NEET UG 2023 G5 Question Paper with Solutions

Time Allowed: 3 Hour 20 Minutes | Maximum Marks: 720 | Total Questions: 200

General Instructions

Read the following instructions very carefully and strictly follow them:

- 1. The Answer Sheet is this Test Booklet. When you are directed to open the Test Booklet, take the Answer Sheet and fill in the particulars in ORIGINAL Copy carefully with blue/black ball pen only.
- 2. The test is of 3 hours 20 minutes duration and the Test Booklet contains 200 multiple-choice questions (four options with a single correct answer) from Physics, Chemistry, and Biology (Botany and Zoology). 50 questions in each subject are divided into two Sections (A and B) as per details given below:
- 3. (a) Section A shall consist of 35 (Thirty-five) questions in each subject (Question Nos. 1 to 35, 51 to 85, 101 to 135 and 151 to 185).
- 4. (b) Section B shall consist of 15 (Fifteen) questions in each subject (Question Nos. 36 to 50, 86 to 100, 136 to 150 and 186 to 200). In Section B, a candidate needs to attempt any 10 (Ten) questions out of 15 (Fifteen) in each subject.
- 5. Candidates are advised to read all 15 questions in each subject of Section B before they start attempting the question paper. In the event of a candidate attempting more than ten questions, the first ten questions answered by the candidate shall be evaluated.
- 6. Each question carries 4 marks. For each correct response, the candidate will get 4 marks. For each incorrect response, one mark will be deducted from the total scores. The maximum marks are 720.
- 7. Rough work is to be done in the space provided for this purpose in the Test Booklet only.
- 8. On completion of the test, the candidate must hand over the Answer Sheet (ORIGINAL and OFFICE Copy) to the Invigilator before leaving the Room/Hall. The candidates are allowed to take away this Test Booklet with them.
- 9. Use of Electronic/Manual Calculator is prohibited.

Physics

- 1. The magnetic energy stored in an inductor of inductance 4 μH carrying a current of 2 A is :
- (A) 8 µJ
- (B) 4 µJ

- (C) 4 mJ
- (D) 8 mJ

Correct Answer: (A) 8 µJ

Solution:

Step 1: Understanding the Question:

The question asks for the magnetic potential energy stored in an inductor given its inductance and the current flowing through it.

Step 2: Key Formula or Approach:

The energy (U) stored in an inductor is given by the formula:

$$U = \frac{1}{2}LI^2$$

where L is the inductance and I is the current.

Step 3: Detailed Explanation:

We are given the following values:

Inductance, $L = 4 \mu H = 4 \times 10^{-6} H$.

Current, I = 2 A.

Substituting these values into the formula:

$$U = \frac{1}{2} \times (4 \times 10^{-6} \text{ H}) \times (2 \text{ A})^{2}$$

$$U = \frac{1}{2} \times 4 \times 10^{-6} \times 4$$

$$U = 2 \times 10^{-6} \times 4$$

$$U = 8 \times 10^{-6} \text{ J}$$

Since $1 \mu J = 10^{-6} J$, the energy stored is $8 \mu J$.

Step 4: Final Answer:

The magnetic energy stored in the inductor is 8 µJ. This corresponds to option (A).

Quick Tip

Always pay close attention to the prefixes of the units, such as micro (μ) which is 10^{-6} and milli (m) which is 10^{-3} . A small mistake in the power of 10 can lead to selecting the wrong option.

2. The half life of a radioactive substance is 20 minutes. In how much time, the activity of substance drops to $\left(\frac{1}{16}\right)^{th}$ of its initial value?

2

- (A) 80 minutes
- (B) 20 minutes
- (C) 40 minutes
- (D) 60 minutes

Correct Answer: (A) 80 minutes

Solution:

Step 1: Understanding the Question:

We are given the half-life of a radioactive substance and asked to find the time it takes for its activity to decrease to a specific fraction (1/16) of its original activity.

Step 2: Key Formula or Approach:

The relationship between the remaining activity (A), the initial activity (A_0), the time elapsed (t), and the half-life ($T_{1/2}$) is given by:

$$\frac{A}{A_0} = \left(\frac{1}{2}\right)^n$$

where n is the number of half-lives, and $n = \frac{t}{T_{1/2}}$.

Step 3: Detailed Explanation:

We are given:

Half-life, $T_{1/2} = 20$ minutes.

The final activity is $\frac{1}{16}$ of the initial activity, so $\frac{A}{A_0} = \frac{1}{16}$.

We can write $\frac{1}{16}$ as a power of $\frac{1}{2}$:

$$\frac{1}{16} = \frac{1}{2^4} = \left(\frac{1}{2}\right)^4$$

Comparing this with the formula $\frac{A}{A_0} = \left(\frac{1}{2}\right)^n$, we find that the number of half-lives, n, is 4. Now, we can find the total time (t) using the relation $n = \frac{t}{T_{1/2}}$:

$$4 = \frac{t}{20 \text{ minutes}}$$

 $t = 4 \times 20 \text{ minutes} = 80 \text{ minutes}$

Step 4: Final Answer:

The time required for the activity to drop to 1/16th of its initial value is 80 minutes. This corresponds to option (A).

Quick Tip

For fractions that are powers of 2 (like 1/2, 1/4, 1/8, 1/16), you can quickly find the number of half-lives. For 1/16, it's 4 half-lives because $2^4 = 16$. Then simply multiply the number of half-lives by the given half-life period.

3. The temperature of a gas is -50° C. To what temperature the gas should be heated so that the rms speed is increased by 3 times?

- (A) 223 K
- (B) 669° C
- (C) 3295° C
- (D) 3097 K

Correct Answer: (C) 3295° C

Solution:

Step 1: Understanding the Question:

The question asks for the final temperature to which a gas must be heated to increase its root-mean-square (rms) speed by a factor of 3. A crucial part of the interpretation is the phrase "increased by 3 times".

Step 2: Key Formula or Approach:

The rms speed (v_{rms}) of gas molecules is directly proportional to the square root of the absolute temperature (T in Kelvin).

$$v_{rms} = \sqrt{\frac{3RT}{M}}$$

Therefore, $v_{rms} \propto \sqrt{T}$.

This gives the relation $\frac{v_2}{v_1} = \sqrt{\frac{T_2}{T_1}}$.

Step 3: Detailed Explanation:

First, convert the initial temperature from Celsius to Kelvin.

Initial temperature, $T_1 = -50^{\circ}\text{C} + 273.15 \approx 223 \text{ K}.$

The phrase "increased by 3 times" means the final speed is the initial speed plus three times the initial speed.

Final speed, $v_2 = v_1 + 3v_1 = 4v_1$.

So, the ratio of the speeds is $\frac{v_2}{v_1} = 4$.

Now, use the proportionality relation:

$$\frac{v_2}{v_1} = \sqrt{\frac{T_2}{T_1}}$$

$$4 = \sqrt{\frac{T_2}{223 \text{ K}}}$$

Squaring both sides of the equation:

$$16 = \frac{T_2}{223 \text{ K}}$$
 $T_2 = 16 \times 223 \text{ K} = 3568 \text{ K}$

The options are given in both Kelvin and Celsius. Option (D) is 3097 K, which is incorrect. Let's convert our answer to Celsius to check the other options.

Final temperature in Celsius, $T_2(^{\circ}C) = 3568 - 273 = 3295^{\circ}C$.

Step 4: Final Answer:

The final temperature required is 3568 K, which is equal to 3295° C. This corresponds to option (C).

Quick Tip

In thermodynamics and gas laws, always convert temperatures to the absolute scale (Kelvin) before using them in formulas. Also, be careful with phrasing: "increased by 3 times" means $x \to x + 3x = 4x$, whereas "increased to 3 times" means $x \to 3x$.

- 4. The ratio of frequencies of fundamental harmonic produced by an open pipe to that of closed pipe having the same length is:
- (A) 3:1
- (B) 1:2
- (C) 2:1
- (D) 1:3

Correct Answer: (C) 2:1

Solution:

Step 1: Understanding the Question:

The question asks for the ratio of the fundamental frequencies of an open organ pipe and a closed organ pipe, given that they have the same length.

Step 2: Key Formula or Approach:

The fundamental frequency (f_o) of an open pipe of length L is given by:

$$f_{\text{open}} = \frac{v}{2L}$$

where v is the speed of sound in air.

The fundamental frequency (f_c) of a closed pipe of length L is given by:

$$f_{\text{closed}} = \frac{v}{4L}$$

Step 3: Detailed Explanation:

We need to find the ratio $\frac{f_{\text{open}}}{f_{\text{closed}}}$. Using the formulas from Step 2:

$$\frac{f_{\text{open}}}{f_{\text{closed}}} = \frac{\frac{v}{2L}}{\frac{v}{4L}}$$

$$\frac{f_{\rm open}}{f_{\rm closed}} = \frac{v}{2L} \times \frac{4L}{v}$$

The terms 'v' and 'L' cancel out.

$$\frac{f_{\text{open}}}{f_{\text{closed}}} = \frac{4}{2} = \frac{2}{1}$$

So, the ratio of the frequencies is 2:1.

Step 4: Final Answer:

The ratio of the fundamental frequency of an open pipe to that of a closed pipe of the same length is 2:1. This corresponds to option (C).

Quick Tip

A simple way to remember this is that a closed pipe can fit only a quarter of a wavelength $(\lambda/4)$ for its fundamental mode, while an open pipe fits half a wavelength $(\lambda/2)$. Since frequency is inversely proportional to wavelength $(f = v/\lambda)$, the open pipe's frequency will be double that of the closed pipe.

- 5. In a series LCR circuit, the inductance L is 10 mH, capacitance C is 1 μF and resistance R is 100 Ω . The frequency at which resonance occurs is :
- (A) 1.59 kHz
- (B) 15.9 rad/s
- (C) 15.9 kHz
- (D) 1.59 rad/s

Correct Answer: (A) 1.59 kHz

Solution:

Step 1: Understanding the Question:

The question asks for the resonance frequency of a series LCR circuit with given values for inductance (L), capacitance (C), and resistance (R).

Step 2: Key Formula or Approach:

Resonance in a series LCR circuit occurs when the inductive reactance (X_L) equals the capacitive reactance (X_C) . The resonance frequency (f_0) is given by the formula:

$$f_0 = \frac{1}{2\pi\sqrt{LC}}$$

The angular resonance frequency (ω_0) is given by:

$$\omega_0 = \frac{1}{\sqrt{LC}}$$

Note that $f_0 = \omega_0/2\pi$. The resistance (R) does not affect the resonance frequency.

Step 3: Detailed Explanation:

First, ensure all units are in the standard SI system.

$$L = 10 \text{ mH} = 10 \times 10^{-3} \text{ H} = 10^{-2} \text{ H}.$$

$$C = 1 \mu F = 1 \times 10^{-6} F.$$

Now, substitute these values into the formula for f_0 :

$$f_0 = \frac{1}{2\pi\sqrt{10^{-2} \times 10^{-6}}}$$

$$f_0 = \frac{1}{2\pi\sqrt{10^{-8}}}$$

$$f_0 = \frac{1}{2\pi \times 10^{-4}}$$

$$f_0 = \frac{10^4}{2\pi} \text{ Hz}$$

Now, calculate the numerical value using $\pi \approx 3.14159$:

$$f_0 = \frac{10000}{2 \times 3.14159} = \frac{10000}{6.28318} \approx 1591.5 \text{ Hz}$$

To express this in kHz, divide by 1000:

$$f_0 \approx 1.5915 \text{ kHz}$$

This value is approximately 1.59 kHz.

Step 4: Final Answer:

The resonance frequency is 1.59 kHz. This corresponds to option (A). Note that options (B) and (D) are in rad/s, which represent angular frequency, not frequency.

Quick Tip

Always check the units in the options. Here, options are given in both kHz (frequency) and rad/s (angular frequency). This is a common way to test if you understand the difference between f and ω . The resistance value is extra information designed to distract you; it's only needed for calculating the quality factor (Q-factor) or bandwidth.

- 6. In hydrogen spectrum, the shortest wavelength in the Balmer series is . The shortest wavelength in the Bracket series is :
- (A) 16
- (B) 2
- (C) 4
- (D) 9

Correct Answer: (C) 4

Solution:

Step 1: Understanding the Question:

The question relates the shortest wavelength of the Balmer series to the shortest wavelength of the Brackett series in the hydrogen spectrum.

Step 2: Key Formula or Approach:

The Rydberg formula gives the wavelength of emitted photons for electron transitions in a hydrogen-like atom:

$$\frac{1}{\lambda} = RZ^2 \left(\frac{1}{n_f^2} - \frac{1}{n_i^2} \right)$$

For hydrogen, Z=1. The shortest wavelength in any series corresponds to the transition from $n_i = \infty$ to the final state n_f . This is also known as the series limit.

For the Balmer series, $n_f = 2$.

For the Brackett series, $n_f = 4$.

