathbf{B} = AB \cos \theta$, where θ is the angle between **A** and **B**. It can be positive, negative or zero depending upon the value of θ . The scalar product of two vectors can be interpreted as the product of magnitude of one vector and component of the other vector along the first vector. For unit vectors:

$$\hat{i} \cdot \hat{i} = \hat{j} \cdot \hat{j} = \hat{k} \cdot \hat{k} = 1 \quad \text{and} \quad \hat{i} \cdot \hat{j} = \hat{j} \cdot \hat{k} = \hat{k} \cdot \hat{i} = 0$$

Scalar products obey the commutative and the distributive laws.

CHAPTER - 7

SYSTEMS OF PARTICLES AND ROTATIONAL MOTION

SUMMARY

1. Ideally, a rigid body is one for which the distances between different particles of the body do not change, even though there are forces on them.
2. A rigid body fixed at one point or along a line can have only rotational motion. A rigid body not fixed in some way can have either pure translation or a combination of translation and rotation.
3. In rotation about a fixed axis, every particle of the rigid body moves in a circle which lies in a plane perpendicular to the axis and has its centre on the axis. Every Point in the rotating rigid body has the same angular velocity at any instant of time.
4. In pure translation, every particle of the body moves with the same velocity at any instant of time.
5. Angular velocity is a vector. Its magnitude is $\omega = d\theta/dt$ and it is directed along the axis of rotation. For rotation about a fixed axis, this vector ω has a fixed direction
6. The vector or cross product of two vector **a** and **b** is a vector written as **a x b**. The magnitude of this vector is $ab \sin \theta$ and its direction is given by the right handed screw or the right hand rule.
7. The linear velocity of a particle of a rigid body rotating about a fixed axis is given by $v = \omega \times r$, where **r** is the position vector of the particle with respect to an origin along the fixed axis. The relation applies even to more general rotation of a rigid body with one point fixed. In that case **r** is the position vector of the particle with respect to the fixed point taken as the origin.

8. The centre of mass of a system of particles is defined as the point whose position vector is

$$R = \frac{\sum m_i r_i}{M}$$

9. Velocity of the centre of mass of a system of particles is given by $V = P/M$, where P is the linear momentum of the system. The centre of mass moves as if all the mass of the system is concentrated at this point and all the external forces act at it. If the total external force on the system is zero, then the total linear momentum of the system is constant.

10. The angular momentum of a system of n particles about the origin is

$$L = \sum_{i=1}^n r_i \times p_i$$

The torque or moment of force on a system of n particles about the origin is

$$\tau = \sum_{i=1}^n r_i \times F_i$$

The force F_i acting on the i^{th} particle includes the external as well as internal forces. Assuming Newton's third law and that forces between any two particles act along the line joining the particles, we can show $\tau_{int} = 0$ and

$$\frac{dL}{dt} = \tau_{ext}$$

11. A rigid body is in mechanical equilibrium if

(1) It is in translational equilibrium, i.e., the total external force on it is zero : $\sum F_i = 0$, and

(2) It is in rotational equilibrium, i.e. the total external torque on it is zero:

$$\sum \tau_i = \sum r_i \times F_i = 0$$

12. The centre of gravity of an extended body is that point where the total gravitational torque on the body is zero.

13. The moment of inertia of a rigid body about an axis is defined by the formula

$I = \sum m_i r_i^2$ where r_i is the perpendicular distance of the i^{th} point of the body from the axis. The kinetic energy of rotation is

$$K = \frac{1}{2} I \omega^2$$

14. The theorem of parallel axes: $I'_z = I_z + Ma^2$, allows us to determine the moment of inertia of a rigid body about an axis as the sum of the moment of inertia of the body about a parallel axis through its centre of mass and the product of mass and square of the perpendicular distance between these two axes.

15. Rotation about a fixed axis is directly analogous to linear motion in respect of kinematics and dynamics.
16. For a rigid body rotating about a fixed axis (say, z-axis) of rotation, $L_z = I\omega$, where I is the moment of inertia about z-axis. In general, the angular momentum L for such a body is not along the axis of rotation. Only if the body is symmetric about the axis of rotation, L is along the axis of rotation. In that case, $|L| = L_z = I\omega$. The angular acceleration of a rigid body rotating about a fixed axis is given by $I\alpha = \tau$. If the external torque τ acting on the body is zero, the component of angular momentum about the fixed axis (say, z-axis), $L_z (= I\omega)$ of such a rotating body is constant.
17. For rolling motion without slipping $v_{cm} = R\omega$, where v_{cm} is the velocity of translation (i.e., of the centre of mass), R is the radius and m is the mass of the body. The kinetic energy of such a rolling body is the sum of kinetic energies of translation and rotation:

$$K = \frac{1}{2}mv_{cm}^2 + \frac{1}{2}I\omega^2$$

CHAPTER - 8

GRAVITATION

SUMMARY

1. Newton's law of universal gravitation states that the gravitational force of attraction between any two particles of masses m_1 and m_2 separated by a distance 'r' has the magnitude

$$F = G \frac{m_1 m_2}{r^2}$$

Where G is the universal gravitational constant, which has the value $6.672 \times 10^{-11} \text{ Nm}^2 \text{ kg}^{-2}$.

2. If we have to find the resultant gravitational force acting on the particle 'm' due to a number of masses M_1, M_2, \dots, M_n etc. we use the principle of superposition. Let F_1, F_2, \dots, F_n be the individual forces due to M_1, M_2, \dots, M_n , each given by the law of gravitation. From the principle of superposition each force acts independently and uninfluenced by the other bodies. The resultant force F_R is then found by vector addition

$$F_R = F_1 + F_2 + \dots + F_n = \sum_{i=1}^n F_i$$

where the symbol ' Σ ' stands for summation.

3. Kepler's laws of planetary motion state that

- (a) All planets move in elliptical orbits with the Sun at one of the focal points.
 (b) The radius vector drawn from the Sun to a planet sweeps out equal areas in equal time intervals. This follows from the fact that the force of gravitation on the planet is central and hence angular momentum is conserved.
 (c) The square of the orbital period of a planet is proportional to the cube of the semimajor axis of the elliptical orbit of the planet.

The period 'T' and radius 'R' of the circular orbit of a planet about the Sun are related

$$T^2 = \left(\frac{4\pi^2}{G M_s} \right) R^3$$

Where M_s is the mass of the Sun. Most planets have nearly circular orbits about the Sun. For elliptical orbits, the above equation is valid if R is replaced by the semi-major axis, a .

4. The acceleration due to gravity.
 (a) At a height h above the earth's surface

$$g(h) = \frac{G M_E}{(R_E + h)^2}$$

$$\approx \frac{G M_E}{R_E^2} \left(1 - \frac{2h}{R_E} \right) \text{ for } h \ll R_E$$

$$g(h) = g(0) \left(1 - \frac{2h}{R_E} \right) \text{ where } g(0) = \frac{G M_E}{R_E^2}$$

- (b) At depth 'd' below the earth's surface is

$$g(d) = \frac{G M_E}{R_E^2} \left(1 - \frac{d}{R_E} \right) = g(0) \left(1 - \frac{d}{R_E} \right)$$

5. The gravitational force is a conservative force, and therefore a potential energy function can be defined. The gravitational potential energy associated with two particles separated by a distance 'r' is given by

$$V = - \frac{G m_1 m_2}{r}$$

where 'V' is taken to be zero at $r \rightarrow \infty$. The total potential energy for a system of particles is the sum of energies for all pairs of particles, with each pair represented by a term of the form given by above equation. This prescription follows from the principle of Superposition.

6. If an isolated system consists of a particle of mass 'm' moving with a speed 'v' in the vicinity of a massive body of mass 'M', the total mechanical energy of the particle is given by

$$E = \frac{1}{2} m v^2 - \frac{G M m}{r}$$

That is, the total mechanical energy is the sum of the kinetic and potential energies. The total energy is a constant of motion.

7. If 'm' moves in a circular orbit of radius 'a' about 'M', where $M \gg m$, the total energy of the system is

$$E = - \frac{G M m}{2a}$$

With the choice of the arbitrary constant in the potential energy given in the point 5., above. The total energy is negative for any bound system, that is, one in which the orbit is closed, such as an elliptical orbit. The kinetic and potential energies are

$$K = \frac{G M m}{2a}$$

$$V = - \frac{G M m}{a}$$

8. The escape speed from the surface of the earth is

$$v_e = \sqrt{\frac{2 G M_E}{R_E}} = \sqrt{2gR_E}$$

and has a value of 11.2 km s^{-1} .

9. If a particle is outside a uniform spherical shell or solid sphere with a spherically symmetric internal mass distribution, the sphere attracts the particle as though the mass of the sphere or shell were concentrated at the centre of the sphere.
10. If a particle is inside a uniform spherical shell, the gravitational force on the particle is zero. If a particle is inside a homogeneous solid sphere, the force on the particle acts toward the centre of the sphere. This force is exerted by the spherical mass interior to the particle.

11. A geostationary (geosynchronous communication) satellite moves in a circular orbit in the equatorial plane at an approximate distance of 4.22×10^4 km from the earth's centre or 36000km from earth's surface.

POINTS TO PONDER

- In considering motion of an object under the gravitational influence of another object the following quantities are conserved:
 - Angular momentum
 - Total mechanical energy
 Linear momentum is not conserved
- Angular momentum conservation leads to Kepler's second law. However, it is not special to the inverse square law of gravitation. It holds for any central force.
- In Kepler's third law $T^2 = K_S R^3$. The constant K_S is the same for all planets in circular orbits. This applies to satellites orbiting the Earth.
- An astronaut experiences weightlessness in a space satellite. This is not because the gravitational force is small at that location in space. It is because both the astronaut and the satellite are in "free fall" towards the Earth.
- The gravitational potential energy associated with two particles separated by a distance 'r' is given by

$$V = - \frac{G m_1 m_2}{r} + \text{constant}$$

The constant can be given any value. The simplest choice is to take it to be zero. With this choice

$$V = - \frac{G m_1 m_2}{r}$$

This choice implies that $V \rightarrow 0$ as $r \rightarrow \infty$ choosing location of zero of the gravitational energy is the same as choosing the arbitrary constant in the potential energy. Note that the gravitational force is not altered by the choice of this constant.

- The total mechanical energy of an object is the sum of its kinetic energy (which is always positive) and the potential energy. Relative to infinity (i.e. if we presume that the potential energy of the object at infinity is zero), the gravitational potential energy of an object is negative. The total energy of a satellite is negative.

7. The commonly encountered expression $m g h$ for the potential energy is actually an approximation to the difference in the gravitational potential energy discussed in the point 6, above.
8. Although the gravitational force between two particles is central, the force between two finite rigid bodies is not necessarily along the line joining their centre of mass. For a spherically symmetric body however the force on a particle external to the body is as if the mass is concentrated at the centre and this force is therefore central.
9. The gravitational force on a particle inside a spherical shell is zero. However, (unlike a metallic shell which shields electrical forces) the shell does not shield other bodies outside it from exerting gravitational forces on a particle inside. Gravitational shielding is not possible.

Note: Please refer to the text book for any clarifications

NCERT Physics Class XI, Part II

Chapter – 9

Mechanical Properties of Solids

(Adopted for educational social service purpose only).

SUMMARY

1. Stress is the restoring force per unit area and strain is the fractional change in dimension. In general there are three types of stresses (a) tensile stress — longitudinal stress (associated with stretching) or compressive stress (associated with compression), (b) shearing stress, and (c) hydraulic stress.

- For small deformations, stress is directly proportional to the strain for many materials. This is known as Hooke's law. The constant of proportionality is called modulus of elasticity. Three elastic module *viz.*, Young's modulus, shear modulus and bulk modulus are used to describe the elastic behavior of objects as they respond to deforming forces that act on them.
A class of solids called elastomers does not obey Hooke's law.
- When an object is under tension or compression, the Hooke's law takes the form
$$F/A = Y\Delta L/L$$
where $\Delta L/L$ is the tensile or compressive strain of the object, F is the magnitude of the applied force causing the strain, A is the cross-sectional area over which F is applied (perpendicular to A) and Y is the Young's modulus for the object. The stress is F/A .
- A pair of forces when applied parallel to the upper and lower faces, the solid deforms so that the upper face moves sideways with respect to the lower. The horizontal displacement ΔL of the upper face is perpendicular to the vertical height L . This type of deformation is called shear and the corresponding stress is the shearing stress. This type of stress is possible only in solids.

In this kind of deformation the Hooke's law takes the form

$$F/A = G \times \Delta L/L$$

Where ΔL is the displacement of one end of object in the direction of the applied force F , and G is the shear modulus.

- When an object undergoes hydraulic compression due to a stress exerted by a surrounding fluid, the Hooke's law takes the form
$$p = B(\Delta V/V),$$
where p is the pressure (hydraulic stress) on the object due to the fluid, $\Delta V/V$ (the volume strain) is the absolute fractional change in the object's volume due to that pressure and B is the bulk modulus of the object.

POINTS TO PONDER

- In the case of a wire, suspended from ceiling and stretched under the action of a weight (F) suspended from its other end, the force exerted by the ceiling on it is equal and opposite to the weight. However, the tension at any cross-section A of the wire is just F

and not $2F$. Hence, tensile stress which is equal to the tension per unit area is equal to F/A .