Step 3: Detailed Explanation:

For the shortest wavelength in the Balmer series (λ), we set $n_f = 2$ and $n_i = \infty$.

$$\frac{1}{\lambda} = R\left(\frac{1}{2^2} - \frac{1}{\infty^2}\right) = R\left(\frac{1}{4} - 0\right) = \frac{R}{4}$$

$$\implies \lambda = \frac{4}{R} \quad -(1)$$

For the shortest wavelength in the Brackett series (let's call it λ_{Br}), we set $n_f = 4$ and $n_i = \infty$.

$$\frac{1}{\lambda_{Br}} = R\left(\frac{1}{4^2} - \frac{1}{\infty^2}\right) = R\left(\frac{1}{16} - 0\right) = \frac{R}{16}$$

$$\implies \lambda_{Br} = \frac{16}{R} \quad -(2)$$

Now, we express λ_{Br} in terms of λ . From equation (1), we have $R = 4/\lambda$. Substitute this into equation (2):

$$\lambda_{Br} = \frac{16}{R} = \frac{16}{(4/\lambda)} = \frac{16\lambda}{4} = 4\lambda$$

Alternatively, by comparing (1) and (2):

$$\lambda_{Br} = \frac{16}{R} = 4 \times \left(\frac{4}{R}\right) = 4\lambda$$

Step 4: Final Answer:

The shortest wavelength in the Brackett series is 4. This corresponds to option (C).

Remember the final energy levels for the first few series: Lyman $(n_f = 1)$, Balmer $(n_f = 2)$, Paschen $(n_f = 3)$, Brackett $(n_f = 4)$. The shortest wavelength (highest energy) in a series is the series limit, corresponding to a transition from $n_i = \infty$. The energy of the final state is $E_n = -13.6/n^2$. The series limit wavelength is inversely proportional to this energy, so $\lambda_{limit} \propto n_f^2$. The ratio would be $\lambda_{Br}/\lambda_{Ba} = (4^2)/(2^2) = 16/4 = 4$.

7. The angular acceleration of a body, moving along the circumference of a circle, is:

- (A) along the axis of rotation
- (B) along the radius, away from centre
- (C) along the radius towards the centre
- (D) along the tangent to its position

Correct Answer: (A) along the axis of rotation

Solution:

Step 1: Understanding the Question:

This is a conceptual question asking for the direction of the angular acceleration vector for a body in circular motion.

Step 2: Detailed Explanation:

Angular velocity $(\vec{\omega})$ and angular acceleration $(\vec{\alpha})$ are axial vectors. Their direction is defined by the axis of rotation.

For a particle moving in a plane (e.g., the x-y plane), the axis of rotation is perpendicular to that plane (e.g., the z-axis).

The direction is determined by the right-hand rule. If you curl the fingers of your right hand in the direction of rotation, your thumb points in the direction of the angular velocity vector $\vec{\omega}$.

Angular acceleration, $\vec{\alpha}$, is defined as the rate of change of angular velocity: $\vec{\alpha} = \frac{d\vec{\omega}}{dt}$.

Since $\vec{\omega}$ is along the axis of rotation, any change in its magnitude (speeding up or slowing down the rotation) will result in an acceleration vector $\vec{\alpha}$ that is also along the same axis.

- If the body is speeding up, $\vec{\alpha}$ is in the same direction as $\vec{\omega}$.
- If the body is slowing down, $\vec{\alpha}$ is in the opposite direction to $\vec{\omega}$.

In both cases, the angular acceleration vector lies along the axis of rotation.

The other options describe linear quantities:

- (C) along the radius towards the centre: Direction of centripetal acceleration.
- (D) along the tangent to its position: Direction of tangential velocity and tangential acceleration.

Step 3: Final Answer:

The angular acceleration of a body moving along the circumference of a circle is along the axis

of rotation. This corresponds to option (A).

Quick Tip

Remember the distinction between linear and angular quantities in circular motion. Linear quantities (position, velocity, acceleration) are in the plane of motion. Angular quantities (angular displacement, angular velocity, angular acceleration) are axial vectors, perpendicular to the plane of motion, along the axis of rotation.

8. An ac source is connected to a capacitor C. Due to decrease in its operating frequency:

- (A) capacitive reactance remains constant
- (B) capacitive reactance decreases.
- (C) displacement current increases.
- (D) displacement current decreases.

Correct Answer: (D) displacement current decreases.

Solution:

Step 1: Understanding the Question:

The question analyzes how the properties of a capacitor in an AC circuit change when the frequency of the AC source is decreased.

Step 2: Key Formula or Approach:

1. The capacitive reactance (X_C) is the opposition offered by a capacitor to the flow of alternating current. It is given by:

$$X_C = \frac{1}{\omega C} = \frac{1}{2\pi f C}$$

where f is the operating frequency.

2. The current (I) in the circuit is given by Ohm's law for AC circuits:

$$I = \frac{V}{X_C}$$

where V is the voltage of the source.

3. In a pure capacitive circuit, the displacement current (I_d) is equal to the conduction current (I).

Step 3: Detailed Explanation:

We are given that the operating frequency, f, decreases.

Let's analyze the effect on capacitive reactance (X_C) :

Since $X_C = \frac{1}{2\pi fC}$, X_C is inversely proportional to f.

As f decreases, X_C increases. Therefore, options (A) and (B) are incorrect.

Now let's analyze the effect on the current. The current in the circuit is $I = \frac{V}{X_C}$.

Since X_C increases and the source voltage V is assumed constant, the current I must decrease.

The displacement current in the capacitor, I_d , is equal to the conduction current, I.

Therefore, as the frequency decreases, the displacement current also decreases. This means option (C) is incorrect and option (D) is correct.

Step 4: Final Answer:

Due to a decrease in operating frequency, the capacitive reactance increases, which causes the current (and thus the displacement current) to decrease. This corresponds to option (D).

Quick Tip

Remember the frequency dependence for capacitors and inductors: - Capacitor: Reactance $X_C \propto 1/f$. Acts as an open circuit (infinite reactance) for DC (f=0). - Inductor: Reactance $X_L \propto f$. Acts as a short circuit (zero reactance) for DC (f=0). This helps in quickly determining the behavior at low and high frequencies.

- 9. The errors in the measurement which arise due to unpredictable fluctuations in temperature and voltage supply are:
- (A) Random errors
- (B) Instrumental errors
- (C) Personal errors
- (D) Least count errors

Correct Answer: (A) Random errors

Solution:

Step 1: Understanding the Question:

The question asks to classify the type of measurement error that results from unpredictable changes in experimental conditions like temperature and voltage.

Step 2: Detailed Explanation:

Let's define the different types of errors listed in the options:

- (A) Random errors: These are errors that occur irregularly and are hence random. They arise from unpredictable fluctuations in experimental conditions (e.g., temperature, voltage supply, mechanical vibrations) or from limitations in reading an instrument. They can be positive or negative. The effect of random errors can be minimized by taking multiple observations and calculating their mean. The description in the question perfectly matches the definition of random errors.
- (B) Instrumental errors: These are a type of systematic error that arise from imperfections in the design or calibration of the measuring instrument. For example, a thermometer that

consistently reads 2°C higher than the actual temperature has an instrumental error.

- (C) Personal errors: These errors arise due to an individual's bias, lack of proper setting of the apparatus, or carelessness in taking observations without observing proper precautions. An example is parallax error. These are also a form of systematic error.
- (D) Least count errors: This is the error associated with the resolution of the instrument. The least count is the smallest value that can be measured by the instrument. This type of error is considered a random error, but "random error" is a broader category that includes other unpredictable fluctuations as well.

The errors described in the question—arising from "unpredictable fluctuations in temperature and voltage supply"—are the classic definition of random errors.

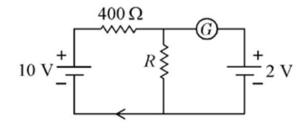
Step 3: Final Answer:

The errors due to unpredictable fluctuations in environmental or experimental conditions are classified as random errors. This corresponds to option (A).

Quick Tip

A good way to distinguish between systematic and random errors is to consider what happens when you repeat the measurement. Systematic errors will consistently affect the measurement in the same direction (e.g., always too high). Random errors will cause the readings to scatter randomly around the true value.

10. If the galvanometer G does not show any deflection in the circuit shown, the value of R is given by:



- (A) 400Ω
- (B) 200Ω
- (C) 50Ω
- (D) 100Ω

Correct Answer: (D) 100 Ω

Solution:

Step 1: Understanding the Question:

We are given a circuit and the condition that the galvanometer (G) shows no deflection. We

need to find the value of the unknown resistance R based on this condition.

Step 2: Key Formula or Approach:

No deflection in the galvanometer means that no current is flowing through it. This implies that the potential difference across the galvanometer is zero. Let the junction point between the $400~\Omega$ resistor and resistor R be P. The condition of no deflection means that the potential at point P is equal to the potential at the other end of the galvanometer, which is connected to the positive terminal of the 2 V battery. This circuit is an application of the potentiometer principle.

Step 3: Detailed Explanation:

Let's assume the negative terminals of both batteries are connected to a common ground with a potential of 0 V. Therefore, the potential at the positive terminal of the 10 V battery is 10 V, and the potential at the positive terminal of the 2 V battery is 2 V.

The resistors 400 Ω and R are connected in series across the 10 V source.

Since no current flows through the galvanometer, we can analyze the main loop consisting of the 10 V battery and the two resistors.

The total resistance in this loop is $R_{total} = 400 \Omega + R$.

The current flowing through this loop is given by Ohm's law:

$$I = \frac{V_{total}}{R_{total}} = \frac{10 \text{ V}}{400 + R}$$

The potential at point P (V_P) is the potential drop across the resistor R.

$$V_P = I \times R = \left(\frac{10}{400 + R}\right) R$$

The condition for no deflection in the galvanometer is that the potential at point P is equal to the potential of the positive terminal of the 2 V battery.

$$V_P = 2 \text{ V}$$

Now, we can set up the equation:

$$2 = \frac{10R}{400 + R}$$

Now, we solve for R:

$$2(400 + R) = 10R$$
$$800 + 2R = 10R$$
$$800 = 10R - 2R$$
$$800 = 8R$$
$$R = \frac{800}{8} = 100 \Omega$$

Step 4: Final Answer:

The value of the resistance R is 100Ω . This corresponds to option (D).

This problem can be quickly solved using the voltage divider rule. The potential at point P is given by $V_P = V_{total} \times \frac{R}{R_{total}}$. Here, $V_P = 2$ V, $V_{total} = 10$ V, and $R_{total} = 400 + R$. So, $2 = 10 \times \frac{R}{400 + R}$, which simplifies to the same equation.

- 11. The potential energy of a long spring when stretched by 2 cm is U. If the spring is stretched by 8 cm, potential energy stored in it will be:
- (A) 16U
- (B) 2U
- (C) 4U
- (D) 8U

Correct Answer: (A) 16U

Solution:

Step 1: Understanding the Question:

The question relates the potential energy stored in a spring at two different extensions. We are given the energy for a 2 cm stretch and asked to find the energy for an 8 cm stretch.

Step 2: Key Formula or Approach:

The potential energy (E) stored in a spring is given by the formula:

$$E = \frac{1}{2}kx^2$$

where k is the spring constant and x is the extension (or compression) from the equilibrium position. From this formula, we can see that the potential energy is directly proportional to the square of the extension: $E \propto x^2$.

Step 3: Detailed Explanation:

Let the initial case be denoted by subscript 1 and the final case by subscript 2.

In the initial case:

Extension, $x_1 = 2$ cm.

Potential energy, $E_1 = U$.

Using the formula: $U = \frac{1}{2}k(2)^2 = \frac{1}{2}k \times 4 = 2k$.

In the final case:

Extension, $x_2 = 8$ cm.

Potential energy, E_2 .

Using the formula: $E_2 = \frac{1}{2}k(8)^2 = \frac{1}{2}k \times 64 = 32k$.

To find E_2 in terms of U, we can take the ratio of the two energies:

$$\frac{E_2}{E_1} = \frac{\frac{1}{2}kx_2^2}{\frac{1}{2}kx_1^2} = \left(\frac{x_2}{x_1}\right)^2$$

14

Substituting the given values:

$$\frac{E_2}{U} = \left(\frac{8 \text{ cm}}{2 \text{ cm}}\right)^2 = (4)^2 = 16$$

$$E_2 = 16U$$

Step 4: Final Answer:

The potential energy stored in the spring when stretched by 8 cm will be 16U. This corresponds to option (A).

Quick Tip

For problems involving ratios with formulas that have squares (like this one, or kinetic energy $E_K = \frac{1}{2}mv^2$), you can solve them very quickly. Here, the extension increases by a factor of 4 (from 2 cm to 8 cm). Since energy is proportional to the square of the extension $(E \propto x^2)$, the energy will increase by a factor of $4^2 = 16$.

12. A vehicle travels half the distance with speed v and the remaining distance with speed 2v. Its average speed is:

- (A) $\frac{3v}{4}$ (B) $\frac{v}{3}$ (C) $\frac{2v}{3}$ (D) $\frac{4v}{3}$

Correct Answer: (D) $\frac{4v}{3}$

Solution:

Step 1: Understanding the Question:

The problem asks for the average speed of a vehicle that covers the first half of a total distance at a speed 'v' and the second half at a speed '2v'.

Step 2: Key Formula or Approach:

Average speed is defined as the total distance traveled divided by the total time taken.

$$\mbox{Average Speed} = \frac{\mbox{Total Distance}}{\mbox{Total Time}}$$

Step 3: Detailed Explanation:

Let the total distance be D.

The vehicle travels the first half of the distance, which is D/2, with speed v.

The time taken for the first half, t_1 , is:

$$t_1 = \frac{\text{Distance}}{\text{Speed}} = \frac{D/2}{v} = \frac{D}{2v}$$

15

The vehicle travels the second half of the distance, which is also D/2, with speed 2v. The time taken for the second half, t_2 , is:

$$t_2 = \frac{\text{Distance}}{\text{Speed}} = \frac{D/2}{2v} = \frac{D}{4v}$$

The total time taken for the journey is $T = t_1 + t_2$.

$$T = \frac{D}{2v} + \frac{D}{4v} = \frac{2D+D}{4v} = \frac{3D}{4v}$$

Now, we can calculate the average speed:

Step 4: Final Answer:

The average speed of the vehicle is $\frac{4v}{3}$. This corresponds to option (D).

Quick Tip

When an object covers two equal distances with speeds v_1 and v_2 , the average speed is the harmonic mean of the two speeds: Average Speed = $\frac{2v_1v_2}{v_1+v_2}$. Here, $v_1=v$ and $v_2=2v$, so the average speed is $\frac{2(v)(2v)}{v+2v}=\frac{4v^2}{3v}=\frac{4v}{3}$.

13. Given below are two statements:

Statement I: Photovoltaic devices can convert optical radiation into electricity.

Statement II : Zener diode is designed to operate under reverse bias in breakdown region.

In the light of the above statements, choose the most appropriate answer from the options given below:

- (A) Statement I is incorrect but Statement II is correct.
- (B) Both Statement I and Statement II are correct.
- (C) Both Statement I and Statement II are incorrect.
- (D) Statement I is correct but Statement II is incorrect.

Correct Answer: (B) Both Statement I and Statement II are correct.

Solution:

Step 1: Understanding the Question:

We need to evaluate the correctness of two separate statements related to semiconductor devices.

Step 2: Detailed Explanation:

Analysis of Statement I:

Photovoltaic devices, commonly known as solar cells, are designed to absorb photons from optical radiation (like sunlight). This absorption creates electron-hole pairs. An internal electric field within the device separates these charge carriers, creating a potential difference and driving a current. Thus, they directly convert optical energy into electrical energy. So, Statement I is correct.

Analysis of Statement II:

A Zener diode is a special type of diode that is heavily doped to have a sharp and well-defined reverse breakdown voltage. It is specifically designed to operate in the reverse breakdown region. When the reverse voltage across the diode reaches the Zener voltage (V_Z) , it starts conducting current without being damaged. This property of maintaining a constant voltage across it makes it useful as a voltage regulator. So, Statement II is correct.

Step 3: Final Answer:

Since both statements are correct descriptions of their respective devices, the correct option is (B).

Quick Tip

Remember the primary application for key semiconductor devices: - Photodiode/Photovoltaic cell: Converts light to electricity. - LED (Light Emitting Diode): Converts electricity to light. - Zener Diode: Used for voltage regulation in reverse breakdown. - Transistor: Used for amplification and switching.

- 14. A bullet is fired from a gun at the speed of 280 m s⁻¹ in the direction 30° above the horizontal. The maximum height attained by the bullet is $(g = 9.8 \text{ m s}^{-2}, \sin 30^{\circ} = 0.5)$:
- (A) 3000 m
- (B) 2800 m
- (C) 2000 m
- (D) 1000 m

Correct Answer: (D) 1000 m

Solution:

Step 1: Understanding the Question:

This is a problem of projectile motion. We are given the initial velocity and angle of projection and are asked to calculate the maximum vertical height reached by the bullet.

Step 2: Key Formula or Approach:

The formula for the maximum height (H) attained by a projectile is:

$$H = \frac{u^2 \sin^2 \theta}{2g}$$

where u is the initial speed, θ is the angle of projection, and g is the acceleration due to gravity.

Step 3: Detailed Explanation:

We are given the following values:

Initial speed, u = 280 m/s.

Angle of projection, $\theta = 30^{\circ}$.

Acceleration due to gravity, $q = 9.8 \text{ m/s}^2$.

We are also given $\sin 30^{\circ} = 0.5$.

Now, substitute these values into the formula:

$$H = \frac{(280)^2 (\sin 30^\circ)^2}{2 \times 9.8}$$

$$H = \frac{(280 \times 280) \times (0.5)^2}{19.6}$$

$$H = \frac{78400 \times 0.25}{19.6}$$

$$H = \frac{19600}{19.6}$$

To simplify the calculation, notice that $19.6 = 2 \times 9.8$ and $196 = 2 \times 98$.

$$H = \frac{19600}{19.6} = \frac{196000}{196} = 1000 \text{ m}$$

Step 4: Final Answer:

The maximum height attained by the bullet is 1000 m. This corresponds to option (D).

Quick Tip

In projectile motion, first, break down the initial velocity into horizontal $(u\cos\theta)$ and vertical $(u \sin \theta)$ components. The maximum height depends only on the vertical component. The time to reach maximum height is $t = u \sin \theta/g$, and using $v^2 = u^2 + 2as$, you get $0^2 = (u \sin \theta)^2 - 2gH$, which gives the same formula for H.

- 15. Two bodies of mass m and 9m are placed at a distance R. The gravitational potential on the line joining the bodies where the gravitational field equals zero, will be (G = gravitational constant):

- $(A) \frac{20Gm}{R}$ $(B) \frac{8Gm}{R}$ $(C) \frac{12Gm}{R}$

(D)
$$-\frac{16\text{Gm}}{R}$$

Correct Answer: (D) $-\frac{16\text{Gm}}{R}$

Solution:

Step 1: Understanding the Question:

First, we need to find the point on the line joining the two masses where the net gravitational field is zero. Then, we need to calculate the net gravitational potential at that point. Note: Gravitational potential is a scalar quantity and is always negative. The options in the question paper missed the negative sign.

Step 2: Find the point of zero gravitational field.

Let the point of zero field be at a distance x from the mass m. The distance from the mass 9m will then be (R-x).

The gravitational field due to a mass M at a distance r is $E = GM/r^2$.

For the net field to be zero, the magnitudes of the fields from the two masses must be equal:

$$E_m = E_{9m}$$

$$\frac{Gm}{x^2} = \frac{G(9m)}{(R-x)^2}$$

Canceling Gm from both sides:

$$\frac{1}{x^2} = \frac{9}{(R-x)^2}$$

Taking the square root of both sides:

$$\frac{1}{x} = \frac{3}{R - x}$$

$$R - x = 3x$$

$$R = 4x \implies x = \frac{R}{4}$$

So, the point is at a distance R/4 from mass m and (R - R/4) = 3R/4 from mass 9m.

Step 3: Calculate the gravitational potential at this point.

The gravitational potential due to a mass M at a distance r is V = -GM/r.

The total potential (V) at the point is the sum of the potentials due to each mass:

$$V = V_m + V_{9m}$$

$$V = \left(-\frac{Gm}{x}\right) + \left(-\frac{G(9m)}{R - x}\right)$$

Substitute the distances we found: x = R/4 and R - x = 3R/4.

$$V = -\frac{Gm}{R/4} - \frac{G(9m)}{3R/4}$$

19

$$V = -\frac{4Gm}{R} - \frac{9 \times 4Gm}{3R}$$

$$V = -\frac{4Gm}{R} - \frac{12Gm}{R}$$

$$V = -\frac{16Gm}{R}$$

Step 4: Final Answer:

The gravitational potential at the point of zero gravitational field is $-\frac{16Gm}{R}$. This corresponds to option (D).

Quick Tip

For two masses m_1 and m_2 , the null point (zero field) is always closer to the smaller mass. The distances from the masses are in the ratio $\sqrt{m_1}$: $\sqrt{m_2}$. Here, the ratio of masses is 1:9, so the ratio of distances will be $\sqrt{1}$: $\sqrt{9} = 1$: 3. So, you can directly find the point by dividing R in the ratio 1:3, which gives distances R/4 and 3R/4.

The minimum wavelength of X-rays produced by an electron accelerated through a potential difference of V volts is proportional to:

- (A) V^2
- (B) \sqrt{V}
- $\begin{array}{c}
 (C) \frac{1}{V} \\
 (D) \frac{1}{\sqrt{V}}
 \end{array}$

Correct Answer: (C) $\frac{1}{V}$

Solution:

Step 1: Understanding the Question:

The question asks for the relationship between the minimum wavelength (λ_{min}) of X-rays produced in an X-ray tube and the accelerating potential difference (V).

Step 2: Key Formula or Approach:

When an electron is accelerated through a potential difference V, it gains kinetic energy $E_k = eV$, where e is the charge of the electron.

When this electron strikes the target, it can lose all its kinetic energy in a single collision, producing an X-ray photon of maximum energy.

The energy of a photon (E) is related to its wavelength (λ) by the formula $E = \frac{hc}{\lambda}$, where h is Planck's constant and c is the speed of light.

The minimum wavelength corresponds to the maximum possible photon energy.

Step 3: Detailed Explanation:

The maximum energy of the X-ray photon (E_{max}) is equal to the kinetic energy of the incident

20

electron:

$$E_{max} = eV$$

The energy of this photon is also given by:

$$E_{max} = \frac{hc}{\lambda_{min}}$$

Equating the two expressions for energy:

$$eV = \frac{hc}{\lambda_{min}}$$

Rearranging the formula to find the minimum wavelength:

$$\lambda_{min} = \frac{hc}{eV}$$

Since h, c, and e are constants, we can see the relationship between λ_{min} and V:

$$\lambda_{min} \propto \frac{1}{V}$$

The minimum wavelength of the X-rays is inversely proportional to the accelerating voltage.

Step 4: Final Answer:

The minimum wavelength of X-rays is proportional to $\frac{1}{V}$. This corresponds to option (C).

Quick Tip

This relationship is also known as the Duane-Hunt law. A useful value to remember is $hc \approx 1240$ eV-nm. This allows for quick calculation of the minimum wavelength in nanometers if the voltage is given in volts: $\lambda_{min}(\text{nm}) = \frac{1240}{V(\text{volts})}$.

17. The venturi-meter works on:

- (A) The principle of perpendicular axes
- (B) Huygen's principle
- (C) Bernoulli's principle
- (D) The principle of parallel axes

Correct Answer: (C) Bernoulli's principle

Solution:

Step 1: Understanding the Question:

This is a direct-knowledge question asking for the scientific principle behind the operation of a

venturi-meter.

Step 2: Detailed Explanation:

A venturi-meter is a device used to measure the rate of flow of a fluid through a pipe. It consists of a constricted section (the throat) between two wider sections.

According to the principle of continuity $(A_1v_1 = A_2v_2)$, the fluid's speed increases as it passes through the narrower throat.

Bernoulli's principle states that for a horizontal flow, an increase in the speed of a fluid occurs simultaneously with a decrease in pressure. The principle is an application of the conservation of energy to fluid flow.

$$P + \frac{1}{2}\rho v^2 + \rho gh = \text{constant}$$

In the venturi-meter, the increased velocity in the throat leads to a lower pressure compared to the wider sections. By measuring this pressure difference, the flow rate can be calculated. Therefore, the operation of a venturi-meter is a direct application of Bernoulli's principle. The other options are unrelated:

- (A) and (D) The principles of perpendicular and parallel axes are theorems related to the moment of inertia in rotational mechanics.
- (B) Huygen's principle is a concept in wave optics that describes how waves propagate.

Step 3: Final Answer:

The venturi-meter works on Bernoulli's principle. This corresponds to option (C).

Quick Tip

Remember the key applications of Bernoulli's principle: venturi-meter, atomizer/sprayer, lift on an airplane wing, and the swinging of a spinning ball (Magnus effect). Associating devices with their underlying principles is key for theory questions.

- 18. Resistance of a carbon resistor determined from colour codes is (22000 \pm 5%) Ω . The colour of third band must be :
- (A) Yellow
- (B) Red
- (C) Green
- (D) Orange

Correct Answer: (D) Orange

Solution:

Step 1: Understanding the Question:

The question asks for the color of the third band of a carbon resistor given its resistance value

in scientific notation.

Step 2: Key Formula or Approach:

For a four-band carbon resistor: - The first band gives the first significant digit. - The second band gives the second significant digit. - The third band is the decimal multiplier (power of 10). - The fourth band indicates the tolerance.

The color code mnemonic is: **B B R O Y** of **G**reat **B**ritain has a **V**ery **G**ood **W**ife. **B**lack (0), **B**rown (1), **R**ed (2), **O**range (3), **Y**ellow (4), **G**reen (5), **B**lue (6), **V**iolet (7), **G**rey (8), **W**hite (9).

Step 3: Detailed Explanation:

The given resistance is 22000Ω .

We first write this in standard form with two significant figures:

$$22000 \Omega = 22 \times 1000 \Omega = 22 \times 10^{3} \Omega$$

Now, let's decode this based on the bands:

- First significant digit is 2. The color for 2 is **Red**.
- Second significant digit is 2. The color for 2 is **Red**.
- The multiplier is 10^3 . The color for a multiplier of 10^3 is **Orange**.
- The tolerance is \pm 5%, which corresponds to the color **Gold**.

The question specifically asks for the color of the third band, which represents the multiplier. The multiplier is 10^3 , so the third band is **Orange**.

Step 4: Final Answer:

The colour of the third band must be Orange. This corresponds to option (D).

Quick Tip

When decoding a resistance value, first write it in the form $AB \times 10^{C}$. A gives the first color, B gives the second, and the power C gives the third color. In this case, $22000 = 22 \times 10^{3}$. A=2 (Red), B=2 (Red), C=3 (Orange).

- 19. The amount of energy required to form a soap bubble of radius 2 cm from a soap solution is nearly: (surface tension of soap solution = 0.03 N m^{-1})
- (A) $50.1 \times 10^{-4} \text{ J}$
- (B) $30.16 \times 10^{-4} \text{ J}$
- (C) $5.06 \times 10^{-4} \text{ J}$
- (D) $3.01 \times 10^{-4} \text{ J}$

Correct Answer: (D) $3.01 \times 10^{-4} \text{ J}$

Solution:

Step 1: Understanding the Question:

We need to calculate the energy required to create a soap bubble of a given radius. This energy is the surface energy, which is the work done against surface tension to create the bubble's surface area.