2. Hooke's law is valid only in the linear part of stress-strain curve.
3. The Young's modulus and shear modulus are relevant only for solids since only solids have lengths and shapes.
4. Bulk modulus is relevant for solids, liquid and gases. It refers to the change in volume when every part of the body is under the uniform stress so that the shape of the body remains unchanged.
5. Metals have larger values of Young's modulus than alloys and elastomers. A material with large value of Young's modulus requires a large force to produce small changes in its length.
6. In daily life, we feel that a material which stretches more is more elastic, but it is a misnomer. In fact material which stretches to a lesser extent for a given load is considered to be more elastic.
7. In general, a deforming force in one direction can produce strains in other directions also. The proportionality between stress and strain in such situations cannot be described by just one elastic constant. For example, for a wire under longitudinal strain, the lateral dimensions (radius of cross section) will undergo a small change, which is described by another elastic constant of the material (called Poisson ratio).
8. Stress is not a vector quantity since, unlike a force, the stress cannot be assigned a specific direction. Force acting on the portion of a body on a specified side of a section has a definite direction.

Chapter - 10

Mechanical Properties of Fluids

SUMMARY

1. The basic property of a fluid is that it can flow. The fluid does not have any resistance to change of its shape. Thus, the shape of a fluid is governed by the shape of its container.
2. A liquid is incompressible and has a free surface of its own. A gas is compressible and it expands to occupy all the space available to it.
3. If F is the normal force exerted by a fluid on an area A then the average pressure P_{av} is defined as the ratio of the force to area

$$P_{av} = \frac{F}{A}$$

4. The unit of the pressure is the Pascal (Pa). It is the same as N m^{-2} . Other common units of pressure are
 $1 \text{ atm} = 1.01 \times 10^5 \text{ Pa}$
 $1 \text{ bar} = 10^5 \text{ Pa}$
 $1 \text{ torr} = 133 \text{ Pa} = 0.133 \text{ kPa}$
 $1 \text{ mm of Hg} = 1 \text{ torr} = 133 \text{ Pa}$

5. Pascal's law states that: Pressure in a fluid at rest is same at all points which are at the same height. A change in pressure applied to an enclosed fluid is transmitted undiminished to every point of the fluid and the walls of the containing vessel.

6. The pressure in a fluid varies with depth h according to the expression
 $P = P_a + \rho gh$
 where ρ is the density of the fluid, assumed uniform.

7. The volume of an incompressible fluid passing any point every second in a pipe of non uniform crosssection is the same in the steady flow.
 $vA = \text{constant}$ (v is the velocity and A is the area of cross section)
 The equation is due to mass conservation in incompressible fluid flow.

8. Bernoulli's principle states that as we move along a streamline, the sum of the pressure (P), the kinetic energy per unit volume ($\rho v^2/2$) and the potential energy per unit volume (ρgy) remains a constant.

$$P + \rho v^2/2 + \rho gy = \text{constant}$$

The equation is basically the conservation of energy applied to non viscous fluid motion in steady state. There is no fluid which have zero viscosity, so the above statement is true only approximately. The viscosity is like friction and converts the kinetic energy to heat energy.

9. Though shear strain in a fluid does not require shear stress, when a shear stress is applied to a fluid, the motion is generated which causes a shear strain growing with time. The ratio of the shear stress to the time rate of shearing strain is known as coefficient of viscosity, η .

10. Stokes' law states that the viscous drag force F on a sphere of radius a moving with velocity v through a fluid of viscosity is, $F = 6\pi\eta av$.

11. The onset of turbulence in a fluid is determined by a dimensionless parameter is called the Reynolds number given by
 $R_e = \rho v d / \eta$

where d is a typical geometrical length associated with the fluid flow and the other symbols have their usual meaning.

12. Surface tension is a force per unit length (or surface energy per unit area) acting in the plane of interface between the liquid and the bounding surface. It is the extra energy that the molecules at the interface have as compared to the interior.

Chapter – 11

Thermal Properties of Matter

SUMMARY

- Heat is a form of energy that flows between a body and its surrounding medium by virtue of temperature difference between them. The degree of hotness of the body is quantitatively represented by temperature.
- A temperature-measuring device (thermometer) makes use of some measurable property (called thermometric property) that changes with temperature. Different thermometers lead to different temperature scales. To construct a temperature scale, two fixed points are chosen and assigned some arbitrary values of temperature. The two numbers fix the origin of the scale and the size of its unit.

- The Celsius temperature (t_C) and the Fahrenheit temperature (t_F) are related by

$$t_F = (9/5) t_C + 32$$

- The ideal gas equation connecting pressure (P), volume (V) and absolute temperature (T) is :

$$PV = nRT$$

where n is the number of moles and R is the universal gas constant.

- In the absolute temperature scale, the zero of the scale corresponds to the temperature where every substance in nature has the least possible molecular activity. The Kelvin absolute temperature scale (T) has the same unit size as the Celsius scale (T_C), but differs in the origin:

$$T_C = T - 273.15$$

- The coefficient of linear expansion (α_l) and volume expansion (α_v) are defined by the relations:

$$\frac{\Delta l}{l} = \alpha_l \Delta T$$

$$\frac{\Delta V}{V} = \alpha_v \Delta T$$

where Δl and ΔV denote the change in length l and volume V for a change of temperature ΔT . The relation between them is:

$$\alpha_v = 3\alpha_l$$

7. The specific heat capacity of a substance is defined by

$$s = \frac{1}{m} \frac{\Delta Q}{\Delta T}$$

where m is the mass of the substance and ΔQ is the heat required to change its temperature by ΔT . The molar specific heat capacity of a substance is defined by

$$C = \frac{1}{\mu} \frac{\Delta Q}{\Delta T}$$

where μ is the number of moles of the substance.

8. The latent heat of fusion (L_f) is the heat per unit mass required to change a substance from solid into liquid at the same temperature and pressure. The latent heat of vaporization (L_v) is the heat per unit mass required to change a substance from liquid to the vapor state without change in the temperature and pressure.
9. The three modes of heat transfer are conduction, convection and radiation.
10. In conduction, heat is transferred between neighboring parts of a body through molecular collisions, without any flow of matter. For a bar of length L and uniform cross section A with its ends maintained at temperatures T_C and T_D , the rate of flow of heat H is :

$$H = K A \frac{T_C - T_D}{L}$$

where K is the thermal conductivity of the material of the bar.

11. Newton's Law of Cooling says that the rate of cooling of a body is proportional to the excess temperature of the body over the surroundings:

$$\frac{dQ}{dt} = -k(T_2 - T_1)$$

where T_1 is the temperature of the surrounding medium and T_2 is the temperature of the body.

POINTS TO PONDER

1. The relation connecting Kelvin temperature (T) and the Celsius temperature t_c
- $$T = t_c + 273.15$$

and the assignment $T = 273.16$ K for the triple point of water are exact relations (by choice). With this choice, the Celsius temperature of the melting point of water and boiling point of water (both at 1 atm pressure) are very close to, but not exactly equal to 0°C and 100°C respectively. In the original Celsius scale, these latter fixed points

were exactly at 0°C and 100°C (by choice), but now the triple point of water is the preferred choice for the fixed point, because it has a unique temperature.

2. A liquid in equilibrium with vapor has the same pressure and temperature throughout the system; the two phases in equilibrium differ in their molar volume (i.e. density). This is true for a system with any number of phases in equilibrium.
3. Heat transfer always involves temperature difference between two systems or two parts of the same system. Any energy transfer that does not involve temperature difference in some way is not heat.
4. Convection involves flow of matter *within a fluid* due to unequal temperatures of its parts. A hot bar placed under a running tap loses heat by conduction between the surface of the bar and water and not by convection within water.

Chapter – 12

Thermodynamics

SUMMARY

1. The zeroth law of thermodynamics states that 'two systems in thermal equilibrium with a third system separately are in thermal equilibrium with each other'. The Zeroth Law leads to the concept of temperature.
2. Internal energy of a system is the sum of kinetic energies and potential energies of the molecular constituents of the system. It does not include the over-all kinetic energy of the system. Heat and work are two modes of energy transfer to the system. Heat is the energy transfer arising due to temperature difference between the system and the surroundings. Work is energy transfer brought about by other means, such as moving the piston of a cylinder containing the gas, by raising or lowering some weight connected to it.
3. The first law of thermodynamics is the general law of conservation of energy applied to any system in which energy transfer from or to the surroundings (through heat and work) is taken into account. It states that

$$\Delta Q = \Delta U + \Delta W$$

Where ΔQ is the heat supplied to the system, ΔW is the work done by the system and ΔU is the change in internal energy of the system.

4. The specific heat capacity of a substance is defined by

$$s = \frac{1}{m} \frac{\Delta Q}{\Delta T}$$

where m is the mass of the substance and ΔQ is the heat required to change its temperature by ΔT . The molar specific heat capacity of a substance is defined by

$$C = \frac{1}{\mu} \frac{\Delta Q}{\Delta T}$$

where μ is the number of moles of the substance. For a solid, the law of equipartition of energy gives

$$C = 3R$$

which generally agrees with experiment at ordinary temperatures.

Calorie is the old unit of heat. 1 calorie is the amount of heat required to raise the temperature of 1 g of water from 14.5 °C to 15.5 °C. 1 cal = 4.186 J.

5. For an ideal gas, the molar specific heat capacities at constant pressure and volume satisfy the relation

$$C_p - C_v = R$$

where R is the universal gas constant.

6. Equilibrium states of a thermodynamic system are described by state variables. The value of a state variable depends only on the particular state, not on the path used to arrive at that state. Examples of state variables are pressure (P), volume (V), temperature (T), and mass (m). Heat and work are not state variables. An Equation of State (like the ideal gas equation ($PV = \mu RT$)) is a relation connecting different state variables.

7. A quasi-static process is an infinitely slow process such that the system remains in thermal and mechanical equilibrium with the surroundings throughout. In a quasi-static process, the pressure and temperature of the environment can differ from those of the system only infinitesimally.

8. In an isothermal expansion of an ideal gas from volume V_1 to V_2 at temperature T the heat absorbed (Q) equals the work done (W) by the gas, each given by

$$Q = W = \mu R T \ln \left(\frac{V_2}{V_1} \right)$$

9. In an adiabatic process of an ideal gas

$$PV^\gamma = \text{constant}$$

where $\gamma = \frac{C_p}{C_v}$

Work done by an ideal gas in an adiabatic change of state from (P_1, V_1, T_1) to (P_2, V_2, T_2) is

$$W = \frac{\mu R (T_1 - T_2)}{\gamma - 1}$$

10. Heat engine is a device in which a system undergoes a cyclic process resulting in conversion of heat into work. If Q_1 is the heat absorbed from the source, Q_2 is the heat released to the sink, and the work output in one cycle is W , the efficiency η of the engines:

$$\eta = \frac{W}{Q_1} = 1 - \frac{Q_2}{Q_1}$$

11. In a refrigerator or a heat pump, the system extracts heat Q_2 from the cold reservoir and releases Q_1 amount of heat to the hot reservoir, with work W done on the system. The coefficient of performance of a refrigerator is given by

$$\alpha = \frac{Q_2}{W} = \frac{Q_2}{Q_1 - Q_2}$$

12. The second law of thermodynamics disallows some processes consistent with the First Law of Thermodynamics. It states

Kelvin-Planck statement

No process is possible whose sole result is the absorption of heat from a reservoir and complete conversion of the heat into work.

Clausius statement

No process is possible whose sole result is the transfer of heat from a colder object to a hotter object.

Put simply, the Second Law implies that no heat engine can have efficiency η equal to 1 or no refrigerator can have co-efficient of performance α equal to infinity.

13. A process is reversible if it can be reversed such that both the system and the surroundings return to their original states, with no other change anywhere else in the universe. Spontaneous processes of nature are irreversible. The idealized reversible process is a quasi-static process with no dissipative factors such as friction, viscosity, etc.
14. Carnot engine is a reversible engine operating between two temperatures T_1 (source) and T_2 (sink). The Carnot cycle consists of two isothermal processes connected by two adiabatic processes. The efficiency of a Carnot engine is given by

$$\eta = 1 - \frac{T_2}{T_1} \quad (\text{Carnot engine})$$

No engine operating between two temperatures can have efficiency greater than that of the Carnot engine.

15. If $Q > 0$, heat is added to the system
 If $Q < 0$, heat is removed to the system
 If $W > 0$, Work is done by the system
 If $W < 0$, Work is done on the system

Chapter – 13

Kinetic Theory

SUMMARY

1. The ideal gas equation connecting pressure (P), volume (V) and absolute temperature (T) is

$$PV = \mu RT = k_B NT$$

where μ is the number of moles and N is the number of molecules. R and k_B are universal constants.

$$R = 8.314 \text{ J mol}^{-1} \text{ K}^{-1}, \quad k_B = \frac{R}{N_A} = 1.38 \times 10^{-23} \text{ J K}^{-1}$$

Real gases satisfy the ideal gas equation only approximately, more so at low pressures and high temperatures.