Step 2: Key Formula or Approach:

The energy (E) or work done (W) to create a surface is given by:

$$E = T \times \Delta A$$

where T is the surface tension and ΔA is the increase in the surface area.

A crucial point for a soap bubble is that it has two surfaces: an inner surface and an outer surface. Therefore, the total surface area is twice the area of a sphere.

Area of a sphere = $4\pi r^2$.

Total surface area of a soap bubble = $2 \times (4\pi r^2) = 8\pi r^2$.

Step 3: Detailed Explanation:

First, convert all units to SI units.

Radius, r = 2 cm = 0.02 m.

Surface tension, $T = 0.03 \text{ N m}^{-1}$.

The increase in surface area (ΔA) when forming the bubble from a solution (negligible initial area) is the final surface area of the bubble.

$$\Delta A = 8\pi r^2$$

Now, calculate the energy:

$$E = T \times \Delta A = T \times (8\pi r^2)$$

$$E = 0.03 \times 8 \times \pi \times (0.02)^2$$

$$E = 0.24 \times \pi \times 0.0004$$

$$E = 0.24 \times 3.14159 \times 0.0004$$

$$E = 0.75398 \times 0.0004$$

$$E \approx 0.00030159 \text{ J}$$

Expressing this in scientific notation:

$$E \approx 3.0159 \times 10^{-4} \text{ J}$$

This value is very close to $3.01 \times 10^{-4} \text{ J}.$

Step 4: Final Answer:

The energy required is approximately 3.01×10^{-4} J. This corresponds to option (D).

A very common mistake is to forget that a soap bubble has two surfaces. A liquid drop has only one surface $(A = 4\pi r^2)$, but a bubble has both an inside and an outside surface $(A = 8\pi r^2)$. Always check if the problem is about a drop or a bubble.

- 20. A full wave rectifier circuit consists of two p-n junction diodes, a centre-tapped transformer, capacitor and a load resistance. Which of these components remove the ac ripple from the rectified output?
- (A) Load resistance
- (B) A centre-tapped transformer
- (C) p-n junction diodes
- (D) Capacitor

Correct Answer: (D) Capacitor

Solution:

Step 1: Understanding the Question:

The question asks about the function of different components in a full-wave rectifier circuit, specifically which component is used for filtering or smoothing the output.

Step 2: Detailed Explanation:

Let's analyze the role of each component in a full-wave rectifier:

- Centre-tapped transformer: Steps down the AC voltage and provides two outputs that are 180° out of phase. This is essential for full-wave rectification.
- **p-n junction diodes:** These act as one-way gates for current. In a full-wave rectifier, they conduct on alternate half-cycles of the AC input, converting the AC into a pulsating DC output.
- Load resistance (R_L) : The output voltage is developed across this resistor. It's where the rectified power is delivered.
- Capacitor: The output from the diodes is a pulsating DC voltage, not a steady one. It contains a significant AC component, known as ripple. A capacitor is connected in parallel with the load resistance to act as a filter. The capacitor charges up when the voltage is rising and then discharges slowly through the load when the voltage is falling. This process smooths out the pulsating DC, significantly reducing the AC ripple and producing a more stable DC output.

Step 3: Final Answer:

The capacitor is the component used to remove the AC ripple from the rectified output. This corresponds to option (D).

In rectifier circuits, the process of converting AC to DC is called rectification (done by diodes). The process of smoothing the pulsating DC output to reduce ripples is called filtering (commonly done by a capacitor, forming a simple low-pass filter).

21. The net magnetic flux through any closed surface is:

- (A) Negative
- (B) Zero
- (C) Positive
- (D) Infinity

Correct Answer: (B) Zero

Solution:

Step 1: Understanding the Question:

This question asks for the value of the net magnetic flux passing through any arbitrary closed surface.

Step 2: Key Formula or Approach:

This is a direct question about Gauss's law for magnetism, which is one of Maxwell's equations. The law is mathematically stated as:

$$\oint_{S} \vec{B} \cdot d\vec{S} = 0$$

where \vec{B} is the magnetic field, and the integral is taken over any closed surface S.

Step 3: Detailed Explanation:

Gauss's law for magnetism states that the net magnetic flux out of any closed surface is zero. The physical reason for this is that magnetic field lines are always continuous closed loops. They do not start or end at any point. This means that for any closed surface, the number of magnetic field lines entering the surface is always equal to the number of magnetic field lines leaving it.

This is also a statement of the experimental fact that magnetic monopoles (isolated north or south poles) do not exist. If they did, they would act as sources or sinks of magnetic field lines, and the flux through a surface enclosing a monopole would be non-zero. Since no magnetic monopoles have ever been observed, the net magnetic flux through any closed surface is always zero.

Step 4: Final Answer:

The net magnetic flux through any closed surface is always zero. This corresponds to option (B).

Contrast this with Gauss's law for electricity: $\oint_S \vec{E} \cdot d\vec{S} = Q_{enc}/\epsilon_0$. Electric flux is non-zero if there's a net charge inside, because electric field lines can start and end on charges (electric monopoles exist). Magnetic flux is always zero because magnetic monopoles do not exist.

- 22. The work functions of Caesium (Cs), Potassium (K) and Sodium (Na) are 2.14 eV, 2.30 eV and 2.75 eV respectively. If incident electromagnetic radiation has an incident energy of 2.20 eV, which of these photosensitive surfaces may emit photoelectrons?
- (A) Na only
- (B) Cs only
- (C) Both Na and K
- (D) K only

Correct Answer: (B) Cs only

Solution:

Step 1: Understanding the Question:

The question is about the photoelectric effect. We are given the work functions of three different metals and the energy of the incident photons. We need to determine which metal(s) will exhibit the photoelectric effect.

Step 2: Key Formula or Approach:

The condition for the photoelectric effect to occur is that the energy of the incident photon (E) must be greater than or equal to the work function (ϕ) of the material.

$$E \ge \phi$$

The work function is the minimum energy required to remove an electron from the surface of the material.

Step 3: Detailed Explanation:

We are given the following data:

Energy of incident radiation, E = 2.20 eV.

Work function of Caesium, $\phi_{Cs} = 2.14 \text{ eV}$.

Work function of Potassium, $\phi_K = 2.30 \text{ eV}$.

Work function of Sodium, $\phi_{Na} = 2.75 \text{ eV}.$

Now, we check the condition $E \ge \phi$ for each metal:

- For Caesium (Cs): E = 2.20 eV and $\phi_{Cs} = 2.14 \text{ eV}$. Since 2.20 eV > 2.14 eV, the condition is satisfied. Caesium will emit photoelectrons.

- For Potassium (K): E = 2.20 eV and $\phi_K = 2.30 \text{ eV}$. Since 2.20 eV < 2.30 eV, the condition is not satisfied. Potassium will not emit photoelectrons.
- For Sodium (Na): E = 2.20 eV and $\phi_{Na} = 2.75$ eV. Since 2.20 eV < 2.75 eV, the condition is not satisfied. Sodium will not emit photoelectrons.

Therefore, only Caesium will emit photoelectrons.

Step 4: Final Answer:

Only Caesium (Cs) has a work function less than the incident energy, so only it will emit photoelectrons. This corresponds to option (B).

Quick Tip

The photoelectric effect is like an "energy toll". The photon's energy is the money you have, and the work function is the toll price. You can only pass (emit an electron) if your money is greater than or equal to the toll. Any extra money becomes the kinetic energy of the electron.

- 23. A football player is moving southward and suddenly turns eastward with the same speed to avoid an opponent. The force that acts on the player while turning is:
- (A) along south-west
- (B) along eastward
- (C) along northward
- (D) along north-east

Correct Answer: (D) along north-east

Solution:

Step 1: Understanding the Question:

The question asks for the direction of the net force acting on a player who changes their velocity. According to Newton's second law, the direction of the net force is the same as the direction of the acceleration, which is the direction of the change in velocity.

Step 2: Key Formula or Approach:

Force $\vec{F} = m\vec{a}$.

Acceleration $\vec{a} = \frac{\Delta \vec{v}}{\Delta t} = \frac{\vec{v}_f - \vec{v}_i}{\Delta t}$.

The direction of the force is the same as the direction of the change in velocity, $\Delta \vec{v} = \vec{v}_f - \vec{v}_i$.

Step 3: Detailed Explanation:

Let's set up a coordinate system. Let the north direction be the positive y-axis and the east direction be the positive x-axis.

The initial velocity $(\vec{v_i})$ is southward with speed v. In vector form:

$$\vec{v}_i = -v\hat{j}$$

The final velocity (\vec{v}_f) is eastward with the same speed v. In vector form:

$$\vec{v}_f = v\hat{i}$$

Now, calculate the change in velocity $\Delta \vec{v}$:

$$\Delta \vec{v} = \vec{v}_f - \vec{v}_i = (v\hat{i}) - (-v\hat{j})$$
$$\Delta \vec{v} = v\hat{i} + v\hat{j}$$

The vector $v\hat{i} + v\hat{j}$ has a component in the positive x-direction (east) and a component in the positive y-direction (north).

A vector with equal positive components in the east and north directions points exactly northeast.

Since the force \vec{F} is in the same direction as $\Delta \vec{v}$, the force acts along the north-east direction.

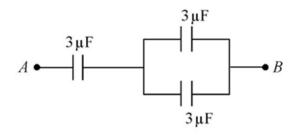
Step 4: Final Answer:

The force that acts on the player is directed along the north-east. This corresponds to option (D).

Quick Tip

You can solve this graphically. Draw the initial velocity vector \vec{v}_i pointing down (south). Draw the final velocity vector \vec{v}_f pointing right (east). The change in velocity is $\vec{v}_f - \vec{v}_i$, which is the same as $\vec{v}_f + (-\vec{v}_i)$. The vector $-\vec{v}_i$ points up (north). Adding \vec{v}_f (east) and $-\vec{v}_i$ (north) using the parallelogram law gives a resultant vector pointing north-east.

24. The equivalent capacitance of the system shown in the following circuit is:



- $(A) 9 \mu F$
- (B) 2 μF
- (C) $3 \mu F$
- (D) $6 \mu F$

Correct Answer: (B) 2 µF

Solution:

Step 1: Understanding the Question:

The question asks for the equivalent capacitance of a given combination of three capacitors between points A and B.

Step 2: Key Formula or Approach:

- For capacitors in series, the equivalent capacitance C_{eq} is given by: $\frac{1}{C_{eq}} = \frac{1}{C_1} + \frac{1}{C_2} + \dots$ - For capacitors in parallel, the equivalent capacitance C_{eq} is given by: $C_{eq} = C_1 + C_2 + \dots$

Step 3: Detailed Explanation:

The top and bottom capacitors are in parallel with each other. Their combined capacitance is $C_p = 3 + 3 = 6 \mu F$.

This combination is in series with the leftmost capacitor.

The equivalent capacitance C_{eq} is:

$$\frac{1}{C_{eq}} = \frac{1}{3 \,\mu\text{F}} + \frac{1}{C_p} = \frac{1}{3} + \frac{1}{6}$$
$$\frac{1}{C_{eq}} = \frac{2+1}{6} = \frac{3}{6} = \frac{1}{2}$$
$$C_{eq} = 2 \,\mu\text{F}$$

This matches option (B).

So, the top and bottom capacitors are in parallel between nodes 2 and 3. $C_{parallel} = 3 \,\mu\text{F} + 3 \,\mu\text{F} = 6 \,\mu\text{F}$. Now, this parallel combination is in series with the leftmost capacitor, which is between nodes 1 (A) and 2. The total equivalent capacitance between A and B (node 1 and 3) is the series combination of the leftmost capacitor and $C_{parallel}$.

$$\frac{1}{C_{eq}} = \frac{1}{3 \ \mu F} + \frac{1}{6 \ \mu F} = \frac{2+1}{6} = \frac{3}{6} = \frac{1}{2}$$

$$C_{eq} = 2 \ \mu F$$

Step 4: Final Answer:

The top and bottom 3 μ F capacitors are in parallel, giving an equivalent of 6 μ F. This combination is in series with the leftmost 3 μ F capacitor. The total equivalent capacitance is 2 μ F. This corresponds to option (B).

Quick Tip

When a circuit diagram is ambiguous, try to redraw it by labeling the nodes (junctions). All points connected by a simple wire are at the same potential and can be considered the same node. This clarification often reveals simple series and parallel combinations.

25. An electric dipole is placed at an angle of 30° with an electric field of intensity 2×10^5 N C⁻¹. It experiences a torque equal to 4 N m. Calculate the magnitude of

charge on the dipole, if the dipole length is 2 cm.

- (A) 2 mC
- (B) 8 mC
- (C) 6 mC
- (D) 4 mC

Correct Answer: (A) 2 mC

Solution:

Step 1: Understanding the Question:

The question provides the torque experienced by an electric dipole in a uniform electric field, along with the field strength, the angle, and the dipole length. We need to find the magnitude of the charge on the dipole.

Step 2: Key Formula or Approach:

The torque (τ) on an electric dipole in a uniform electric field (E) is given by:

$$\tau = pE\sin\theta$$

where p is the magnitude of the electric dipole moment and θ is the angle between the dipole moment and the electric field.

The electric dipole moment p is defined as:

$$p = q \times d$$

where q is the magnitude of the charge and d is the separation between the charges (dipole length).

Step 3: Detailed Explanation:

First, let's list the given values and convert them to SI units.

Torque, $\tau = 4$ N m.

Electric field, $E = 2 \times 10^5$ N/C.

Angle, $\theta = 30^{\circ}$.

Dipole length, d = 2 cm = 0.02 m.

We know that $\sin 30^{\circ} = 0.5$.

First, we use the torque formula to find the dipole moment p.

$$\tau = pE \sin \theta$$

$$4 = p \times (2 \times 10^5) \times \sin 30^\circ$$

$$4 = p \times (2 \times 10^5) \times 0.5$$

$$4 = p \times 10^5$$

$$p = \frac{4}{10^5} = 4 \times 10^{-5} \text{ C m}$$

Now that we have the dipole moment, we can find the charge q using $p = q \times d$.

$$4 \times 10^{-5} = q \times 0.02$$
$$q = \frac{4 \times 10^{-5}}{0.02} = \frac{4 \times 10^{-5}}{2 \times 10^{-2}}$$
$$q = 2 \times 10^{-3} \text{ C}$$

Since 1 mC = 10^{-3} C, the charge is:

$$q = 2 \text{ mC}$$

Step 4: Final Answer:

The magnitude of the charge on the dipole is 2 mC. This corresponds to option (A).

Quick Tip

Remember to convert all units to the standard SI system before calculation (e.g., cm to m). Also, know the values of sine for common angles like 0° , 30° , 45° , 60° , and 90° . Maximum torque occurs at 90° (sin $90^{\circ} = 1$) and zero torque occurs at 0° and 180° .

26. If $\oint_S \vec{E} \cdot d\vec{s} = 0$ over a surface, then :

- (A) the electric field inside the surface is necessarily uniform.
- (B) the number of flux lines entering the surface must be equal to the number of flux lines leaving it.
- (C) the magnitude of electric field on the surface is constant.
- (D) all the charges must necessarily be inside the surface.

Correct Answer: (B) the number of flux lines entering the surface must be equal to the number of flux lines leaving it.

Solution:

Step 1: Understanding the Question:

The question presents a condition based on Gauss's law for electricity and asks for the correct physical interpretation of that condition. The condition is that the total electric flux through a closed surface is zero.

Step 2: Key Formula or Approach:

Gauss's law states that the total electric flux (Φ_E) through any closed surface (a Gaussian surface) is equal to the net electric charge (Q_{enc}) enclosed by the surface, divided by the permittivity of free space (ϵ_0) .

$$\Phi_E = \oint_S \vec{E} \cdot d\vec{s} = \frac{Q_{enc}}{\epsilon_0}$$

Step 3: Detailed Explanation:

We are given the condition $\oint_S \vec{E} \cdot d\vec{s} = 0$. From Gauss's law, this means:

$$\frac{Q_{enc}}{\epsilon_0} = 0 \implies Q_{enc} = 0$$

This tells us that the net charge enclosed within the surface is zero. This could mean there are no charges inside, or there are equal amounts of positive and negative charge inside.

Now let's evaluate the given options:

- (A) The electric field inside is not necessarily uniform. For example, consider an electric dipole placed at the center of a sphere. The net charge inside is zero, so the flux is zero, but the electric field inside is certainly not uniform.
- (B) Electric flux is a measure of the number of electric field lines passing through a surface. A positive flux corresponds to lines leaving the surface, and a negative flux corresponds to lines entering the surface. If the total flux is zero, it means the net number of lines crossing the surface is zero. This implies that the number of flux lines entering the surface must be equal to the number of flux lines leaving it. This statement is a direct and correct interpretation of zero net flux.
- (C) The magnitude of the electric field on the surface is not necessarily constant. A dipole at the center of a sphere would produce a non-constant field on the surface, yet the total flux would be zero.
- (D) This is incorrect. The condition implies that the net charge inside is zero. There could be charges outside the surface, or there could be a combination of positive and negative charges inside that sum to zero. It does not mean all charges must be inside. In fact, if there is a charge outside, it will contribute to the field at the surface, but its net flux through the closed surface will be zero.

Therefore, the most accurate conclusion is that the incoming and outgoing flux are balanced.

Step 4: Final Answer:

If the net electric flux over a surface is zero, it means that the number of flux lines entering the surface is equal to the number of flux lines leaving it. This corresponds to option (B).

Quick Tip

Remember that Gauss's Law relates the flux through a closed surface to the enclosed charge. Charges outside the surface contribute to the electric field at the surface but do not contribute to the net flux through the surface. Zero net flux implies zero net enclosed charge.

27. For Young's double slit experiment, two statements are given below: Statement I: If screen is moved away from the plane of slits, angular separation of the fringes remains constant. Statement II: If the monochromatic source is replaced by another monochromatic source of larger wavelength, the angular separation of fringes decreases.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (D) Statement I is true but Statement II is false.

Solution:

Step 1: Understanding the Question:

We need to evaluate two statements concerning the angular separation of fringes in a Young's double-slit experiment (YDSE).

Step 2: Key Formula or Approach:

In YDSE, the angular position (θ) of the nth bright fringe is given by $d \sin \theta = n\lambda$. For small angles, $\sin \theta \approx \theta$, so $d\theta \approx n\lambda$.

The angular separation (θ_{sep}) between consecutive bright fringes is the change in angle when n changes by 1.

$$\theta_{sep} = \frac{\lambda}{d}$$

where λ is the wavelength of light and d is the distance between the slits.

Step 3: Detailed Explanation:

Analysis of Statement I:

The formula for angular separation is $\theta_{sep} = \lambda/d$. This formula depends only on the wavelength (λ) and the slit separation (d). It does not depend on the distance between the slits and the screen (D). Therefore, if the screen is moved away from the plane of slits (i.e., D is changed), the angular separation of the fringes remains constant. So, Statement I is true.

Analysis of Statement II:

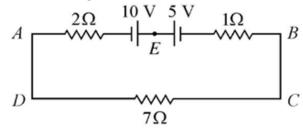
The formula is $\theta_{sep} = \lambda/d$. This shows that the angular separation is directly proportional to the wavelength ($\theta_{sep} \propto \lambda$). If the monochromatic source is replaced by another source of larger wavelength, the angular separation of the fringes will increase, not decrease. So, Statement II is false.

Step 4: Final Answer:

Based on the analysis, Statement I is true and Statement II is false. This corresponds to option (D).

Remember the difference between fringe width $(\beta = \lambda D/d)$ and angular separation $(\theta = \lambda/d)$. Fringe width depends on the screen distance D, but angular separation does not. This is a common point of confusion.

28. The magnitude and direction of the current in the following circuit is



- (A) 1.5 A from B to A through E
- (B) 0.2 A from B to A through E
- (C) 0.5 A from A to B through E
- (D) $\frac{5}{9}$ A from A to B through E

Correct Answer: (C) 0.5 A from A to B through E

Solution:

Step 1: Understanding the Question:

We need to find the magnitude and direction of the current flowing in the given series circuit containing two batteries and three resistors.

Step 2: Key Formula or Approach:

We can use Kirchhoff's Voltage Law (KVL), which states that the algebraic sum of potential differences around any closed loop is zero. The net current in a simple series circuit is given by $I = \frac{\text{Net EMF}}{\text{Total Resistance}}$.

Step 3: Detailed Explanation:

First, identify the components in the series loop.

Resistors: $R_1 = 2 \Omega$, $R_2 = 1 \Omega$, $R_3 = 7 \Omega$.

Total Resistance, $R_{total} = R_1 + R_2 + R_3 = 2 + 1 + 7 = 10 \Omega$.

Batteries: $V_1 = 10 \text{ V}, V_2 = 5 \text{ V}.$

The batteries are connected in series. Let's observe their polarity. The positive terminal of the 10 V battery and the positive terminal of the 5 V battery are connected together (at point E). This means they are connected in opposition.

The net EMF is the difference between their voltages:

$$E_{net} = E_1 - E_2 = 10 \text{ V} - 5 \text{ V} = 5 \text{ V}$$

35

The direction of the current will be determined by the stronger battery (10 V). The 10 V battery will push the current out of its positive terminal, which means the current will flow in the anti-clockwise direction (A \rightarrow E \rightarrow B \rightarrow C \rightarrow D \rightarrow A).

Now, calculate the magnitude of the current using Ohm's law for the entire circuit:

$$I = \frac{E_{net}}{R_{total}} = \frac{5 \text{ V}}{10 \Omega} = 0.5 \text{ A}$$

The direction of the current is anti-clockwise. In the segment AEB, the current flows from A to E and then to B. So the direction is from A to B through E.

Step 4: Final Answer:

The current is 0.5 A, and it flows from A to B through E. This corresponds to option (C).

Quick Tip

When batteries are in series, check their polarities. If they are connected positive-to-negative, their EMFs add up. If they are connected positive-to-positive or negative-to-negative, they oppose each other, and their net EMF is the difference. The current flows in the direction of the stronger battery.

29. The ratio of radius of gyration of a solid sphere of mass M and radius R about its own axis to the radius of gyration of the thin hollow sphere of same mass and radius about its axis is:

- (A) 5:2
- (B) 3:5
- (C) 5:3
- (D) 2:5
- (E) $\sqrt{3} : \sqrt{5}$

Correct Answer: (E) $\sqrt{3}$: $\sqrt{5}$

Solution:

Step 1: Understanding the Question:

The question asks for the ratio of the radius of gyration of a solid sphere to that of a hollow sphere, both having the same mass M and radius R, rotating about an axis passing through their centers.

Step 2: Key Formula or Approach:

The moment of inertia (I) is related to the radius of gyration (k) by the formula $I = Mk^2$, which means $k = \sqrt{\frac{I}{M}}$.

- Moment of inertia of a solid sphere about its axis: $I_{solid} = \frac{2}{5}MR^2$.
- Moment of inertia of a thin hollow sphere about its axis: $I_{hollow} = \frac{2}{3}MR^2$.

Step 3: Detailed Explanation:

First, find the radius of gyration for the solid sphere (k_{solid}) :

$$k_{solid}^2 = \frac{I_{solid}}{M} = \frac{\frac{2}{5}MR^2}{M} = \frac{2}{5}R^2$$
$$k_{solid} = \sqrt{\frac{2}{5}}R$$

Next, find the radius of gyration for the hollow sphere (k_{hollow}) :

$$k_{hollow}^2 = \frac{I_{hollow}}{M} = \frac{\frac{2}{3}MR^2}{M} = \frac{2}{3}R^2$$
$$k_{hollow} = \sqrt{\frac{2}{3}}R$$

Now, find the required ratio:

$$\frac{k_{solid}}{k_{hollow}} = \frac{\sqrt{\frac{2}{5}}R}{\sqrt{\frac{2}{3}}R} = \sqrt{\frac{2/5}{2/3}} = \sqrt{\frac{2}{5} \times \frac{3}{2}} = \sqrt{\frac{3}{5}}$$

The ratio is $\sqrt{3}$: $\sqrt{5}$.

Step 4: Final Answer:

The calculated ratio is $\sqrt{3}$: $\sqrt{5}$.

Quick Tip

Be very careful to distinguish between moment of inertia (I) and radius of gyration (k). They are related by $I = Mk^2$. A common error is to state the ratio of I instead of k, or the ratio of k^2 instead of k. Always double-check what is being asked.

30. A 12 V, 60 W lamp is connected to the secondary of a step down transformer, whose primary is connected to ac mains of 220 V. Assuming the transformer to be ideal, what is the current in the primary winding?

- (A) 0.37 A
- (B) 0.27 A
- (C) 2.7 A
- (D) 3.7 A

Correct Answer: (B) 0.27 A

Solution:

Step 1: Understanding the Question:

We are given the voltage and power ratings of a lamp connected to the secondary coil of an ideal transformer. The primary voltage is also given. We need to find the current in the primary coil.

Step 2: Key Formula or Approach:

For an ideal transformer, the efficiency is 100%, which means the power input to the primary coil is equal to the power output from the secondary coil.

$$P_{primary} = P_{secondary}$$

Also, the power in a circuit is given by $P = V \times I$.

So,
$$V_p I_p = V_s I_s = P_{lamp}$$
.

Step 3: Detailed Explanation:

We are given:

Power of the lamp (connected to secondary), $P_s = 60 \text{ W}$.

Voltage of the primary, $V_p = 220 \text{ V}$.

We need to find the primary current, I_p .

For an ideal transformer, $P_p = P_s$.

So, the power drawn by the primary winding is also 60 W.

Using the power formula for the primary coil:

$$P_p = V_p \times I_p$$
60 W = 220 V × I_p

Solving for I_p :

$$I_p = \frac{60}{220} = \frac{6}{22} = \frac{3}{11} \text{ A}$$

Calculating the decimal value:

$$I_p \approx 0.2727... \text{ A}$$

This is approximately 0.27 A.

Step 4: Final Answer:

The current in the primary winding is 0.27 A. This corresponds to option (B).

Quick Tip

The keyword here is "ideal transformer". This immediately tells you that Power In = Power Out. You don't need to calculate the secondary current or the turns ratio. Just equate the primary power (V_pI_p) to the secondary power (which is given as 60 W).

31. Light travels a distance x in time t_1 in air and 10x in time t_2 in another denser medium. What is the critical angle for this medium?

(A)
$$\sin^{-1}\left(\frac{10t_1}{t_2}\right)$$

(B)
$$\sin^{-1}\left(\frac{t_2}{t_1}\right)$$

(A)
$$\sin^{-1} \left(\frac{10t_1}{t_2} \right)$$

(B) $\sin^{-1} \left(\frac{t_2}{t_1} \right)$
(C) $\sin^{-1} \left(\frac{10t_2}{t_1} \right)$
(D) $\sin^{-1} \left(\frac{t_1}{10t_2} \right)$

(D)
$$\sin^{-1}\left(\frac{t_1}{10t_2}\right)$$

Correct Answer: (A) $\sin^{-1}\left(\frac{10t_1}{t_2}\right)$

Solution:

Step 1: Understanding the Question:

We are given information about the speed of light in air and a denser medium. We need to find the critical angle for the interface between the denser medium and air.