2. Kinetic theory of an ideal gas gives the relation

$$P = \frac{1}{3} n m \overline{v^2}$$

where n is number density of molecules, m the mass of the molecule and $\overline{v^2}$ is the mean of squared speed. Combined with the ideal gas equation it yields a kinetic interpretation of temperature.

$$\frac{1}{2} m \overline{v^2} = \frac{3}{2} k_B T, \quad v_{\text{rms}} = (\overline{v^2})^{1/2} = \sqrt{\frac{3k_B T}{m}}$$

This tells us that the temperature of a gas is a measure of the average kinetic energy of a molecule, independent of the nature of the gas or molecule. In a mixture of gases at a fixed temperature the heavier molecule has the lower average speed.

3. The translational kinetic energy

$$E = \frac{3}{2} k_B NT.$$

This leads to a relation

$$PV = \frac{2}{3} E$$

4. The law of equipartition of energy states that if a system is in equilibrium at absolute temperature, the total energy is distributed equally in different energy modes of absorption, the energy in each mode being equal to $\frac{1}{2} k_B T$. Each translational and rotational degree of freedom corresponds to one energy mode of absorption and has energy $\frac{1}{2} k_B T$. Each vibration frequency has two modes of energy (kinetic and potential) with corresponding energy equal to

$$2 \times \frac{1}{2} k_B T = k_B T.$$

5. Using the law of equipartition of energy, the molar specific heats of gases can be determined and the values are in agreement with the experimental values of specific heats of several gases. The agreement can be improved by including vibration modes of motion.
6. The mean free path l is the average distance covered by a molecule between two successive collisions :

$$l = \frac{1}{\sqrt{2} n \pi d^2}$$

Where n is the number density and d the diameter of the molecule.

Chapter – 14

Oscillations

SUMMARY

- The motion that repeats itself is called periodic motion.
- The period T is the time required for one complete oscillation, or cycle. It is related to the frequency ν by,

$$T = \frac{1}{\nu}$$

The frequency of periodic or oscillatory motion is the number of oscillations per unit time. In the SI, it is measured in hertz :

$$1 \text{ hertz} = 1 \text{ Hz} = 1 \text{ oscillation per second} = 1 \text{ s}^{-1}$$

- In simple harmonic motion (SHM), the displacement $x(t)$ of a particle from its equilibrium position is given by,

$$x(t) = A \cos(\omega t + \phi) \quad (\text{displacement}),$$

in which A is the amplitude of the displacement, the quantity $(\omega t + \phi)$ is the phase of the motion, and ϕ is the *phase constant*. The angular frequency is related to the period and frequency of the motion by,

$$\omega = \frac{2\pi}{T} = 2\pi\nu \quad (\text{angular frequency}).$$

- Simple harmonic motion can also be viewed as the projection of uniform circular motion on the diameter of the circle in which the latter motion occurs.
- The particle velocity and acceleration during SHM as functions of time are given by,

$$v(t) = -\omega A \sin(\omega t + \phi) \quad (\text{Velocity}),$$

$$a(t) = -\omega^2 A \cos(\omega t + \phi)$$

$$= -\omega^2 x(t) \quad (\text{acceleration}),$$

Thus we see that both velocity and acceleration of a body executing simple harmonic motion are periodic functions, having the velocity amplitude $v_m = \omega A$ and acceleration amplitude $a_m = \omega^2 A$, respectively.

- The force acting in a simple harmonic motion is proportional to the displacement and is always directed towards the centre of motion.
- A particle executing simple harmonic motion has, at any time, kinetic energy $K = \frac{1}{2}mv^2$ and potential energy $U = \frac{1}{2}kx^2$. If no friction is present the mechanical energy of the system, $E = K + U$ always remains constant even though K and U change with time.
- A particle of mass m oscillating under the influence of Hooke's law restoring force given by $F = -kx$ exhibits simple harmonic motion with

$$\omega = \sqrt{\frac{k}{m}} \quad (\text{angular frequency})$$

$$T = 2\pi \sqrt{\frac{m}{k}} \quad (\text{period})$$

Such a system is also called a linear oscillator.

- The motion of a simple pendulum swinging through small angles is approximately simple harmonic. The period of oscillation is given by,

$$T = 2\pi \sqrt{\frac{L}{g}}$$

10. The mechanical energy in a real oscillating system decreases during oscillations because external forces, such as drag, inhibit the oscillations and transfer mechanical energy to thermal energy. The real oscillator and its motion are then said to be damped. If the damping forces given by $F_d = -bv$, where v is the velocity of the oscillator and 'b' is a damping constant, then the displacement of the oscillator is given by,

$$x(t) = A e^{-\frac{bt}{2m}} \cos(\omega' t + \phi)$$

where ω' , the angular frequency of the damped oscillator, is given by

$$\omega' = \sqrt{\frac{k}{m} - \frac{b^2}{4m^2}}$$

If the damping constant is small then $b \ll w$, where w is the angular frequency of the un damped oscillator. The mechanical energy E of the damped oscillator is given by

$$E(t) = \frac{1}{2} k A^2 e^{-bt/m}$$

11. If an external force with angular frequency ω_d acts on an oscillating system with natural angular frequency ω , the system oscillates with angular frequency ω_d . The amplitude of oscillations is the greatest when

$$\omega_d = \omega$$

a condition called resonance.

Chapter – 15

Waves

SUMMARY

1. Mechanical waves can exist in material media and are governed by Newton's Laws.
2. Transverse waves are waves in which the particles of the medium oscillate perpendicular to the direction of wave propagation.
3. Longitudinal waves are waves in which the particles of the medium oscillate along the direction of wave propagation.
4. Progressive wave is a wave that moves from one point of medium to another.
5. The displacement in a sinusoidal wave propagating in the positive x direction is given by

$$y(x, t) = a \sin(kx - \omega t + \phi)$$

Where a is the amplitude of the wave, k is the angular wave number, ω is the angular frequency, $(kx - \omega t + \phi)$ is the phase, and ϕ is the phase constant or phase angle.

6. Wavelength of a progressive wave is the distance between two consecutive points of the same phase at a given time. In a stationary wave, it is twice the distance between two consecutive nodes or antinodes.
7. Period T of oscillation of a wave is defined as the time any element of the medium takes to move through one complete oscillation. It is related to the angular frequency through the relation

$$T = \frac{2\pi}{\omega}$$

8. Frequency ' ν ' of a wave is defined as $1/T$ and is related to angular frequency by

$$\nu = \frac{\omega}{2\pi}$$

9. Speed of a progressive wave is given by $v = \frac{\omega}{k} = \frac{\lambda}{T} = \lambda\nu$

10. The speed of a transverse wave on a stretched string is set by the properties of the string. The speed on a string with tension ' T ' and linear mass density μ is

$$v = \sqrt{\frac{T}{\mu}}$$

11. Sound waves are longitudinal mechanical waves that can travel through solids, liquids or gases. The speed ' v ' of sound wave in a fluid having bulk modulus ' B ' and density ' ρ ' is

$$v = \sqrt{\frac{B}{\rho}}$$

The speed of longitudinal waves in a metallic bar is

$$v = \sqrt{\frac{Y}{\rho}}$$

For gases, since $B = \gamma P$, the speed of sound is

$$v = \sqrt{\frac{\gamma P}{\rho}}$$

12. When two or more waves traverse simultaneously in the same medium, the displacement of any element of the medium is the algebraic sum of the displacements due to each wave. This is known as the principle of superposition of waves

$$y = \sum_{i=1}^n f_i(x - vt)$$

13. Two sinusoidal waves on the same string exhibit interference, adding or cancelling according to the principle of superposition. If the two are travelling in the same direction and have the same amplitude a and frequency but differ in phase by a phase constant ϕ , the result is a single wave with the same frequency ω :

$$y(x, t) = \left[2a \cos \frac{1}{2}\phi \right] \sin \left[kx - \omega t + \frac{1}{2}\phi \right]$$

If $\phi = 0$ or an integral multiple of 2π , the waves are exactly in phase and the interference is constructive; if $\phi = \pi$, they are exactly out of phase and the interference is destructive.

14. A travelling wave, at a rigid boundary or a closed end, is reflected with a phase reversal but the reflection at an open boundary takes place without any phase change.
For an incident wave

$$y_i(x, t) = a \sin(kx - \omega t)$$

the reflected wave at a rigid boundary is

$$y_r(x, t) = -a \sin(kx + \omega t)$$

For reflection at an open boundary

$$y_r(x, t) = a \sin(kx + \omega t)$$

15. The interference of two identical waves moving in opposite directions produces standing waves. For a string with fixed ends, the standing wave is given by

$$y(x, t) = [2a \sin kx] \cos \omega t$$

Standing waves are characterized by fixed locations of zero displacement called nodes and fixed locations of maximum displacements called antinodes. The separation between two consecutive nodes or antinodes is $\lambda/2$.

A stretched string of length L fixed at both the ends vibrates with frequencies given by

$$v = \frac{nv}{2L}, \quad n = 1, 2, 3, \dots$$

The set of frequencies given by the above relation are called the normal modes of oscillation of the system. The oscillation mode with lowest frequency is called the fundamental mode or the first harmonic. The second harmonic is the oscillation mode within $n = 2$ and so on.

A pipe of length ' L ' with one end closed and other end open (such as air columns) vibrates with frequencies given by

$$v = \left(n + \frac{1}{2}\right) \frac{v}{2L}, \quad n = 0, 1, 2, 3, \dots$$

The set of frequencies represented by the above relation are the normal modes of oscillation of such a system. The lowest frequency given by $v/4L$ is the fundamental mode or the first harmonic.

16. A string of length 'L' fixed at both ends or an air column closed at one end and open at the other end or open at both the ends, vibrates with certain frequencies called their normal modes. Each of these frequencies is a resonant frequency of the system.
17. Beats arise when two waves having slightly different frequencies, v_1 and v_2 and comparable amplitudes, are superposed. The beat frequency is

$$v_{beat} = v_1 \sim v_2$$

18. The Doppler effect is a change in the observed frequency of a wave when the source (S) or the observer (O) or both move(s) relative to the medium. For sound the observed frequency 'v' is given in terms of the source frequency v_0 by

$$v = v_0 \left[\frac{v+v_0}{v+v_s} \right]$$

here v is the speed of sound through the medium, v_0 is the velocity of observer relative to the medium, and v_s is the source velocity relative to the medium. In using this formula, velocities in the direction OS should be treated as positive and those opposite to it should be taken to be negative.

Note: Please refer to the text book for any clarifications

NCERT Physics XII, Part 1

CHAPTER – 1

(Adopted for educational social service purpose only)

Electric Charges and Fields

SUMMARY

1. Electric and magnetic forces determine the properties of atoms, molecules and bulk matter.
2. From simple experiments on frictional electricity, one can infer that there are two types of charges in nature; and that like charges repel and unlike charges attract. By convention, the charge on a glass rod rubbed with silk is positive; that on a plastic rod rubbed with fur is then negative.
3. Conductors allow movement of electric charge through them, insulators do not. In metals, the mobile charges are electrons; in electrolytes both positive and negative ions are mobile.
4. Electric charge has three basic properties: quantisation, additivity and conservation.

Quantisation of electric charge means that total charge (q) of a body is always an integral multiple of a basic quantum of charge (e) i.e. $q = n e$, where $n = 0, \pm 1, \pm 2, \pm 3$. Proton and electron have charges $+e, -e$, respectively. For macroscopic charges for which n is a very large number, quantisation of charge can be ignored.

Additivity of electric charges means that the total charge of a system is the algebraic sum (i.e., the sum taking into account proper signs) of all individual charges in the system.

Conservation of electric charges means that the total charge of an isolated system remains unchanged with time. This means that when bodies are charged through friction, there is a transfer of electric charge from one body to another, but no creation or destruction of charge.

5. **Coulomb's Law:** The mutual electrostatic force between two point charges q_1 and q_2 is proportional to the product $q_1 q_2$ and inversely proportional to the square of the distance r_{21} separating them. Mathematically,

$$F_{21} = \text{force on } q_2 \text{ due to } q_1 = \frac{k(q_1 q_2)}{r_{21}^2} \hat{r}_{21}$$

where \hat{r}_{21} is a unit vector in the direction from q_1 to q_2 and $k = \frac{1}{4\pi\epsilon_0}$ is the constant of proportionality.