Step 2: Key Formula or Approach:

The speed of light is distance/time. The refractive index of a medium (n) is the ratio of the speed of light in vacuum/air (c) to the speed of light in the medium (v): n = c/v. The critical angle (θ_c) for light going from a denser medium (refractive index n) to a rarer medium (like air, refractive index ≈ 1) is given by Snell's law: $n \sin \theta_c = 1 \sin 90^\circ$, which simplifies to:

$$\sin \theta_c = \frac{1}{n}$$

Step 3: Detailed Explanation:

First, calculate the speed of light in air (c) and in the medium (v).

Speed in air, $c = \frac{\text{distance}}{\text{time}} = \frac{x}{t_1}$. Speed in the medium, $v = \frac{\text{distance}}{\text{time}} = \frac{10x}{t_2}$. Next, calculate the refractive index (n) of the denser medium with respect to air.

$$n = \frac{c}{v} = \frac{x/t_1}{10x/t_2} = \frac{x}{t_1} \times \frac{t_2}{10x} = \frac{t_2}{10t_1}$$

Now, find the critical angle using the formula $\sin \theta_c = 1/n$.

$$\sin \theta_c = \frac{1}{\left(\frac{t_2}{10t_1}\right)} = \frac{10t_1}{t_2}$$

Therefore, the critical angle is:

$$\theta_c = \sin^{-1}\left(\frac{10t_1}{t_2}\right)$$

Step 4: Final Answer:

The critical angle for the medium is $\sin^{-1}\left(\frac{10t_1}{t_2}\right)$. This corresponds to option (A).

Quick Tip

The critical angle only exists when light travels from a denser medium to a rarer medium. The value of $\sin \theta_c$ is always n_{rarer}/n_{denser} . In this case, it's 1/n. Since it must be a real angle, $\sin \theta_c \leq 1$, which means $n_{rarer} \leq n_{denser}$, confirming the denser-to-rarer condition.

- 32. Let a wire be suspended from the ceiling (rigid support) and stretched by a weight W attached at its free end. The longitudinal stress at any point of cross-sectional area A of the wire is:
- (A) Zero
- (B) 2W/A
- (C) W/A
- (D) W/2A

Correct Answer: (C) W/A

Solution:

Step 1: Understanding the Question:

This is a fundamental definition question. We need to find the expression for longitudinal stress in a wire under a stretching force.

Step 2: Key Formula or Approach:

Stress is defined as the internal restoring force per unit area of a body.

$$Stress = \frac{Restoring\ Force}{Area}$$

In equilibrium, the internal restoring force is equal in magnitude to the external deforming force.

Step 3: Detailed Explanation:

The wire is stretched by a weight W. This weight acts as the external deforming force (tension) on the wire.

At any cross-section of the wire, the wire develops an internal restoring force to counteract this external force. In equilibrium, the magnitude of this restoring force is equal to the tension, which is W.

The cross-sectional area of the wire is given as A.

Therefore, the longitudinal stress (σ) is:

$$\sigma = \frac{\text{Restoring Force}}{\text{Cross-sectional Area}} = \frac{W}{A}$$

(This analysis assumes the weight of the wire itself is negligible compared to W).

Step 4: Final Answer:

The longitudinal stress at any point of the wire is W/A. This corresponds to option (C).

Quick Tip

Stress and pressure have the same dimensions (Force/Area), but stress is a tensor quantity that describes internal forces in a material, while pressure is typically a scalar quantity describing external force on a surface. For a simple stretched wire, stress is just Tension/Area.

33. A Carnot engine has an efficiency of 50% when its source is at a temperature 327° C. The temperature of the sink is :

- (A) 200° C
- (B) 27° C
- (C) 15° C
- (D) 100° C

Correct Answer: (B) 27° C

Solution:

Step 1: Understanding the Question:

We are given the efficiency of a Carnot engine and the temperature of its source. We need to find the temperature of the sink.

Step 2: Key Formula or Approach:

The efficiency (η) of a Carnot engine is given by the formula:

$$\eta = 1 - \frac{T_{sink}}{T_{source}}$$

where T_{sink} and T_{source} are the absolute temperatures of the sink and source, respectively (measured in Kelvin).

Step 3: Detailed Explanation:

First, convert all given temperatures to Kelvin. The conversion is $T(K) = T(^{\circ}C) + 273$.

Source temperature, $T_{source} = 327^{\circ}\text{C} = 327 + 273 = 600 \text{ K}.$

Efficiency, $\eta = 50\% = 0.5$.

Now, substitute these values into the efficiency formula:

$$0.5 = 1 - \frac{T_{sink}}{600}$$

Rearrange the equation to solve for T_{sink} :

$$\frac{T_{sink}}{600} = 1 - 0.5 = 0.5$$

$$T_{sink} = 0.5 \times 600 = 300 \text{ K}$$

The question asks for the temperature of the sink in degrees Celsius. Convert the result back to Celsius:

$$T_{sink}(^{\circ}C) = T_{sink}(K) - 273 = 300 - 273 = 27^{\circ}C$$

Step 4: Final Answer:

The temperature of the sink is 27° C. This corresponds to option (B).

Quick Tip

A very common mistake in thermodynamics problems is forgetting to convert temperatures from Celsius to Kelvin before using them in formulas like the Carnot efficiency or ideal gas law. Always convert to Kelvin first, do the calculation, and then convert back if the answer is required in Celsius.

34. A metal wire has mass (0.4 ± 0.002) g, radius (0.3 ± 0.001) mm and length (5 ± 0.02) cm. The maximum possible percentage error in the measurement of density will nearly be:

- (A) 1.4%
- (B) 1.2%
- (C) 1.3%
- (D) 1.6%

Correct Answer: (D) 1.6%

Solution:

Step 1: Understanding the Question:

We are given the measured values and absolute errors for the mass, radius, and length of a wire. We need to calculate the maximum possible percentage error in the density.

Step 2: Key Formula or Approach:

Density (ρ) is mass (m) divided by volume (V). For a cylindrical wire, the volume is $V = \pi r^2 l$. So, $\rho = \frac{m}{\pi r^2 l}$.

For a quantity $X = \frac{A^a B^b}{C^c}$, the maximum fractional error is given by $\frac{\Delta X}{X} = a \frac{\Delta A}{A} + b \frac{\Delta B}{B} + c \frac{\Delta C}{C}$. Applying this to the density formula, the maximum percentage error is:

$$\frac{\Delta \rho}{\rho} \times 100\% = \left(\frac{\Delta m}{m} + 2\frac{\Delta r}{r} + \frac{\Delta l}{l}\right) \times 100\%$$

Step 3: Detailed Explanation:

Calculate the individual relative errors first:

- For mass (m): m = 0.4 g, $\Delta m = 0.002$ g.

$$\frac{\Delta m}{m} = \frac{0.002}{0.4} = \frac{2}{400} = 0.005$$

- For radius (r): $r = 0.3 \text{ mm}, \Delta r = 0.001 \text{ mm}.$

$$\frac{\Delta r}{r} = \frac{0.001}{0.3} = \frac{1}{300} \approx 0.00333$$

- For length (l): l = 5 cm, $\Delta l = 0.02$ cm.

$$\frac{\Delta l}{l} = \frac{0.02}{5} = \frac{2}{500} = 0.004$$

Now, substitute these into the error formula:

$$\frac{\Delta \rho}{\rho} = \frac{\Delta m}{m} + 2\frac{\Delta r}{r} + \frac{\Delta l}{l}$$
$$\frac{\Delta \rho}{\rho} = 0.005 + 2 \times \left(\frac{1}{300}\right) + 0.004$$
$$\frac{\Delta \rho}{\rho} = 0.005 + 0.00667 + 0.004 = 0.01567$$

To get the percentage error, multiply by 100:

Percentage Error =
$$0.01567 \times 100\% = 1.567\%$$

This value is nearly 1.6%.

Step 4: Final Answer:

The maximum possible percentage error in the measurement of density is nearly 1.6%. This corresponds to option (D).

Quick Tip

When calculating percentage error, remember to multiply the relative error of each quantity by the power to which it is raised in the formula. For density of a cylinder $(\rho = m/(\pi r^2 l))$, the power of r is 2, so its relative error is multiplied by 2.

35. In a plane electromagnetic wave travelling in free space, the electric field component oscillates sinusoidally at a frequency of 2.0×10^{10} Hz and amplitude 48 Vm⁻¹. Then the amplitude of oscillating magnetic field is : (Speed of light in free space = 3×10^8 m s⁻¹)

- (A) 1.6×10^{-6} T
- (B) 1.6×10^{-9} T
- (C) 1.6×10⁻⁸T
- (D) 1.6×10^{-7} T

Correct Answer: (D) 1.6×10^{-7} T

Solution:

Step 1: Understanding the Question:

We are given the amplitude of the electric field component of an electromagnetic wave in free space. We need to find the amplitude of the magnetic field component.

Step 2: Key Formula or Approach:

For an electromagnetic wave travelling in a vacuum (free space), the ratio of the amplitudes of the electric field (E_0) and the magnetic field (B_0) is equal to the speed of light (c).

$$\frac{E_0}{B_0} = c$$

Step 3: Detailed Explanation:

We are given the following values:

Amplitude of the electric field, $E_0 = 48 \text{ V/m}$.

Speed of light in free space, $c = 3 \times 10^8$ m/s.

The frequency of oscillation is also given, but it is not needed to find the magnetic field amplitude.

Rearranging the formula to solve for the magnetic field amplitude B_0 :

$$B_0 = \frac{E_0}{c}$$

Substitute the given values:

$$B_0 = \frac{48 \text{ V/m}}{3 \times 10^8 \text{ m/s}}$$

$$B_0 = 16 \times 10^{-8} \text{ T}$$

To match the format of the options, we can write this in standard scientific notation:

$$B_0 = 1.6 \times 10^{-7} \text{ T}$$

Step 4: Final Answer:

The amplitude of the oscillating magnetic field is 1.6×10^{-7} T. This corresponds to option (D).

Quick Tip

A simple way to remember the relationship is E=cB. The frequency information is extraneous for this specific question and is sometimes included to test if you can identify the relevant data.

36. A bullet from a gun is fired on a rectangular wooden block with velocity u. When bullet travels 24 cm through the block along its length horizontally, velocity of bullet becomes u/3. Then it further penetrates into the block in the same direction before coming to rest exactly at the other end of the block. The total

length of the block is:

- (A) 30 cm
- (B) 27 cm
- (C) 24 cm
- (D) 28 cm

Correct Answer: (B) 27 cm

Solution:

Step 1: Understanding the Question:

A bullet enters a wooden block and decelerates. We are given its initial velocity, its velocity after traveling a certain distance, and that it eventually comes to rest at the other end. We need to find the total length of the block. We assume the block offers constant retardation (constant negative acceleration).

Step 2: Key Formula or Approach:

We can use the third equation of motion, $v^2 = u^2 + 2as$, as it relates initial velocity (u), final velocity (v), acceleration (a), and distance (s). Alternatively, the work-energy theorem can be used.

Using the Work-Energy Theorem (easier method):

Work done by the retarding force = Change in Kinetic Energy. Let the constant retarding force be F.

Step 3: Detailed Explanation:

Part 1: Bullet travels the first 24 cm.

- Initial velocity = u.
- Final velocity = u/3.
- Distance, $s_1 = 24$ cm.
- Work done by the retarding force $= -F \times s_1$. Change in KE $= \frac{1}{2}m(\frac{u}{3})^2 \frac{1}{2}mu^2 = \frac{1}{2}m(\frac{u^2}{9} u^2) = -\frac{1}{2}m\frac{8u^2}{9}$. So, $-F \times 24 = -\frac{4mu^2}{9} \implies F \times 24 = \frac{4mu^2}{9}$ (Equation 1)

Part 2: Bullet travels the remaining distance until it stops.

- Initial velocity = u/3.
- Final velocity = 0.
- Let the remaining distance be s_2 .

- Work done by the retarding force $= -F \times s_2$. Change in KE $= 0 \frac{1}{2}m(\frac{u}{3})^2 = -\frac{1}{2}m\frac{u^2}{9}$. So, $-F \times s_2 = -\frac{mu^2}{18} \implies F \times s_2 = \frac{mu^2}{18}$ (Equation 2)

Solving for s_2 :

Divide Equation 1 by Equation 2:

$$\frac{F \times 24}{F \times s_2} = \frac{4mu^2/9}{mu^2/18}$$

$$\frac{24}{s_2} = \frac{4}{9} \times 18 = 8$$

$$s_2 = \frac{24}{8} = 3 \text{ cm}$$

Total Length of the Block:

Total length = $s_1 + s_2 = 24 \text{ cm} + 3 \text{ cm} = 27 \text{ cm}$.

Step 4: Final Answer:

The total length of the wooden block is 27 cm. This corresponds to option (B).

Quick Tip

For problems involving changes in speed over distances with constant force, the workenergy theorem $(W_{net} = \Delta KE)$ is often more direct than using kinematic equations, as it avoids calculating acceleration explicitly.