In SI units, the unit of charge is coulomb. The experimental value of the constant ϵ_0 is

$$\epsilon_0 = 8.854 \times 10^{-12} \text{ C}^2 \text{ N}^{-1} \text{ m}^{-2}$$

The approximate value of k is

$$k = 9 \times 10^9 \text{ N m}^2 \text{ C}^{-2}$$

6. The ratio of electric force and gravitational force between a proton and an electron is

$$\frac{ke^2}{G m_e m_p} \cong 2.4 \times 10^{39}$$

7. **Superposition Principle:** The principle is based on the property that the forces with which two charges attract or repel each other are not affected by the presence of a third (or more) additional charge(s). For an assembly of charges q_1, q_2, q_3, \dots the force on any charge, say q_1 , is the vector sum of the force on q_1 due to q_2 , the force on q_1 due to q_3 , and so on. For each pair, the force is given by the Coulomb's law for two charges stated earlier.
8. The electric field E at a point due to a charge configuration is the force on a small positive test charge ' q ' placed at the point divided by the magnitude of the charge. Electric field due to a point charge q has a magnitude $|q|/4\pi\epsilon_0 r^2$; it is radially outwards from q , if q is positive and radially inwards if q is negative. Like Coulomb force, electric field also satisfies superposition principle.
9. An electric field line is a curve drawn in such a way that the tangent at each point on the curve gives the direction of electric field at that point. The relative closeness of field lines indicates the relative strength of electric field at different points; they crowd near each other in regions of strong electric field and are far apart where the electric field is weak. In regions of constant electric field, the field lines are uniformly spaced parallel straight lines.
10. Some of the important properties of field lines are:
 (i) Field lines are continuous curves without any breaks.
 (ii) Two field lines cannot cross each other.
 (iii) Electrostatic field lines start at positive charges and end at negative charges they cannot form closed loops.
11. An electric dipole is a pair of equal and opposite charges q and $-q$ separated by some distance $2a$. Its dipole moment vector ' p ' has magnitude $2qa$ and is in the direction of the dipole axis from $-q$ to q .
12. Field of an electric dipole in its equatorial plane (i.e., the plane perpendicular to its axis and passing through its centre) at a distance ' r ' from the centre:

$$E = \frac{-P}{4\pi\epsilon_0} \frac{1}{(a^2+r^2)^{3/2}}$$

$$\cong \frac{-P}{4\pi\epsilon_0 r^3}, \text{ for } r \gg a$$

Dipole electric field on the axis at a distance ' r ' from the centre:

$$E = \frac{2Pr}{4\pi\epsilon_0(r^2 - a^2)^2}$$

$$\cong \frac{2P}{4\pi\epsilon_0 r^3}, \text{ for } r \gg a$$

The $1/r^3$ dependence of dipole electric fields should be noted in contrast to the $1/r^2$ dependence of electric field due to a point charge.

13. In a uniform electric field E , a dipole experiences a torque τ given by

$$\tau = p \times E$$
 but experiences no net force.
14. The flux $\Delta \phi$ of electric field 'E' through a small area element ΔS is given by

$$\Delta \phi = E \cdot \Delta S$$

The vector area element ΔS is

$$\Delta S = \Delta S \hat{n}$$

where ΔS is the magnitude of the area element and \hat{n} is normal to the area element, which can be considered planar for sufficiently small ΔS .

For an area element of a closed surface, \hat{n} is taken to be the direction of outward normal, by convention.

15. **Gauss's law:** The flux of electric field through any closed surface 'S' is $1/\epsilon_0$ times the total charge enclosed by 'S'. The law is especially useful for determining electric field 'E', when the source distribution has simple symmetry:
- (i) Thin infinitely long straight wire of uniform linear charge density λ

$$E = \frac{\lambda}{2\pi\epsilon_0 r} \hat{n}$$

Where 'r' is the perpendicular distance of the point from the wire and \hat{n} is the radial unit vector in the plane normal to the wire passing through the point.

- (ii) Infinite thin plane sheet of uniform surface charge density σ

$$E = \frac{\sigma}{2\epsilon_0} \hat{n}$$

Where \hat{n} is a unit vector normal to the plane, outward on either side.

- (iii) Thin spherical shell of uniform surface charge density σ

$$E = \frac{q}{4\pi\epsilon_0 r^2} \hat{r} \quad (r \geq R)$$

$$E = 0 \quad (r < R)$$

Where 'r' is the distance of the point from the centre of the shell and 'R' the radius of the shell, 'q' is the total charge of the shell: $q = 4\pi R^2 \sigma$.

The electric field outside the shell is as though the total charge is concentrated at the centre. The same result is true for a solid sphere of uniform volume charge density. The field is zero at all points inside the shell.

Chapter – 2

Electrostatic Potential and Capacitance

SUMMARY

1. Electrostatic force is a conservative force. Work done by an external force (equal and opposite to the electrostatic force) in bringing a charge 'q' from a point 'R' to a point 'P' is $V_P - V_R$, which is the difference in potential energy of charge 'q' between the final and initial points.
2. Potential at a point is the work done per unit charge (by an external agency) in bringing a charge from infinity to that point. Potential at a point is arbitrary to within an additive constant, since it is the potential difference between two points which is physically significant. If potential at infinity is chosen to be zero; potential at a point with position vector 'r' due to a point charge 'Q' placed at the origin is given by

$$V(r) = \frac{1}{4\pi\epsilon_0} \frac{Q}{r}$$

3. The electrostatic potential at a point with position vector \mathbf{r} due to a point dipole of dipole moment \mathbf{p} placed at the origin is

$$V(r) = \frac{1}{4\pi\epsilon_0} \frac{\mathbf{p} \cdot \hat{\mathbf{r}}}{r^2}$$

The result is true also for a dipole (with charges $-q$ and q separated by $2a$) for $r \gg a$.

4. For a charge configuration q_1, q_2, \dots, q_n with position vectors $\mathbf{r}_1, \mathbf{r}_2, \dots, \mathbf{r}_n$, the potential at a point P is given by the superposition principle

$$V = \frac{1}{4\pi\epsilon_0} \left(\frac{q_1}{r_{1P}} + \frac{q_2}{r_{2P}} + \dots + \frac{q_n}{r_{nP}} \right)$$

Where r_{1P} is the distance between q_1 and P, as and so on.

5. An equipotential surface is a surface over which potential has a constant value. For a point charge, concentric spheres centered at a location of the charge are equipotential surfaces. The electric field \mathbf{E} at a point is perpendicular to the equipotential surface through the point. \mathbf{E} is in the direction of the steepest decrease of potential.
6. Potential energy stored in a system of charges is the work done (by an external agency) in assembling the charges at their locations. Potential energy of two charges q_1, q_2 at r_1, r_2 is given by

$$U = \frac{1}{4\pi\epsilon_0} \frac{q_1 q_2}{r_{12}}$$

Where r_{12} is distance between q_1 and q_2 .

7. The potential energy of a charge 'q' in an external potential $V(r)$ is $qV(r)$. The potential energy of a dipole moment \mathbf{p} in a uniform electric field \mathbf{E} is $-\mathbf{p}\cdot\mathbf{E}$.
8. Electrostatics field \mathbf{E} is zero in the interior of a conductor; just outside the surface of a charged conductor, \mathbf{E} is normal to the surface given by

$$\mathbf{E} = \frac{\sigma}{\epsilon_0} \hat{n}$$
 where \hat{n} is the unit vector along the outward normal to the surface and σ is the surface charge density. Charges in a conductor can reside only at its surface. Potential is constant within and on the surface of a conductor. In a cavity within a conductor (with no charges), the electric field is zero.

9. A capacitor is a system of two conductors separated by an insulator. Its capacitance is defined by $C = Q/V$, where Q and $-Q$ are the charges on the two conductors and V is the potential difference between them. C is determined purely geometrically, by the shapes, sizes and relative positions of the two conductors. The unit of capacitance is farad:

$1 \text{ F} = 1 \text{ C V}^{-1}$. For a parallel plate capacitor (with vacuum between the plates),

$$C = \epsilon_0 \frac{A}{d}$$

Where 'A' is the area of each plate and 'd' the separation between them.

10. If the medium between the plates of a capacitor is filled with an insulating substance (dielectric), the electric field due to the charged plates induces a net dipole moment in the dielectric. This effect, called polarisation, gives rise to a field in the opposite direction. The net electric field inside the dielectric and hence the potential difference between the plates is thus reduced. Consequently, the capacitance 'C' increases from its value C_0 when there is no medium (vacuum),

$$C = KC_0$$

Where K is the dielectric constant of the insulating substance.

11. For capacitors in the series combination, the total capacitance C is given by

$$\frac{1}{C} = \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} + \dots$$

In the parallel combination, the total capacitance C is:

$$C = C_1 + C_2 + C_3 + \dots$$

Where C_1, C_2, C_3, \dots are individual capacitances.

12. The energy U stored in a capacitor of capacitance C , with charge Q and voltage V is

$$U = \frac{1}{2}QV = \frac{1}{2}CV^2 = \frac{1}{2}\frac{Q^2}{C}$$

The electric energy density (energy per unit volume) in a region with electric field is $(1/2)\epsilon_0 E^2$.

13. A Van de Graff generator consists of a large spherical conducting shell (a few metre in diameter). By means of a moving belt and suitable brushes, charge is continuously transferred to the shell and potential difference of the order of several million volts is built up, which can be used for accelerating charged particles.

Chapter – 3

Current Electricity

SUMMARY

- Current through a given area of a conductor is the net charge passing per unit time through the area.
- To maintain a steady current, we must have a closed circuit in which an external agency moves electric charge from lower to higher potential energy. The work done per unit charge by the source in taking the charge from lower to higher potential energy (i.e., from one terminal of the source to the other) is called the electromotive force, or emf, of the source. Note that the emf is not a force; it is the voltage difference between the two terminals of a source in open circuit.

3. Ohm's law: The electric current I flowing through a substance is proportional to the voltage V across its ends, i.e., $V \propto I$ or $V = RI$, where ' R ' is called the resistance of the substance. The unit of resistance is ohm: $1\Omega = 1 \text{ V A}^{-1}$.
4. The resistance R of a conductor depends on its length ' l ' and constant cross-sectional area A through the relation,

$$R = \frac{\rho l}{A}$$

Where ρ , called resistivity is a property of the material and depends on temperature and pressure.

5. Electrical resistivity of substances varies over a very wide range: Metals have low resistivity, in the range of $10^{-8}\Omega \text{ m}$ to $10^{-6}\Omega \text{ m}$. Insulators like glass and rubber have 10^{22} to 10^{24} times greater resistivity. Semiconductors like Si and Ge lie roughly in the middle range of resistivity on a logarithmic scale.
6. In most substances, the carriers of current are electrons; in some cases, for example, ionic crystals and electrolytic liquids, positive and negative ions carry the electric current.
7. Current density \mathbf{j} gives the amount of charge flowing per second per unit area normal to the flow,

$$\mathbf{j} = nq \mathbf{v}_d$$

Where ' n ' is the number density (number per unit volume) of charge carriers each of charge q , and \mathbf{v}_d is the drift velocity of the charge carriers. For electrons $q = -e$. If \mathbf{j} is normal to a cross-sectional area A and is constant over the area, the magnitude of the current ' I ' through the area is $nev_d A$.

8. Using $E = V/l$, $I = nev_d A$, and Ohm's law, one obtains

$$\frac{eE}{m} = \rho \frac{ne^2}{m} v_d$$

The proportionality between the force E on the electrons in a metal due to the external field ' E ' and the drift velocity v_d (not acceleration) can be understood, if we assume that the electrons suffer collisions with ions in the metal, which deflect them randomly. If such collisions occur on an average at a time interval τ ,

$$v_d = a\tau = eE\tau/m$$

Where a is the acceleration of the electron. This gives

$$\rho = \frac{m}{ne^2\tau}$$

9. In the temperature range in which resistivity increases linearly with temperature, the temperature coefficient of resistivity ' α ' is defined as the fractional increase in resistivity per unit increase in temperature.
10. Ohm's law is obeyed by many substances, but it is not a fundamental law of nature. It fails if
- 'V' depends on 'I' non-linearly.
 - The relation between 'V' and 'I' depends on the sign of V for the same absolute value of 'V'.
 - The relation between 'V' and 'I' is non-unique. An example of (a) is when ' ρ ' increases with 'I' (even if temperature is kept fixed). A rectifier combines features (a) and (b).

11. When a source of emf ε is connected to an external resistance R, the voltage V_{ext} across R is given by

$$V_{\text{ext}} = IR = \frac{\varepsilon}{R+r} R$$

Where 'r' is the internal resistance of the source.

12. (a) Total resistance 'R' of 'n' resistors connected in series is given by
 $R = R_1 + R_2 + \dots + R_n$
- (b) Total resistance 'R' of 'n' resistors connected in parallel is given by
 $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \dots + \frac{1}{R_n}$
13. Kirchhoff's Rules –
- Junction Rule:** At any junction of circuit elements, the sum of currents entering the junction must equal the sum of currents leaving it.
 - Loop Rule:** The algebraic sum of changes in potential around any closed loop must be zero.
14. The Wheatstone bridge is an arrangement of four resistances – R_1, R_2, R_3, R_4 as shown in the text. The null-point condition is given by

$$\frac{R_1}{R_2} = \frac{R_3}{R_4}$$

using which the value of one resistance can be determined, knowing the other three resistances.

15. The potentiometers a device to compare potential differences. Since the method involves a condition of no current flow, the device can be used to measure potential difference; internal resistance of a cell and compare emf's of two sources.

POINTS TO PONDER

1. Current is a scalar although we represent current with an arrow. Currents do not obey the law of vector addition. That current is scalar also follows from its definition. The current 'I' through an area of cross-section is given by the scalar product of two vectors:

$$I = \mathbf{j} \cdot \Delta \mathbf{S}$$

Where \mathbf{j} and $\Delta \mathbf{S}$ are vectors.

2. Refer to V-I curves of a resistor and a diode as drawn in the text. A resistor obeys Ohm's law while a diode does not. The assertion that $V = IR$ is a statement of Ohm's law is not true. This equation defines resistance and it may be applied to all conducting devices whether they obey Ohm's law or not. The Ohm's law asserts that the plot of 'I' versus 'V' is linear i.e., R is independent of V.