37. In the figure shown here, what is the equivalent focal length of the combination of lenses (Assume that all layers are thin)?

$$n_1 = 1.5$$

$$R_1 = R_2 = 20 \text{ cm}$$

$$R_2 = 1.6$$

- (A) 50 cm
- (B) 40 cm
- (C) -40 cm
- (D) -100 cm

Correct Answer: (D) -100 cm

Solution:

Step 1: Understanding the Question:

The setup consists of three thin lenses in contact, forming a composite lens. We need to find the equivalent focal length of this combination.

46

Lens 1: Plano-convex lens $(n_1 = 1.5, R_1 = 20 \text{ cm}, R_2 = \infty)$.

Lens 2: Biconcave lens $(n_2 = 1.6, R_1 = -20 \text{ cm}, R_2 = +20 \text{ cm}).$

Lens 3: Plano-convex lens $(n_1 = 1.5, R_1 = \infty, R_2 = -20 \text{ cm}).$

Step 2: Key Formula or Approach:

We will use the Lens Maker's Formula for each lens and then the formula for the combination of thin lenses in contact.

Lens Maker's Formula: $\frac{1}{f} = (n-1)\left(\frac{1}{R_1} - \frac{1}{R_2}\right)$.

Combination Formula: $\frac{1}{f_{eq}} = \frac{1}{f_1} + \frac{1}{f_2} + \frac{1}{f_3}$.

Step 3: Detailed Explanation:

Focal length of Lens 1 (f_1) :

$$\frac{1}{f_1} = (1.5 - 1) \left(\frac{1}{+20} - \frac{1}{\infty} \right) = 0.5 \times \frac{1}{20} = \frac{1}{40} \implies f_1 = +40 \text{ cm}$$

Focal length of Lens 2 (f_2) :

$$\frac{1}{f_2} = (1.6 - 1) \left(\frac{1}{-20} - \frac{1}{+20} \right) = 0.6 \times \left(-\frac{2}{20} \right) = 0.6 \times \left(-\frac{1}{10} \right) = -0.06 = -\frac{6}{100} = -\frac{3}{50}$$

Focal length of Lens 3 (f_3) :

This lens is identical to Lens 1 but flipped.

$$\frac{1}{f_3} = (1.5 - 1) \left(\frac{1}{\infty} - \frac{1}{-20} \right) = 0.5 \times \frac{1}{20} = \frac{1}{40} \implies f_3 = +40 \text{ cm}$$

Equivalent Focal Length (f_{eq}) :

$$\frac{1}{f_{eq}} = \frac{1}{f_1} + \frac{1}{f_2} + \frac{1}{f_3} = \frac{1}{40} - \frac{3}{50} + \frac{1}{40}$$
$$\frac{1}{f_{eq}} = \frac{2}{40} - \frac{3}{50} = \frac{1}{20} - \frac{3}{50}$$
$$\frac{1}{f_{eq}} = \frac{5 - 6}{100} = -\frac{1}{100}$$
$$f_{eq} = -100 \text{ cm}$$

Step 4: Final Answer:

The equivalent focal length of the combination is -100 cm. This corresponds to option (D).

Quick Tip

Pay close attention to the sign convention for radii of curvature. For light incident from the left: surfaces curved to the right have positive R, and surfaces curved to the left have negative R. A plane surface has $R = \infty$.

38. A horizontal bridge is built across a river. A student standing on the bridge throws a small ball vertically upwards with a velocity 4 m s^{-1} . The ball strikes the water surface after 4 s. The height of bridge above water surface is (Take g=10 m s⁻²):

- (A) 68 m
- (B) 56 m
- (C) 60 m
- (D) 64 m

Correct Answer: (D) 64 m

Solution:

Step 1: Understanding the Question:

This is a one-dimensional kinematics problem under constant acceleration (gravity). We need to find the displacement of the ball from the bridge to the water surface, which gives the height of the bridge.

Step 2: Key Formula or Approach:

We use the second equation of motion, which relates displacement (s), initial velocity (u), time (t), and acceleration (a):

$$s = ut + \frac{1}{2}at^2$$

We will set up a coordinate system with the origin at the bridge and the upward direction as positive.

Step 3: Detailed Explanation:

Let the point from where the ball is thrown be the origin (s=0).

- Initial velocity, u = +4 m/s (upwards is positive).
- Acceleration, $a=-g=-10~\mathrm{m/s^2}$ (gravity acts downwards).
- Time of flight, t = 4 s.

We need to find the final position (displacement) 's' of the ball when it hits the water.

$$s = (4)(4) + \frac{1}{2}(-10)(4)^{2}$$
$$s = 16 + (-5)(16)$$
$$s = 16 - 80$$
$$s = -64 \text{ m}$$

The negative sign indicates that the final position is 64 m below the initial position (the bridge). Therefore, the height of the bridge above the water is 64 m.

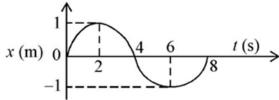
Step 4: Final Answer:

The height of the bridge above the water surface is 64 m. This corresponds to option (D).

Quick Tip

Displacement is a vector quantity. A clear choice of coordinate system and consistent use of signs for velocity, acceleration, and displacement is crucial for solving kinematics problems correctly. Here, the net displacement is negative, which directly gives the height of the bridge.

39. The x-t graph of a particle performing simple harmonic motion is shown in the figure. The acceleration of the particle at t = 2 s is:



(A)
$$\frac{\pi^2}{16}$$
 m s⁻²

(B)
$$\frac{\pi^2}{8}$$
 m s⁻²

(C)
$$-\frac{\pi^2}{8}$$
 m s⁻²

(A)
$$\frac{\pi^2}{16}$$
 m s⁻²
(B) $\frac{\pi^2}{8}$ m s⁻²
(C) $-\frac{\pi^2}{8}$ m s⁻²
(D) $-\frac{\pi^2}{16}$ m s⁻²

Correct Answer: (D) $-\frac{\pi^2}{16}$ m s⁻²

Solution:

Step 1: Understanding the Question:

We are given the position-time (x-t) graph for a particle in SHM and asked to find its acceleration at a specific time, t = 2 s.

Step 2: Key Formula or Approach:

The acceleration (a) in SHM is related to the displacement (x) by the formula:

$$a = -\omega^2 x$$

where ω is the angular frequency. We can find x from the graph and ω from the time period T, using $\omega = 2\pi/T$.

Step 3: Detailed Explanation:

- 1. Determine parameters from the graph:
- Amplitude (A): The maximum displacement is 1 m. So, A = 1 m.
- Time Period (T): The graph shows one complete oscillation takes 8 seconds (from 0 to 8 s). So, T = 8 s.
- Position at t = 2 s: From the graph, at t = 2 s, the particle is at its maximum positive displacement, so x(t=2) = +1 m.

49

2. Calculate Angular Frequency (ω):

$$\omega = \frac{2\pi}{T} = \frac{2\pi}{8} = \frac{\pi}{4} \text{ rad/s}$$

3. Calculate Acceleration at t = 2 s:

Using the formula $a = -\omega^2 x$:

$$a(t=2) = -\omega^2 \times x(t=2)$$

$$a(t=2) = -\left(\frac{\pi}{4}\right)^2 \times (1)$$

$$a(t=2) = -\frac{\pi^2}{16} \text{ m s}^{-2}$$

Step 4: Final Answer:

The acceleration of the particle at t = 2 s is $-\frac{\pi^2}{16}$ m s⁻². This corresponds to option (D).

Quick Tip

A key concept in SHM is that the restoring force, and hence acceleration, is always directed towards the equilibrium position and is opposite to the displacement. So, when displacement is positive, acceleration is negative, and vice versa.

- 40. A satellite is orbiting just above the surface of the earth with period T. If d is the density of the earth and G is the universal constant of gravitation, the quantity $\frac{3\pi}{Gd}$ represents:
- (A) \sqrt{T}
- (B) T
- (C) T^2
- (D) T^{3}

Correct Answer: (C) T^2

Solution:

Step 1: Understanding the Question:

The question asks us to relate the given quantity $\frac{3\pi}{Gd}$ to the time period T of a satellite orbiting just above the Earth's surface. This involves using the formula for the orbital period (Kepler's Third Law) and the definition of density.

Step 2: Key Formula or Approach:

- 1. The gravitational force provides the necessary centripetal force for the satellite's orbit: $\frac{GM_em}{R_e^2} = m\omega^2 R_e$, where M_e and R_e are the mass and radius of the Earth.
- 2. The orbital period is $T = \frac{2\pi}{\omega}$.

3. The density of the Earth is $d = \frac{M_e}{V_e} = \frac{M_e}{\frac{4}{3}\pi R_e^3}$.

Step 3: Detailed Explanation:

From the force equation, we get $\omega^2 = \frac{GM_e}{R_e^3}$.

Substituting $\omega = \frac{2\pi}{T}$, we have:

$$\left(\frac{2\pi}{T}\right)^2 = \frac{GM_e}{R_e^3} \implies \frac{4\pi^2}{T^2} = \frac{GM_e}{R_e^3}$$

Rearranging for T^2 , we get Kepler's Third Law:

$$T^2 = \frac{4\pi^2 R_e^3}{GM_e}$$

Now, express the mass of the Earth M_e in terms of its density d:

$$M_e = d \times V_e = d \times \frac{4}{3}\pi R_e^3$$

Substitute this expression for M_e into the equation for T^2 :

$$T^{2} = \frac{4\pi^{2}R_{e}^{3}}{G\left(d \cdot \frac{4}{3}\pi R_{e}^{3}\right)}$$

Cancel the common terms $(4, \pi, R_e^3)$:

$$T^2 = \frac{\pi}{\frac{1}{3}Gd} = \frac{3\pi}{Gd}$$

Thus, the quantity $\frac{3\pi}{Gd}$ is equal to T^2 .

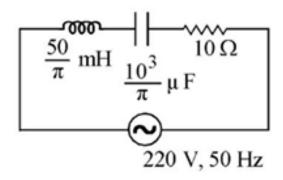
Step 4: Final Answer:

The quantity $\frac{3\pi}{Gd}$ represents T^2 . This corresponds to option (C).

Quick Tip

This derivation shows that for an object orbiting just above a spherical body, the square of the orbital period is inversely proportional to the density of the body. This is a useful result to remember.

41. The net impedance of circuit (as shown in figure) will be:



- (A) 25Ω
- (B) $10\sqrt{2} \Omega$
- (C) 15 Ω
- (D) $5\sqrt{5} \Omega$

Correct Answer: (D) $5\sqrt{5} \Omega$

Solution:

Step 1: Understanding the Question:

We need to calculate the total impedance (Z) of a series LCR circuit with given values for R, L, C, and the source frequency.

Step 2: Key Formula or Approach:

The impedance of a series LCR circuit is given by:

$$Z = \sqrt{R^2 + (X_L - X_C)^2}$$

where $X_L = \omega L = 2\pi f L$ is the inductive reactance and $X_C = \frac{1}{\omega C} = \frac{1}{2\pi f C}$ is the capacitive reactance.

Step 3: Detailed Explanation:

Given values:

$$R = 10 \,\Omega, f = 50 \,\text{Hz}.$$

 $L = \frac{50}{\pi} \,\text{mH} = \frac{50}{\pi} \times 10^{-3} \,\text{H}.$
 $C = \frac{10^3}{\pi} \,\mu\text{F} = \frac{10^3}{\pi} \times 10^{-6} \,\text{F} = \frac{10^{-3}}{\pi} \,\text{F}.$

First, calculate the angular frequency ω :

$$\omega = 2\pi f = 2\pi (50) = 100\pi \text{ rad/s}$$

Next, calculate the reactances:

$$X_L = \omega L = (100\pi) \left(\frac{50}{\pi} \times 10^{-3}\right) = 5000 \times 10^{-3} = 5 \Omega$$
$$X_C = \frac{1}{\omega C} = \frac{1}{(100\pi) \left(\frac{10^{-3}}{\pi}\right)} = \frac{1}{100 \times 10^{-3}} = \frac{1}{0.1} = 10 \Omega$$

Now, calculate the impedance Z:

$$Z = \sqrt{R^2 + (X_L - X_C)^2} = \sqrt{(10)^2 + (5 - 10)^2}$$
$$Z = \sqrt{100 + (-5)^2} = \sqrt{100 + 25} = \sqrt{125}$$
$$Z = \sqrt{25 \times 5} = 5\sqrt{5}\Omega$$

Step 4: Final Answer:

The net impedance of the circuit is $5\sqrt{5} \Omega$. This corresponds to option (D).

Quick Tip

In LCR circuits, always calculate X_L and X_C separately before plugging them into the impedance formula. Be careful with units, especially prefixes like milli (m) and micro (μ). The voltage of the source (220 V) is not needed to calculate impedance.

42. The radius of inner most orbit of hydrogen atom is 5.3×10^{-11} m. What is the radius of third allowed orbit of hydrogen atom?

- (A) 4.77 Å
- (B) 0.53 Å
- (C) 1.06 Å
- (D) 1.59 Å

Correct Answer: (A) 4.77 Å

Solution:

Step 1: Understanding the Question:

This question applies the Bohr model for the hydrogen atom. We are given the radius of the first orbit (ground state) and asked to find the radius of the third orbit.

Step 2: Key Formula or Approach:

According to the Bohr model, the radius of the n-th allowed orbit (r_n) for a hydrogen-like atom is given by:

$$r_n = r_1 \times \frac{n^2}{Z}$$

For the hydrogen atom, the atomic number Z=1. So, the formula simplifies to:

$$r_n = r_1 \times n^2$$

53

where r_1 is the radius of the first orbit (the Bohr radius).