Equation $E = \rho j$ leads to another statement of Ohm's law, i.e., a conducting material obeys Ohm's law when the resistivity of the material does not depend on the magnitude and direction of applied electric field.

3. Homogeneous conductors like silver or semiconductors like pure germanium or germanium containing impurities obey Ohm's law within some range of electric field values. If the field becomes too strong, there are departures from Ohm's law in all cases.
4. Motion of conduction electrons in electric field 'E' is the sum of (i) motion due to random collisions and (ii) that due to E. The motion due to random collisions averages to zero and does not contribute to v_d (Chapter 11, Textbook of Class XI). v_d , thus is only due to applied electric field on the electron.
5. The relation $\mathbf{j} = \rho \mathbf{v}$ should be applied to each type of charge carriers separately. In a conducting wire, the total current and charge density arises from both positive and negative charges:

$$\mathbf{j} = \rho_+ \mathbf{v}_+ + \rho_- \mathbf{v}_-$$

$$\rho = \rho_+ + \rho_-$$

Now in a neutral wire carrying electric current,

$$\rho_+ = -\rho_-$$

Further, $v_+ \sim 0$ which gives

$$\rho = 0$$

$$\mathbf{j} = \rho_- \mathbf{v}_-$$

Thus, the relation $j = \rho v$ does not apply to the total current charge density.

6. Kirchhoff's junction rule is based on conservation of charge and the outgoing currents add up and are equal to incoming current at a junction. Bending or reorienting the wire does not change the validity of Kirchhoff's junction rule.

Chapter – 4

Moving Charges and Magnetism

SUMMARY

1. The total force on a charge 'q' moving with velocity 'v' in the presence of magnetic and electric fields B and E respectively is called the Lorentz force. It is given by the expression:

$$F = q (v \times B + E)$$

The magnetic force $q (v \times B)$ is normal to 'v' and work done by it is zero.

2. A straight conductor of length 'l' and carrying a steady current 'I' experiences a force 'F' in a uniform external magnetic field 'B',

$$F = I l \times B$$

where $|l| = l$ and the direction of l is given by the direction of the current.

3. In a uniform magnetic field B, a charge 'q' executes a circular orbit in a plane normal to B. Its frequency of uniform circular motion is called the cyclotron frequency and is given by:

$$v_c = \frac{qB}{2\pi m}$$

This frequency is independent of the particle's speed and radius. This fact is exploited in a machine, the cyclotron, which is used to accelerate charged particles.

4. The Biot-Savart law asserts that the magnetic field dB due to an element dl carrying a steady current 'I' at a point 'P' at a distance 'r' from the current element is:

$$dB = \frac{\mu_0}{4\pi} I \frac{dl \times r}{r^3}$$

To obtain the total field at 'P', we must integrate this vector expression over the entire length of the conductor.

5. The magnitude of the magnetic field due to a circular coil of radius 'R' carrying a current 'I' at an axial distance 'x' from the centre is

$$B = \frac{\mu_0 I R^2}{2(x^2 + R^2)^{3/2}}$$

At the centre this reduces to

$$B = \frac{\mu_0 I}{2R}$$

6. **Ampere's Circuital Law:** Let an open surface 'S' be bounded by a loop 'C'. Then the Ampere's law states that $\oint_C \mathbf{B} \cdot d\mathbf{l} = \mu_0 I$ where 'I' refers to the current passing through 'S'. The sign of 'I' is determined from the right-hand rule. If 'B' is directed along the tangent to every point on the perimeter 'L' of a closed curve and is constant in magnitude along perimeter then,

$$BL = \mu_0 I_e$$

where I_e is the net current enclosed by the closed circuit.

7. The magnitude of the magnetic field at a distance R from a long, straight wire carrying a current I is given by:

$$B = \frac{\mu_0 I}{2\pi R}$$

The field lines are circles concentric with the wire.

8. The magnitude of the field 'B' inside a long solenoid carrying a current 'I' is

$$B = \mu_0 nI$$

Where 'n' is the number of turns per unit length. For a toroid one obtains,

$$B = \frac{\mu_0 NI}{2\pi r}$$

Where 'N' is the total number of turns and 'r' is the average radius.

9. Parallel currents attract and anti-parallel currents repel.
10. A planar loop carrying a current 'I', having 'N' closely wound turns, and an area 'A' possesses a magnetic moment 'm' where,
 $m = N I A$

and the direction of 'm' is given by the right-hand thumb rule : curl the palm of your right hand along the loop with the fingers pointing in the direction of the current. The thumb sticking out gives the direction of m (and A).

When this loop is placed in a uniform magnetic field B, the force F on it is: $F = 0$
And the torque on it is,

$$\tau = m \times B$$

In a moving coil galvanometer, this torque is balanced by a counter torque due to a spring, yielding

$$K \phi = NI AB$$

where ϕ is the equilibrium deflection and 'k' the torsion constant of the spring.

11. An electron moving around the central nucleus has a magnetic moment μ_l given by:

$$\mu_l = \frac{e}{2m} l$$

where 'l' is the magnitude of the angular momentum of the circulating electron about the central nucleus. The smallest value of μ_l is called the Bohr magneton μ_B and it is $\mu_B = 9.27 \times 10^{-24} \text{ J/T}$

12. A moving coil galvanometer can be converted into an ammeter by introducing a shunt resistance r_s , of small value in parallel. It can be converted into a voltmeter by introducing a resistance of a large value in series.

POINTS TO PONDER

- Electrostatic field lines originate at a positive charge and terminate at a negative charge or fade at infinity. Magnetic field lines always form closed loops.
- The discussion in this Chapter holds only for steady currents which do not vary with time.
When currents vary with time Newton's third law is valid only if momentum carried by the electromagnetic field is taken into account.
- Recall the expression for the Lorentz force,
 $F = q(\mathbf{v} \times \mathbf{B} + \mathbf{E})$

This velocity dependent force has occupied the attention of some of the greatest scientific thinkers. If one switches to a frame with instantaneous velocity 'v', the magnetic part of the force vanishes. The motion of the charged particle is then

explained by arguing that there exists an appropriate electric field in the new frame. We shall not discuss the details of this mechanism. However, we stress that the resolution of this paradox implies that electricity and magnetism are linked phenomena (electromagnetism) and that the Lorentz force expression does not imply a universal preferred frame of reference in nature.

4. Ampere's Circuital law is not independent of the Biot-Savart law. It can be derived from the Biot-Savart law. Its relationship to the Biot-Savart law is similar to the relationship between Gauss's law and Coulomb's law.

Chapter – 5

Magnetism and Matter

SUMMARY

- The science of magnetism is old. It has been known since ancient times that magnetic materials tend to point in the north-south direction; like magnetic poles repel and unlike ones attract; and cutting a bar magnetic two leads to two smaller magnets. Magnetic poles cannot be isolated.
- When a bar magnet of dipole moment 'm' is placed in a uniform magnetic field 'B',
 - The force on it is zero,
 - The torque on it is $m \times B$,
 - Its potential energy is $-m \cdot B$, where we choose the zero of energy at the orientation when m is perpendicular to B.
- Consider a bar magnet of size 'l' and magnetic moment 'm', at a distance 'r' from its mid-point, where $r \gg l$, the magnetic field 'B' due to this bar,

$$B = \frac{\mu_0 m}{2\pi r^3} \quad (\text{along axis})$$

$$= -\frac{\mu_0 m}{4\pi r^3} \quad (\text{along equator})$$

4. Gauss's law for magnetism states that the net magnetic flux through any closed surface is zero

$$\phi_B = \sum_{\text{elements } \Delta S} B \cdot \Delta S = 0$$

5. The earth's magnetic field resembles that of a (hypothetical) magnetic dipole located at the centre of the earth. The pole near the geographic north pole of the earth is called the north magnetic pole. Similarly, the pole near the geographic South Pole is called the south magnetic pole. This dipole is aligned making a small angle with the rotation axis of the earth. The magnitude of the field on the earth's surface $\approx 4 \times 10^{-5}$ T.

6. Three quantities are needed to specify the magnetic field of the earth on its surface – the horizontal component, the magnetic declination, and the magnetic dip. These are known as the elements of the earth's magnetic field.
7. Consider a material placed in an external magnetic field B_0 . The magnetic intensity is defined as,

$$H = \frac{B_0}{\mu_0}$$

The magnetisation M of the material is its dipole moment per unit volume. The magnetic field B in the material is,

$$B = \mu_0 (H + M)$$

8. For a linear material $M = \chi H$. So that $B = \mu H$ and is called the magnetic susceptibility of the material. The three quantities, χ , the relative magnetic permeability μ_r , and the magnetic permeability μ are related as follows:

$$\mu = \mu_0 \mu_r$$

$$\mu_r = 1 + \chi$$

9. Magnetic materials are broadly classified as: diamagnetic, paramagnetic, and ferromagnetic. For diamagnetic materials χ is negative and small and for paramagnetic materials it is positive and small. Ferromagnetic materials have large ' χ ' and are characterised by non-linear relation between B and H . They show the property of hysteresis.
10. Substances, which at room temperature retain their ferromagnetic property for a long period of time are called permanent magnets.

POINTS TO PONDER

1. A satisfactory understanding of magnetic phenomenon in terms of moving charges/currents was arrived at after 1800 AD. But technological exploitation of the directional properties of magnets predates this scientific understanding by two thousand years. Thus, scientific understanding is not a necessary condition for engineering applications. Ideally, science and engineering go hand-in-hand, one leading and assisting the other in tandem.

2. Magnetic monopoles do not exist. If you slice a magnet in half, you get two smaller magnets. On the other hand, isolated positive and negative charges exist. There exists a smallest unit of charge, for example, the electronic charge with value $|e| = 1.6 \times 10^{-19}$ C. All other charges are integral multiples of this smallest unit charge. In other words, charge is quantised. We do not know why magnetic monopoles do not exist or why electric charge is quantised.
3. A consequence of the fact that magnetic monopoles do not exist is that the magnetic field lines are continuous and form closed loops. In contrast, the electrostatic lines of force begin on a positive charge and terminate on the negative charge (or fade out at infinity).
4. The earth's magnetic field is not due to a huge bar magnet inside it. The earth's core is hot and molten. Perhaps convective currents in this core are responsible for the earth's magnetic field. As to what 'dynamo' effect sustains this current, and why the earth's field reverses polarity every million years or so, we do not know.
5. A miniscule difference in the value of, the magnetic susceptibility, yields radically different behaviour: diamagnetic versus paramagnetic. For diamagnetic materials $\chi = -10^{-5}$ whereas $\chi = +10^{-5}$ for paramagnetic materials.
6. There exists a perfect diamante, namely, a superconductor. This is a metal at very low temperatures. In this case, $\chi = -1$, $\mu_r = 0$, $\mu = 0$. The external magnetic field is totally expelled. Interestingly, this material is also a perfect conductor. However, there exists no classical theory which ties these two properties together. A quantum-mechanical theory by Bardeen, Cooper, and Schrieffer (BCS theory) explains these effects. The BCS theory was proposed in 1957 and was eventually recognised by a Nobel Prize in physics in 1970.
7. The phenomenon of magnetic hysteresis is reminiscent of similar behaviour concerning the elastic properties of materials. Strain may not be proportional to stress; here B (or M) are not linearly related. The stress-strain curve exhibits hysteresis and area enclosed by it represents the energy dissipated per unit volume. A similar interpretation can be given to the B - H magnetic hysteresis curve.
8. Diamagnetism is universal. It is present in all materials. But it is weak and hard to detect if the substance is para- or ferromagnetic.
9. We have classified materials as diamagnetic, paramagnetic, and ferromagnetic. However, there exist additional types of magnetic material such as ferric magnetic, anti-ferromagnetic, spin glass, etc. with properties which are exotic and mysterious.

Chapter – 6

Electromagnetic Induction

SUMMARY

- The magnetic flux through a surface of area 'A' placed in a uniform magnetic field 'B' is defined as,

$$\Phi_B = B \cdot A = BA \cos \theta$$

Where 'θ' is the angle between B and A.

- Faraday's laws of induction imply that the emf induced in a coil of 'N' turns is directly related to the rate of change of flux through it,

$$\varepsilon = -N \frac{d\Phi_B}{dt}$$

Here Φ_B is the flux linked with one turn of the coil. If the circuit is closed, a current $I = \varepsilon/R$ is set up in it, where 'R' is the resistance of the circuit.

- Lenz's law states that the polarity of the induced emf is such that it tends to produce a current which opposes the change in magnetic flux that produces it. The negative sign in the expression for Faraday's law indicates this fact.

- When a metal rod of length 'l' is placed normal to a uniform magnetic field B and moved with a velocity 'v' perpendicular to the field, the induced emf (called motional emf) across its ends is

$$\varepsilon = Blv$$

- Changing magnetic fields can set up current loops in nearby metal (any conductor) bodies. They dissipate electrical energy as heat. Such currents are called eddy currents.

- Inductance is the ratio of the flux-linkage to current. It is equal to $N\Phi/I$.

- A changing current in a coil (coil 2) can induce an emf in a nearby coil (coil 1). This relation is given by,

$$\varepsilon_1 = -M_{12} \frac{di_2}{dt}$$

The quantity M_{12} is called mutual inductance of coil 1 with respect to coil 2. One can similarly define M_{21} . There exists a general equality,

$$M_{12} = M_{21}$$

8. When a current in a coil changes, it induces a back emf in the same coil. The self-induced emf is given by,

$$\varepsilon = -L \frac{di}{dt}$$

L is the self-inductance of the coil. It is a measure of the inertia of the coil against the change of current through it.

9. The self-inductance of a long solenoid, the core of which consists of a magnetic material of permeability μ_r , is given by

$$L = \mu_r \mu_0 n^2 Al$$

Where ' A ' is the area of cross-section of the solenoid, ' l ' its length and n the number of turns per unit length.

10. In an ac generator, mechanical energy is converted to electrical energy by virtue of electromagnetic induction. If coil of ' N ' turn and area ' A ' is rotated at ' v ' revolutions per second in a uniform magnetic field ' B ', then the motional emf produced is

$$\varepsilon = NBA (2\pi v) \sin (2\pi vt)$$

where we have assumed that at time $t = 0$ s, the coil is perpendicular to the field.

Chapter - 7

Alternating Current

If the secondary coil has a greater number of turns than the primary ($N_s > N_p$), the voltage is stepped up ($V_s > V_p$). This type of arrangement is called a step-up transformer. However, in this arrangement, there is less current in the secondary than in the primary ($N_p/N_s < 1$ and $I_s < I_p$). For example, if the primary coil of a transformer has 100 turns and the secondary has 200 turns, $N_s/N_p = 2$ and $N_p/N_s = 1/2$. Thus, a 220V input at 10A will step-up to 440 V output at 5.0 A.

If the secondary coil has less turns than the primary ($N_s < N_p$), we have a step-down transformer. In this case, $V_s < V_p$ and $I_s > I_p$. That is, the voltage is stepped down, or reduced, and the current is increased. The equations obtained above apply to ideal transformers (without any energy losses). But in actual transformers, small energy losses do occur due to the following reasons:

(i) **Flux Leakage:** There is always some flux leakage; that is, not all of the flux due to primary passes through the secondary due to poor design of the core or the air gaps in the core. It can be reduced by winding the primary and secondary coils one over the other.

(ii) **Resistance of the windings:** The wire used for the windings has some resistance and so, energy is lost due to heat produced in the wire (I^2R). In high current, low voltage windings, these are minimised by using thick wire.

(iii) **Eddy currents:** The alternating magnetic flux induces eddy currents in the iron core and causes heating. The effect is reduced by having a laminated core.

(iv) **Hysteresis:** The magnetisation of the core is repeatedly reversed by the alternating magnetic field. The resulting expenditure of energy in the core appears as heat and is kept to a minimum by using a magnetic material which has a low hysteresis loss.

The large scale transmission and distribution of electrical energy over long distances is done with the use of transformers. The voltage output of the generator is stepped-up (so that current is reduced and consequently, the I^2R loss is cut down). It is then transmitted over long distances to an area sub-station near the consumers. There the voltage is stepped down. It is further stepped down at distributing sub-stations and utility poles before a power supply of 240 V reaches our homes.

SUMMARY

1. An alternating voltage $v = v_m \sin \omega t$ applied to a resistor 'R' drives a current $i = i_m \sin \omega t$ in the resistor, $i_m = \frac{v_m}{R}$. The current is in phase with the applied voltage.
2. For an alternating current $i = i_m \sin \omega t$ passing through a resistor 'R', the average power loss P (averaged over a cycle) due to joule heating is $(1/2)i_m^2 R$. To express it in the same form as the dc power ($P = I^2 R$), a special value of current is used. It is called root mean square (rms) current and is denoted by I:

$$I = \frac{i_m}{\sqrt{2}} = 0.707 i_m$$

Similarly, the rms voltage is defined by

$$V = \frac{v_m}{\sqrt{2}} = 0.707 v_m$$

We have $P = IV = I^2 R$

3. An ac voltage $v = v_m \sin \omega t$ applied to a pure inductor L, drives a current in the inductor $i = i_m \sin(\omega t - \frac{\pi}{2})$, where $i_m = v_m / X_L$. $X_L = \omega L$ is called inductive reactance. The current in the inductor lags the voltage by $\pi/2$. The average power supplied to an inductor over one complete cycle is zero.
4. An ac voltage $v = v_m \sin \omega t$ applied to a capacitor drives a current in the capacitor: $i = i_m \sin(\omega t + \frac{\pi}{2})$. Here, $i_m = \frac{v_m}{X_C}$, $X_C = \frac{1}{\omega C}$ is called capacitive reactance.

The current through the capacitor is $\pi/2$ ahead of the applied voltage. As in the case of inductor, the average power supplied to a capacitor over one complete cycle is zero.

5. For a series RLC circuit driven by voltage $v = v_m \sin \omega t$, the current is given by $i = i_m \sin(\omega t + \phi)$

where
$$i_m = \frac{v_m}{\sqrt{R^2 + (X_C - X_L)^2}}$$

and
$$\phi = \tan^{-1} \frac{X_C - X_L}{R}$$

$Z = \sqrt{R^2 + (X_C - X_L)^2}$ is called the impedance of the circuit.

The average power loss over a complete cycle is given by

$$P = VI \cos \phi$$

The term $\cos \phi$ is called the power factor.

6. In a purely inductive or capacitive circuit, $\cos \phi = 0$ and no power is dissipated even though a current is flowing in the circuit. In such cases, current is referred to as a wattless current.
7. The phase relationship between current and voltage in an a circuit can be shown conveniently by representing voltage and current by rotating vectors called phases. A phase is a vector which rotates about the origin with angular speed ' ω '. The magnitude of a phase or represents the amplitude or peak value of the quantity (voltage or current) represented by the pHS or.

The analysis of an ac circuit is facilitated by the use of a phase diagram.

8. An interesting characteristic of a series RLC circuit is the phenomenon of resonance. The circuit exhibits resonance, i.e., the amplitude of the current is maximum at the resonant frequency, $\omega_0 = \frac{1}{\sqrt{LC}}$. The quality factor 'Q' defined by

$Q = \frac{\omega_0 L}{R} = \frac{1}{\omega_0 CR}$ is an indicator of the sharpness of the resonance, the higher value of 'Q' indicating sharper peak in the current.

9. A circuit containing an inductor L and a capacitor C (initially charged) with no ac source and no resistors exhibits free oscillations. The charge 'q' of the capacitor satisfies the equation of simple harmonic motion:

$$\frac{d^2q}{dt^2} + \frac{1}{LC}q = 0$$

and therefore, the frequency ' ω ' of free oscillation is $\omega_0 = \frac{1}{\sqrt{LC}}$. The energy in the system oscillates between the capacitor and the inductor but their sum or the total energy is constant in time.

10. A transformer consists of an iron core on which are bound a primary coil of N_p turns and a secondary coil of N_s turns. If the primary coil is connected to an ac source, the primary and secondary voltages are related by

$$V_s = \left(\frac{N_s}{N_p}\right) V_p$$

and the currents are related by

$$I_s = \left(\frac{N_p}{N_s}\right) I_p$$

If the secondary coil has a greater number of turns than the primary, the voltage is stepped-up ($V_s > V_p$). This type of arrangement is called a step-up transformer. If the secondary coil has turns less than the primary, we have a step-down transformer.

POINTS TO PONDER

1. When a value is given for ac voltage or current, it is ordinarily the rms value. The voltage across the terminals of an outlet in your room is normally 240 V. This refers to the *rms* value of the voltage. The amplitude of this voltage is

$$v_m = \sqrt{2}V = \sqrt{2}(240) = 340 \text{ V.}$$

2. The power rating of an element used in ac circuits refers to its average power rating.
3. The power consumed in a circuit is never negative.
4. Both alternating current and direct current are measured in amperes. But how is the ampere defined for an alternating current? It cannot be derived from the mutual attraction of two parallel wires carrying currents, as the dc ampere is derived. An ac current changes direction with the source frequency and the attractive force would average to zero. Thus, the ac ampere must be defined in terms of some property that is independent of the direction of the current. Joule heating is such a property, and there is one ampere of rms value of alternating current in a circuit if the current produces the same average heating effect as one ampere of dc current would produce under the same conditions.
5. In an ac circuit, while adding voltages across different elements, one should take care of their phases properly. For example, if V_R and V_C are voltages across R and C, respectively in an RC circuit, then the total voltage across RC combination is $V_{RC} = \sqrt{V_R^2 + V_C^2}$ and not $V_R + V_C$ since V_C is $\pi/2$ out of phase of V_R .

6. Though in a phase diagram, voltage and current are represented by vectors, these quantities are not really vectors themselves. They are scalar quantities. It so happens that the amplitudes and phases of harmonically varying scalars combine mathematically in the same way as do the projections of rotating vectors of corresponding magnitudes and directions. The 'rotating vectors' that represent harmonically varying scalar quantities are introduced only to provide us with a simple way of adding these quantities using a rule that we already know as the law of vector addition.
7. There are no power losses associated with pure capacitances and pure inductances in an ac circuit. The only element that dissipates energy in an ac circuit is the resistive element.
8. In a RLC circuit, resonance phenomenon occur when $X_L = X_C$ or $\omega_0 = \frac{1}{\sqrt{LC}}$. For resonance to occur, the presence of both 'L' and 'C' elements in the circuit is a must. With only one of these (L or C) elements, there is no possibility of voltage cancellation and hence, no resonance is possible.
9. The power factor in a RLC circuit is a measure of how close the circuit is to expending the maximum power.
10. In generators and motors, the roles of input and output are reversed. In a motor, electric energy is the input and mechanical energy is the output. In a generator, mechanical energy is the input and electric energy is the output. Both devices simply transform energy from one form to another.
11. A transformer (step-up) changes a low-voltage into a high-voltage. This does not violate the law of conservation of energy. The current is reduced by the same proportion.
12. The choice of whether the description of an oscillatory motion is by means of sines or cosines or by their linear combinations is unimportant, since changing the zero-time position transforms the one to the other.

Chapter – 8

Electromagnetic Waves

SUMMARY

1. Maxwell found an inconsistency in the Ampere's law and suggested the existence of an additional current, called displacement current, to remove his inconsistency. This displacement current is due to time-varying electric field and is given by

$$i_d = \epsilon_0 \frac{d\phi_E}{dt}$$

and acts as a source of magnetic field in exactly the same way as conduction current.

2. An accelerating charge produces electromagnetic waves. An electric charge oscillating harmonically with frequency 'ν', produces electromagnetic waves of the same frequency 'ν'. An electric dipole is a basic source of electromagnetic waves.
3. Electromagnetic waves with wavelength of the order of a few metres were first produced and detected in the laboratory by Hertz in 1887. He thus verified a basic prediction of Maxwell's equations.
4. Electric and magnetic fields oscillate sinusoidal in space and time in an electromagnetic wave. The oscillating electric and magnetic fields, 'E' and 'B' are perpendicular to each other, and to the direction of propagation of the electromagnetic wave. For a wave of frequency 'ν', wavelength 'λ', propagating along z-direction, we have

$$E = E_x(t) = E_0 \sin(kz - \omega t)$$

$$= E_0 \sin \left[2\pi \left(\frac{z}{\lambda} - \nu t \right) \right] = E_0 \sin \left[\left(2\pi \left(\frac{z}{\lambda} - \frac{t}{T} \right) \right) \right]$$

$$B = B_y(t) = B_0 \sin(kz - \omega t)$$

$$= B_0 \sin \left[2\pi \left(\frac{z}{\lambda} - \nu t \right) \right] = B_0 \sin \left[\left(2\pi \left(\frac{z}{\lambda} - \frac{t}{T} \right) \right) \right]$$

They are related by $E_0/B_0 = c$.

5. The speed 'c' of electromagnetic wave in vacuum is related to μ_0 and ϵ_0 (the free space permeability and permittivity constants) as follows:

$c = 1 / \sqrt{\mu_0 \epsilon_0}$. The value of 'c' equals the speed of light obtained from optical measurements.

Light is an electromagnetic wave; c is, therefore, also the speed of light. Electromagnetic waves other than light also have the same velocity 'c' in free space.

The speed of light, or of electromagnetic waves in a material medium is given by $v = 1 / \sqrt{\mu \epsilon}$

where 'μ' is the permeability of the medium and 'ε' its permittivity.

6. Electromagnetic waves carry energy as they travel through space and this energy is shared equally by the electric and magnetic fields.

Electromagnetic waves transport momentum as well. When these waves strike a surface, a pressure is exerted on the surface. If total energy transferred to a surface in time 't' is 'U', total momentum delivered to this surface is $p = U/c$.

7. The spectrum of electromagnetic waves stretches, in principle, over an infinite range of wavelengths. Different regions are known by different names; γ -rays, X-rays, ultraviolet rays, visible rays, infrared rays, microwaves and radio waves in order of increasing wavelength from 10^{-2} Å or 10^{-12} m to 10^6 m.

They interact with matter via their electric and magnetic fields which set in oscillation charges present in all matter. The detailed interaction and so the mechanism of absorption, scattering, etc., depend on the wavelength of the electromagnetic wave, and the nature of the atoms and molecule in the medium.

POINTS TO PONDER

1. The basic difference between various types of electromagnetic waves lies in their wavelengths or frequencies since all of them travel through vacuum with the same speed. Consequently, the waves differ considerably in their mode of interaction with matter.
2. Accelerated charged particles radiate electromagnetic waves. The wavelength of the electromagnetic wave is often correlated with the characteristic size of the system that radiates. Thus, gamma radiation, having wavelength of 10^{-14} m to 10^{-15} m, typically originate from an atomic nucleus. X-rays are emitted from heavy atoms. Radio waves are produced by accelerating electrons in a circuit. A transmitting antenna can most efficiently radiate waves having a wavelength of about the same size as the antenna. Visible radiation emitted by atoms is, however, much longer in wavelength than atomic size.
3. The oscillating fields of an electromagnetic wave can accelerate charges and can produce oscillating currents. Therefore, an apparatus designed to detect electromagnetic waves is based on this fact. Hertz original 'receiver' worked in exactly this way. The same basic principle is utilised in practically all modern receiving devices. High frequency electromagnetic waves are detected by other means based on the physical effects they produce on interacting with matter.
4. Infrared waves, with frequencies lower than those of visible light, vibrate not only the electrons, but entire atoms or molecules of a substance. This vibration increases the internal energy and consequently, the temperature of the substance. This is why infrared waves are often called heat waves.
5. The centre of sensitivity of our eyes coincides with the centre of the wavelength distribution of the sun. It is because humans have evolved with visions most sensitive to the strongest wavelengths from the sun.

Note: Please refer to the text book for any clarifications

NCERT PHYSICS XII – PART II

Chapter –9

(Adopted for educational social service purpose only)**Ray Optics and Optical Instruments****SUMMARY**

- Reflection is governed by the equation $\angle i = \angle r'$ and refraction by the Snell's law, $\sin i / \sin r = n$, where the incident ray, reflected ray, refracted ray and normal lie in the same plane. Angles of incidence, reflection and refraction are i , r' and r , respectively.
- The critical angle of incidence i_c for a ray incident from a denser to rarer medium, is that angle for which the angle of refraction is 90° . For $i > i_c$, total internal reflection occurs. Multiple internal reflections in diamond ($i_c \cong 24.4^\circ$), totally reflecting prisms and mirage, are some examples of total internal reflection. Optical fibres consist of glass fibres coated with a thin layer of material of lower refractive index. Light incident at an angle at one end comes out at the other, after multiple internal reflections, even if the fibre is bent.
- Cartesian sign convention:** Distances measured in the same direction as the incident light are positive; those measured in the opposite direction are negative. All distances are measured from the pole/opticcentre of the mirror/lens on the principal axis. The heights measured upwards above x -axis and normal to the principal axis of the mirror/lens are taken as positive. The heights measured downwards are taken as negative.
- Mirror equation:**

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

Where u and v are object and image distances, respectively and f is the focal length of the mirror. F is (approximately) half the radius of curvature R . f is negative for concave mirror; f is positive for a convex mirror.
- For a prism of the angle A , of refractive index n_2 placed in a medium of refractive index n_1

$$n_{21} = \frac{n_2}{n_1} = \frac{\sin[(A + D_m)/2]}{\sin(A/2)}$$

where D_m is the angle of minimum deviation.
- For refraction through a spherical interface (from medium 1 to 2 of refractive index n_1 and n_2 , respectively)

$$\frac{n_2}{v} - \frac{n_1}{u} = \frac{n_2 - n_1}{R}$$

Thin lens formula

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{f}$$

Lens maker's formula

$$\frac{1}{f} = \frac{(n_2 - n_1)}{n_1} \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$$

R_1 and R_2 are the radii of curvature of the lens surfaces. f is positive for a converging lens; f is negative for a diverging lens. The power of a lens $P = 1/f$.

The SI unit for power of a lens is dioptre (D): $1 \text{ D} = 1 \text{ m}^{-1}$.

If several thin lenses of focal length f_1, f_2, f_3, \dots are in contact, the effective focal length of their combination, is given by

$$\frac{1}{f} = \frac{1}{f_1} + \frac{1}{f_2} + \frac{1}{f_3} + \dots$$

The total power of a combination of several lenses is $P = P_1 + P_2 + P_3 + \dots$

7. Dispersion is the splitting of light into its constituent colours.
8. **The Eye:** The eye has a convex lens of focal length about 2.5 cm. This focal length can be varied somewhat so that the image is always formed on the retina. This ability of the eye is called accommodation. In a defective eye, if the image is focussed before the retina (myopia), a diverging corrective lens is needed; if the image is focussed beyond the retina (hypermetropia), a converging corrective lens is needed.

Astigmatism is corrected by using cylindrical lenses.

9. Magnifying power m of a simple microscope is given by $m = 1 + (D/f)$, where $D = 25 \text{ cm}$ is the least distance of distinct vision and f is the focal length of the convex lens. If the image is at infinity, $m = D/f$. For a compound microscope, the magnifying power is given by $m = m_e \times m_o$ where $m_e = 1 + (D/f_e)$, is the magnification due to the eyepiece and m_o is the magnification produced by the objective.

Approximately,

$$m = \frac{L}{f_o} \times \frac{D}{f_e}$$

Where f_o and f_e are the focal lengths of the objective and eyepiece, respectively, and L is the distance between their focal points.

10. Magnifying power m of a telescope is the ratio of the angle β subtended at the eye by the image to the angle α subtended at the eye by the object.

$$m = \frac{\beta}{\alpha} = \frac{f_o}{f_e}$$

Where f_o and f_e are the focal lengths of the objective and eyepiece, respectively.

Chapter –10

Wave Optics

SUMMARY

1. Huygens' principle tells us that each point on a wavefront is a source of secondary waves, which add up to give the wavefront at a later time.
2. Huygens' construction tells us that the new wavefront is the forward envelope of the secondary waves. When the speed of light is independent of direction, the secondary waves are spherical. The rays are then perpendicular to both the wavefronts and the time of travel is the same measured along any ray. This principle leads to the well known laws of reflection and refraction.
3. The principle of superposition of waves applies whenever two or more sources of light illuminate the same point. When we consider the intensity of light due to these sources at the given point, there is an interference term in addition to the sum of the individual intensities. But this term is important only if it has a non-zero average, which occurs only if the sources have the same frequency and a stable phase difference.
4. Young's double slit of separation d gives equally spaced fringes of angular separation λ/d . The source, mid-point of the slits, and central bright fringe lie in a straight line. An extended source will destroy the fringes if it subtends angle more than λ/d at the slits.
5. A single slit of width a gives a diffraction pattern with a central maximum. The intensity falls to zero at angles of $\pm \frac{\lambda}{a}, \pm \frac{2\lambda}{a}$, etc..with successively weaker secondary maxima in between. Diffraction limits the angular resolution of a telescope to λ/D where D is the diameter. Two stars closer than this give strongly overlapping images. Similarly, a microscope objective subtending angle 2β at the focus, in a medium of refractive index n , will just separate two objects spaced at a distance $\lambda/(2n \sin \beta)$, which is the resolution limit of a microscope. Diffraction determines the limitations of the concept of light rays. A beam of width a travels a distance a^2/λ , called the Fresnel distance, before it starts to spread out due to diffraction.
6. Natural light, e.g., from the sun is unpolarised. This means the electric vector takes all possible directions in the transverse plane, rapidly and randomly, during a measurement. A polaroid transmits only one component (parallel to a special axis). The resulting light is called linearly polarised or plane polarised. When this kind of light is viewed through a

second polaroid whose axis turns through 2π , two maxima and minima of intensity are seen. Polarised light can also be produced by reflection at a special angle (called the Brewster angle) and by scattering through $\pi/2$ in the earth's atmosphere.

Chapter –11

Dual Nature of Radiation and Matter

SUMMARY

1. The minimum energy needed by an electron to come out from a metal surface is called the work function of the metal. Energy (greater than the work function (ϕ_0)) required for electron emission from the metal surface can be supplied by suitably heating or applying strong electric field or irradiating it by light of suitable frequency.
2. Photoelectric effect is the phenomenon of emission of electrons by metals when illuminated by light of suitable frequency. Certain metals respond to ultraviolet light while others are sensitive even to the visible light. Photoelectric effect involves conversion of light energy into electrical energy. It follows the law of conservation of energy. The photoelectric emission is an instantaneous process and possesses certain special features.
3. Photoelectric current depends on (i) the intensity of incident light, (ii) the potential difference applied between the two electrodes, and (iii) the nature of the emitter material.
4. The stopping potential (V_0) depends on (i) the frequency of incident light, and (ii) the nature of the emitter material. For a given frequency of incident light, it is independent of its intensity. The stopping potential is directly related to the maximum kinetic energy of electrons emitted:

$$e V_0 = (1/2) m v_{max}^2 = K_{max}.$$

5. Below a certain frequency (threshold frequency) V_0 , characteristic of the metal, no photoelectric emission takes place, no matter how large the intensity may be.
6. The classical wave theory could not explain the main features of photoelectric effect. Its picture of continuous absorption of energy from radiation could not explain the independence of K_{max} on intensity, the existence of V_0 and the instantaneous nature of the process. Einstein explained these features on the basis of photon picture of light. According to this, light is composed of discrete packets of energy called quanta or photons. Each photon carries an energy $E (= h\nu)$ and momentum $p (= h/\lambda)$, which depend on the frequency (ν) of incident light and not on its intensity. Photoelectric emission from the metal surface occurs due to absorption of a photon by an electron.

7. Einstein's photoelectric equation is in accordance with the energy conservation law as applied to the photon absorption by an electron in the metal. The maximum kinetic energy $(1/2)mv_{max}^2$ is equal to the photon energy ($h\nu$) minus the work function ϕ_0 ($=h\nu_0$) of the target metal:

$$\frac{1}{2}mv_{max}^2 = h\nu - \phi_0 = h(\nu - \nu_0)$$

This photoelectric equation explains all the features of the photoelectric effect. Millikan's first precise measurements confirmed the Einstein's photoelectric equation and obtained an accurate value of Planck's constant h . This led to the acceptance of particle or photon description(nature) of electromagnetic radiation, introduced by Einstein.

8. Radiation has dual nature: wave and particle. The nature of experiment determines whether a wave or particle description is best suited for understanding the experimental result. Reasoning that radiation and matter should be symmetrical in nature, Louis Victor de Broglie attributed a wave-like character to matter (material particles). The waves associated with the moving material particles are called matter waves or de Broglie waves.
9. The de Broglie wavelength (λ) associated with a moving particle is related to its momentum p as: $\lambda = h/p$. The dualism of matter is inherent in the de Broglie relation which contains a wave concept (λ) and a particle concept(p). The de Broglie wavelength is independent of the charge and nature of the material particle. It is significantly measurable (of the order of the atomic-planes spacing in crystals) only in case of sub-atomic particles like electrons, protons, etc. (due to smallness of their masses and hence, momenta). However, it is indeed very small, quite beyond measurement, in case of macroscopic objects, commonly encountered in everyday life.
10. Electron diffraction experiments by Davisson and Germer, and by G. P. Thomson, as well as many later experiments, have verified and confirmed the wave-nature of electrons. The de Broglie hypothesis of matter waves supports the Bohr's concept of stationary orbits.

Chapter – 12

Atoms

SUMMARY

1. Atom, as a whole, is electrically neutral and therefore contains equal amount of positive and negative charges.
2. In Thomson's model, an atom is a spherical cloud of positive charges with electrons embedded in it.
3. In Rutherford's model, most of the mass of the atom and all its positive charge are concentrated in a tiny nucleus (typically one by ten thousand the size of an atom), and the electrons revolve around it.
4. Rutherford nuclear model has two main difficulties in explaining the structure of atom: (a) It predicts that atoms are unstable because the accelerated electrons revolving around the nucleus must spiral into the nucleus. This contradicts the stability of matter. (b) It cannot explain the characteristic line spectra of atoms of different elements.
5. Atoms of each element are stable and emit characteristic spectrum. The spectrum consists of a set of isolated parallel lines termed as line spectrum. It provides useful information about the atomic structure.
6. The atomic hydrogen emits a line spectrum consisting of various series. The frequency of any line in a series can be expressed as a difference of two terms;

Lyman series: $\nu = Rc \left(\frac{1}{1^2} - \frac{1}{n^2} \right); n = 2, 3, 4, \dots$

Balmer series: $\nu = Rc \left(\frac{1}{2^2} - \frac{1}{n^2} \right); n = 3, 4, 5, \dots$

Paschen series: $\nu = Rc \left(\frac{1}{3^2} - \frac{1}{n^2} \right); n = 4, 5, 6, \dots$

Brackett series: $\nu = Rc \left(\frac{1}{4^2} - \frac{1}{n^2} \right); n = 5, 6, 7, \dots$

Pfund series: $\nu = Rc \left(\frac{1}{5^2} - \frac{1}{n^2} \right); n = 6, 7, 8, \dots$
7. To explain the line spectra emitted by atoms, as well as the stability of atoms, Neil's Bohr proposed a model for hydrogenic (single electron) atoms. He introduced three postulates and laid the foundations of quantum mechanics:

- (a) In a hydrogen atom, an electron revolves in certain stable orbits (called stationary orbits) without the emission of radiant energy.
- (b) The stationary orbits are those for which the angular momentum is some integral multiple of $h/2\pi$. (Bohr's quantisation condition.) That is $L = nh/2\pi$, where n is an integer called a quantum number.
- (c) The third postulate states that an electron might make a transition from one of its specified non-radiating orbits to another of lower energy. When it does so, a photon is emitted having energy equal to the energy difference between the initial and final states. The frequency (ν) of the emitted photon is then given by
- $$h\nu = E_i - E_f$$

An atom absorbs radiation of the same frequency the atom emits, in which case the electron is transferred to an orbit with a higher value of n .

$$E_i + h\nu = E_f$$

8. As a result of the quantisation condition of angular momentum, the electron orbits the nucleus at only specific radii. For a hydrogen atom it is given by

$$r_n = \left(\frac{n^2}{m}\right) \left(\frac{h}{2\pi}\right)^2 \frac{4\pi\epsilon_0}{e^2}$$

The total energy is also quantised:

$$E_n = -\frac{me^4}{8n^2\epsilon_0^2h^2}$$

$$= -13.6 \text{ eV} / n^2$$

The $n = 1$ state is called ground state. In hydrogen atom the ground state energy is -13.6 eV. Higher values of n correspond to excited states ($n > 1$). Atoms are excited to these higher states by collisions with other atoms or electrons or by absorption of a photon of right frequency.

9. de Broglie's hypothesis that electrons have a wavelength $\lambda = h/mv$ gave an explanation for Bohr's quantised orbits by bringing in the wave-particle duality. The orbits correspond to circular standing waves in which the circumference of the orbit equals a whole number of wavelengths.
10. Bohr's model is applicable only to hydrogenic (single electron) atoms. It cannot be extended to even two electron atoms such as helium. This model is also unable to explain for the relative intensities of the frequencies emitted even by hydrogenic atoms.

Chapter –13

Nuclei

SUMMARY

1. An atom has a nucleus. The nucleus is positively charged. The radius of the nucleus is smaller than the radius of an atom by a factor of 10^4 . More than 99.9% mass of the atom is concentrated in the nucleus.
2. On the atomic scale, mass is measured in atomic mass units (u). By definition, 1 atomic mass unit (1u) is $1/12^{\text{th}}$ mass of one atom of ^{12}C ; $1\text{u} = 1.660563 \times 10^{-27} \text{ kg}$.
3. A nucleus contains a neutral particle called neutron. Its mass is almost the same as that of proton.
4. The atomic number Z is the number of protons in the atomic nucleus of an element. The mass number A is the total number of protons and neutrons in the atomic nucleus; $A = Z+N$; Here N denotes the number of neutrons in the nucleus.

A nuclear species or a nuclide is represented as ^A_ZX , where X is the chemical symbol of the species. Nuclides with the same atomic number Z , but different neutron number N are called isotopes. Nuclides with the same A are isobars and those with the same N are isotones.

Most elements are mixtures of two or more isotopes. The atomic mass of an element is a weighted average of the masses of its isotopes. The masses are the relative abundances of the isotopes.

5. A nucleus can be considered to be spherical in shape and assigned a radius. Electron scattering experiments allow determination of the nuclear radius; it is found that radii of nuclei fit the formula

$$R = R_0 A^{1/3},$$

Where $R_0 = \text{a constant} = 1.2 \text{ fm}$. This implies that the nuclear density is independent of A . It is of the order of 10^{17} kg/m^3 .

6. Neutrons and protons are bound in a nucleus by the short-range strong nuclear force. The nuclear force does not distinguish between neutron and proton.
7. The nuclear mass M is always less than the total mass, Σm , of its constituents. The difference in mass of a nucleus and its constituents is called the mass defect,

$$\Delta M = \{Z m_p + (A - Z) m_n\} - M$$

Using Einstein's mass energy relation, we express this mass difference in terms of energy as

$$\Delta E_b = \Delta M c^2$$

The energy ΔE_b represents the *binding energy* of the nucleus. In the mass number range $A = 30$ to 170, the binding energy per nucleon is nearly constant, about 8 MeV/nucleon.

8. Energies associated with nuclear processes are about a million times larger than chemical process.

9. The Q -value of a nuclear process is

$Q =$ final kinetic energy – initial kinetic energy.

Due to conservation of mass-energy, this is also,

$$Q = (\text{sum of initial masses} - \text{sum of final masses}) c^2$$

10. Radioactivity is the phenomenon in which nuclei of a given species transform by giving out α or β rays or γ rays; α -rays are helium nuclei; β -rays are electrons. γ -rays are electromagnetic radiation of wavelengths shorter than X -rays.

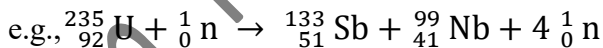
11. Law of radioactive decay : $N(t) = N(0) e^{-\lambda t}$

where λ is the decay constant or disintegration constant.

The half-life $T_{1/2}$ of a radionuclide is the time in which N has been reduced to one-half of its initial value. The mean life τ is the time at which N has been reduced to e^{-1} of its initial value

$$T_{1/2} = \frac{\ln 2}{\lambda} = \tau \ln 2$$

12. Energy is released when less tightly bound nuclei are transmuted into more tightly bound nuclei. In fission, a heavy nucleus like ${}_{92}^{235}\text{U}$ breaks into two smaller fragments,



13. The fact that more neutrons are produced in fission than are consumed gives the possibility of a chain reaction with each neutron that is produced triggering fission. The chain reaction is uncontrolled and rapid in a nuclear bomb explosion. It is controlled and steady in a nuclear reactor. In a reactor, the value of the neutron multiplication factor k is maintained at 1.

14. In fusion, lighter nuclei combine to form a larger nucleus. Fusion of hydrogen nuclei into helium nuclei is the source of energy of all stars including our sun.

Chapter – 14

Semiconductor Electronics: Materials, Devices and Simple circuits

SUMMARY

1. Semiconductors are the basic materials used in the present solid state electronic devices like diode, transistor, ICs, etc.
2. Lattice structure and the atomic structure of constituent elements decide whether a particular material will be insulator, metal or semiconductor.
3. Metals have low resistivity (10^{-2} to $10^{-8}\Omega\text{m}$), insulators have very high resistivity ($>10^8 \Omega\text{m}$), while semiconductors have intermediate values of resistivity.
4. Semiconductors are elemental (Si, Ge) as well as compound (GaAs, CdS, etc.).
5. Pure semiconductors are called 'intrinsic semiconductors'. The presence of charge carriers (electrons and holes) is an 'intrinsic' property of the material and these are obtained as a result of thermal excitation. The number of electrons (n_e) is equal to the number of holes (n_h) in intrinsic semiconductors. Holes are essentially electron vacancies with an effective positive charge.
6. The number of charge carriers can be changed by 'doping' of a suitable impurity in pure semiconductors. Such semiconductors are known as extrinsic semiconductors. These are of two types (n-type and p-type).
7. In n-type semiconductors, $n_e \gg n_h$ while in p-type semiconductors $n_h \gg n_e$.
8. n-type semiconducting Si or Ge is obtained by doping with pentavalent atoms (donors) like As, Sb, P, etc., while p-type Si or Ge can be obtained by doping with trivalent atom (acceptors) like B, Al, In etc.
9. $n_e n_h = n_i^2$ in all cases. Further, the material possesses an overall charge neutrality.
10. There are two distinct bands of energies (called valence band and conduction band) in which the electrons in a material lie. Valence band energies are low as compared to conduction band energies. All energy levels in the valence band are filled while energy levels in the conduction band may be fully empty or partially filled. The electrons in the conduction band are free to move in a solid and are responsible for the conductivity. The extent of conductivity depends upon the energy gap (E_g) between the top of valence band (E_V) and the bottom of the conduction band E_C . The electrons from valence band can be excited by heat, light or electrical energy to the conduction band and thus, produce a change in the current flowing in a semiconductor.
11. For insulators $E_g > 3 \text{ eV}$, for semiconductors E_g is 0.2 eV to 3 eV, while for metals $E_g \approx 0$.

12. p-n junction is the 'key' to all semiconductor devices. When such a junction is made, a 'depletion layer' is formed consisting of immobile ion-cores devoid of their electrons or holes. This is responsible for a junction potential barrier.
13. By changing the external applied voltage, junction barriers can be changed. In forward bias (n-side is connected to negative terminal of the battery and p-side is connected to the positive), the barrier is decreased while the barrier increases in reverse bias. Hence, forward bias current is more (mA) while it is very small (μA) in a p-n junction diode.
14. Diodes can be used for rectifying an ac voltage (restricting the ac voltage to one direction). With the help of a capacitor or a suitable filter, a dc voltage can be obtained.
15. There are some special purpose diodes.
16. Zener diode is one such special purpose diode. In reverse bias, after a certain voltage, the current suddenly increases (breakdown voltage) in a Zener diode. This property has been used to obtain voltage regulation.
17. p-n junctions have also been used to obtain many photonic or optoelectronic devices where one of the participating entity is 'photon':
 - (a) Photodiodes in which photon excitation results in a change of reverse saturation current which helps us to measure light intensity;
 - (b) Solar cells which convert photon energy into electricity;
 - (c) Light Emitting Diode and Diode Laser in which electron excitation by a bias Voltage results in the generation of light.
18. Transistor is an n-p-n or p-n-p junction device. The central block (thin and lightly doped) is called 'Base' while the other electrodes are 'Emitter' and 'Collectors'. The emitter-base junction is forward biased while collector-base junction is reverse biased.
19. The transistors can be connected in such a manner that either C or E or B is common to both the input and output. This gives the three configurations in which a transistor is used: Common Emitter (CE), Common Collector (CC) and Common Base (CB). The plot between I_C and V_{CE} for fixed I_B is called output characteristics while the plot between I_B and V_{BE} with fixed V_{CE} is called input characteristics. The important transistor parameters for CE-configuration are:

$$\text{input resistance, } r_i = \left(\frac{\Delta V_{BE}}{\Delta I_B} \right)_{V_{CE}}$$

$$\text{output resistance, } r_o = \left(\frac{\Delta V_{CE}}{\Delta I_C} \right)_{I_B}$$

current amplification factor, $\beta = \left(\frac{\Delta I_C}{\Delta I_B} \right)_{V_{CE}}$

20. Transistor can be used as an amplifier and oscillator. In fact, an oscillator can also be considered as a self-sustained amplifier in which a part of output is fed-back to the input in the same phase (positive feedback). The voltage gain of a transistor amplifier in common emitter configuration is: $A_v = \left(\frac{v_o}{v_i} \right) = \beta \frac{R_C}{R_B}$, where R_C and R_B are respectively the resistances in collector and base sides of the circuit.
21. When the transistor is used in the cut off or saturation state, it acts as a switch.
22. There are some special circuits which handle the digital data consisting of 0 and 1 levels. This forms the subject of Digital Electronics.
23. The important digital circuits performing special logic operations are called logic gates. These are: OR, AND, NOT, NAND, and NOR gates.
24. In modern day circuit, many logical gates or circuits are integrated in one single 'Chip'. These are known as Integrated circuits (IC).

Chapter-15

Communication Systems

SUMMARY

1. Electronic communication refers to the faithful transfer of information or message (available in the form of electrical voltage and current) from one point to another point.
2. Transmitter, transmission channel and receiver are three basic units of a communication system.
3. Two important forms of communication system are: Analog and Digital. The information to be transmitted is generally in continuous waveform for the former while for the latter it has only discrete or quantised levels.
4. Every message signal occupies a range of frequencies. The bandwidth of a message signal refers to the band of frequencies, which are necessary for satisfactory transmission of the information contained in the signal. Similarly, any practical communication system permits transmission of a range of frequencies only, which is referred to as the bandwidth of the system.
5. Low frequencies cannot be transmitted to long distances. Therefore, they are superimposed on a high frequency carrier signal by a process known as modulation.

6. In modulation, some characteristic of the carrier signal like amplitude, frequency or phase varies in accordance with the modulating or message signal. Correspondingly, they are called Amplitude Modulated (AM), Frequency Modulated (FM) or Phase Modulated (PM) waves.
7. Pulse modulation could be classified as: Pulse Amplitude Modulation(PAM), Pulse Duration Modulation (PDM) or Pulse Width Modulation(PWM) and Pulse Position Modulation (PPM).
8. For transmission over long distances, signals are radiated into space using devices called antennas. The radiated signals propagate as electromagnetic waves and the mode of propagation is influenced by the presence of the earth and its atmosphere. Near the surface of the earth, electromagnetic waves propagate as surface waves. Surface wave propagation is useful up to a few MHz frequencies.
9. Long distance communication between two points on the earth is achieved through reflection of electromagnetic waves by ionosphere. Such waves are called sky waves. Sky wave propagation takes place upto frequency of about 30 MHz. Above this frequency, electromagnetic waves essentially propagate as space waves. Space waves are used for line-of-sight communication and satellite communication.
10. If an antenna radiates electromagnetic waves from a height h_T , then the range d_T is given by $\sqrt{2Rh_T}$ where R is the radius of the earth.
11. Amplitude modulated signal contains frequencies $(\omega_c - \omega_m)$, ω_c and $(\omega_c + \omega_m)$.
12. Amplitude modulated waves can be produced by application of the message signal and the carrier wave to a non-linear device, followed by a band pass filter.
13. AM detection, which is the process of recovering the modulating signal from an AM waveform, is carried out using a rectifier and an envelope detector.