Step 3: Detailed Explanation:

Given values:

- Radius of the innermost orbit (n=1), $r_1 = 5.3 \times 10^{-11}$ m.
- We need to find the radius of the third orbit, so n=3.

Substitute the values into the formula:

$$r_3 = r_1 \times (3)^2 = r_1 \times 9$$

 $r_3 = (5.3 \times 10^{-11} \text{ m}) \times 9 = 47.7 \times 10^{-11} \text{ m}$

The options are in Angstroms (Å). We know that $1 \text{ Å} = 10^{-10} \text{ m}$. Converting the result to Angstroms:

$$r_3 = 4.77 \times 10^{-10} \text{ m} = 4.77 \text{ Å}$$

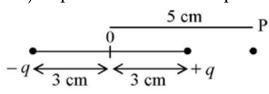
Step 4: Final Answer:

The radius of the third allowed orbit of the hydrogen atom is 4.77 Å. This corresponds to option (A).

Quick Tip

Remember the scaling laws in the Bohr model for hydrogen (Z=1): - Radius: $r_n \propto n^2$ - Energy: $E_n \propto -1/n^2$ - Velocity: $v_n \propto 1/n$ These proportionalities are very useful for ratio-based problems.

43. An electric dipole is placed as shown in the figure. The electric potential (in 10² V) at point P due to the dipole is $(\epsilon_0$ =permittivity of free space and $\frac{1}{4\pi\epsilon_0} = K$):



- (A) $(\frac{8}{3})qK$ (B) $(\frac{3}{8})qK$ (C) $(\frac{5}{8})qK$
- (D) $(\frac{8}{5})qK$

Correct Answer: (B) $(\frac{3}{8})qK$

Solution:

Step 1: Understanding the Question:

We have an electric dipole consisting of charges -q and +q. We need to find the net electric potential at a point P located on the axis of the dipole.

54

Step 2: Key Formula or Approach:

The electric potential is a scalar quantity. The total potential at a point due to a system of charges is the algebraic sum of the potentials due to individual charges.

The potential V at a distance r from a point charge Q is given by:

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

Step 3: Detailed Explanation:

The dipole consists of:

- Charge $Q_1 = -q$ at x = -3 cm.
- Charge $Q_2 = +q$ at x = +3 cm.

The point P is located at a distance of 5 cm from the origin along the positive x-axis.

Let's find the distance of point P from each charge:

- Distance from +q to P, $r_2 = 5 \text{ cm} 3 \text{ cm} = 2 \text{ cm}$.
- Distance from -q to P, $r_1 = 5$ cm -(-3 cm) = 8 cm.

Now, calculate the total potential V_P at point P:

$$V_P = V_1 + V_2 = \frac{KQ_1}{r_1} + \frac{KQ_2}{r_2}$$

$$V_P = \frac{K(-q)}{r_1} + \frac{K(+q)}{r_2} = Kq\left(\frac{1}{r_2} - \frac{1}{r_1}\right)$$

We can use the distances in cm since we are looking for an expression that is a ratio.

$$V_P = Kq\left(\frac{1}{2} - \frac{1}{8}\right)$$

$$V_P = Kq\left(\frac{4-1}{8}\right) = Kq\left(\frac{3}{8}\right)$$

Step 4: Final Answer:

The electric potential at point P is $(\frac{3}{8})qK$. This corresponds to option (B).

Quick Tip

For potential calculations, always remember it's a scalar sum. Pay careful attention to the signs of the charges. The distances must be calculated from the location of each charge to the point of interest.

44. A wire carrying a current I along the positive x-axis has length L. It is kept in a magnetic field $\vec{B} = (2\hat{i} + 3\hat{j} - 4\hat{k})$ T. The magnitude of the magnetic force acting on the wire is :

55

- (A) $\sqrt{3}$ IL
- (B) 3 IL
- (C) $\sqrt{5}$ IL

(D) 5 IL

Correct Answer: (D) 5 IL

Solution:

Step 1: Understanding the Question:

The question asks for the magnitude of the magnetic force on a straight current-carrying wire of length L placed in a uniform magnetic field.

Step 2: Key Formula or Approach:

The magnetic force \vec{F} on a wire of length vector \vec{L} carrying current I in a magnetic field \vec{B} is given by the formula:

$$\vec{F} = I(\vec{L} \times \vec{B})$$

The magnitude of the force is $|\vec{F}|$.

Step 3: Detailed Explanation:

The wire carries current I along the positive x-axis and has length L. So, the length vector \vec{L} can be written as:

$$\vec{L} = L\hat{i}$$

The magnetic field is given as:

$$\vec{B} = (2\hat{i} + 3\hat{j} - 4\hat{k}) \mathrm{T}$$

Now, we calculate the cross product $\vec{L} \times \vec{B}$:

$$\vec{L} \times \vec{B} = (\hat{L}\hat{i}) \times (2\hat{i} + 3\hat{j} - 4\hat{k})$$

$$\vec{L} \times \vec{B} = L(\hat{i} \times 2\hat{i} + \hat{i} \times 3\hat{j} - \hat{i} \times 4\hat{k})$$

Using the properties of cross products of unit vectors $(\hat{i} \times \hat{i} = 0, \hat{i} \times \hat{j} = \hat{k}, \hat{i} \times \hat{k} = -\hat{j})$:

$$\vec{L} \times \vec{B} = L(0 + 3\hat{k} - 4(-\hat{j}))$$

$$\vec{L} \times \vec{B} = L(4\hat{j} + 3\hat{k})$$

Now, substitute this into the force formula:

$$\vec{F} = I(L(4\hat{j} + 3\hat{k})) = IL(4\hat{j} + 3\hat{k})$$

The magnitude of the force is:

$$|\vec{F}| = \sqrt{(4IL)^2 + (3IL)^2}$$

$$|\vec{F}| = \sqrt{16I^2L^2 + 9I^2L^2} = \sqrt{25I^2L^2}$$

$$|\vec{F}| = 5IL$$

Step 4: Final Answer:

The magnitude of the magnetic force acting on the wire is 5 IL.

Quick Tip

When finding the magnetic force on a straight wire, remember that only the components of the magnetic field perpendicular to the wire contribute to the force. Here, the wire is along the x-axis, so the j and k components of \vec{B} are the perpendicular components. The magnitude of the perpendicular field is $\sqrt{3^2 + (-4)^2} = \sqrt{9 + 16} = \sqrt{25} = 5$. The force is then $F = I \cdot L \cdot B_{\perp} = I \cdot L \cdot 5 = 5IL$.

- 45. Calculate the maximum acceleration of a moving car so that a body lying on the floor of the car remains stationary. The coefficient of static friction between the body and the floor is 0.15 (g = 10 m s^{-2}).
- (A) 50 ms^{-2}
- (B) 1.2 ms^{-2}
- (C) 150 ms^{-2}
- (D) 1.5 ms^{-2}

Correct Answer: (D) 1.5 ms^{-2}

Solution:

Step 1: Understanding the Question:

We need to find the maximum acceleration of a car such that an object on its floor does not slide. This means the force causing the object to accelerate with the car is the static friction force, and we are looking for the limit where this force is at its maximum.

Step 2: Key Formula or Approach:

For the body to remain stationary relative to the car, the force accelerating it must be provided by static friction. The force required to accelerate the body is F = ma.

The maximum available static friction force is $f_{s,max} = \mu_s N$, where N is the normal force.

For the body not to slip, $F \leq f_{s,max}$. The maximum acceleration occurs when $ma_{max} = f_{s,max}$.

Step 3: Detailed Explanation:

Let 'm' be the mass of the body.

The car is accelerating horizontally. In the frame of the car, the body experiences a pseudo force, $F_{pseudo} = ma$, in the direction opposite to the acceleration.

This pseudo force is balanced by the force of static friction, f_s , acting in the direction of acceleration.

So, $ma = f_s$.

The body remains stationary as long as the static friction required is less than or equal to its

maximum value, $f_{s,max}$.

$$f_s \leq f_{s,max}$$

The normal force N on the body is equal to its weight, N = mg.

The maximum static friction is $f_{s,max} = \mu_s N = \mu_s mg$.

For the maximum acceleration a_{max} , the required force is maximum, which equals the maximum static friction.

$$ma_{max} = f_{s,max}$$

$$ma_{max} = \mu_s mg$$

Canceling 'm' from both sides:

$$a_{max} = \mu_s g$$

Given values are $\mu_s = 0.15$ and $g = 10 \,\mathrm{m/s}^2$.

$$a_{max} = 0.15 \times 10 = 1.5 \,\mathrm{m/s}^2$$

Step 4: Final Answer:

The maximum acceleration of the moving car is 1.5 m s^{-2} .

Quick Tip

This is a standard problem. The maximum possible acceleration for an object resting on a horizontal surface without slipping is simply $a_{max} = \mu_s g$. This simple formula can save you time in exams.

- 46. Two thin lenses are of same focal lengths (f), but one is convex and the other one is concave. When they are placed in contact with each other, the equivalent focal length of the combination will be:
- (A) Infinite
- (B) Zero
- (C) f/4
- (D) f/2

Correct Answer: (A) Infinite

Solution:

Step 1: Understanding the Question:

The question asks for the equivalent focal length of a combination of a convex lens and a concave lens, both having the same magnitude of focal length f, when they are in contact.

Step 2: Key Formula or Approach:

For two thin lenses in contact, the equivalent focal length F_{eq} is given by the formula:

$$\frac{1}{F_{eq}} = \frac{1}{f_1} + \frac{1}{f_2}$$

Alternatively, the power of the combination is the sum of individual powers: $P_{eq} = P_1 + P_2$.

Step 3: Detailed Explanation:

Let the focal length of the convex lens be f_1 and the concave lens be f_2 .

According to the sign convention: The focal length of a convex lens is positive, so $f_1 = +f$. The focal length of a concave lens is negative, so $f_2 = -f$.

Using the formula for the combination of lenses:

$$\frac{1}{F_{eq}} = \frac{1}{f_1} + \frac{1}{f_2}$$

Substituting the values:

$$\frac{1}{F_{eq}} = \frac{1}{+f} + \frac{1}{-f} = \frac{1}{f} - \frac{1}{f} = 0$$

So, we have $\frac{1}{F_{eq}} = 0$. This implies that the equivalent focal length F_{eq} is:

$$F_{eq} = \frac{1}{0} = \infty$$

The power of the combination is $P_{eq} = \frac{1}{F_{eq}} = 0$. A system with zero power and infinite focal length behaves like a plane glass slab.

Step 4: Final Answer:

The equivalent focal length of the combination will be infinite.

Quick Tip

Remember that power is the reciprocal of focal length (P = 1/f). The power of a convex lens is positive, and for a concave lens, it's negative. A combination of a convex and a concave lens of equal focal lengths has a net power of zero, behaving like a glass slab with infinite focal length.

47. A very long conducting wire is bent in a semi-circular shape from A to B as shown in figure. The magnetic field at point P for steady current configuration is given by:

59

- (A) $\frac{\mu_0 i}{4R} \left[1 \frac{2}{\pi} \right]$ pointed into the page
- (B) $\frac{\mu_0 i}{4R}$ pointed into the page
- (C) $\frac{\mu_0 i}{4R}$ pointed away from the page

(D) $\frac{\mu_0 i}{4R} \left[1 - \frac{2}{\pi}\right]$ pointed away from the page

Correct Answer: (D) $\frac{\mu_0 i}{4R} \left[1 - \frac{2}{\pi} \right]$ pointed away from the page

Solution:

Step 1: Understanding the Question:

The diagram shows a current configuration consisting of a very long straight wire and a semi-circular loop. We need to find the net magnetic field at the center of the semi-circle, P. We must use the principle of superposition.

Step 2: Key Formula or Approach:

The total magnetic field at P is the vector sum of the magnetic field due to the long straight wire $(B_{straight})$ and the semi-circular arc (B_{arc}) . 1. Magnetic field due to a long straight wire at a distance R: $B = \frac{\mu_0 i}{2\pi R}$.

2. Magnetic field at the center of a semi-circular arc of radius R: $B = \frac{\mu_0 i}{4R}$. The direction of the fields is determined by the right-hand thumb rule.

Step 3: Detailed Explanation:

The system can be considered as a superposition of two parts: 1. A very long straight wire carrying current i. 2. A semi-circular loop of radius R carrying current i.

Field due to the semi-circular arc (B_{arc}): The current in the arc flows from A to B (counter-clockwise). Using the right-hand thumb rule (curling fingers in the direction of current), the thumb points out of the page. The magnitude is:

$$B_{arc} = \frac{1}{2} \times \text{(Field of a full circle)} = \frac{1}{2} \times \left(\frac{\mu_0 i}{2R}\right) = \frac{\mu_0 i}{4R}$$

So, $\vec{B}_{arc} = \frac{\mu_0 i}{4R}$ (away from the page, or \odot).

Field due to the long straight wire ($B_{straight}$): The current in the straight part of the wire flows from right to left. Point P is at a perpendicular distance R from this wire. Using the right-hand thumb rule (pointing the thumb to the left), the fingers curl into the page at point P. The magnitude is:

$$B_{straight} = \frac{\mu_0 i}{2\pi R}$$

So, $\vec{B}_{straight} = \frac{\mu_0 i}{2\pi R}$ (into the page, or \otimes).

Net Magnetic Field (B_{net}): The two fields are in opposite directions. Let's take the direction 'away from the page' as positive.

$$B_{net} = B_{arc} - B_{straight}$$

$$B_{net} = \frac{\mu_0 i}{4R} - \frac{\mu_0 i}{2\pi R}$$

Factoring out the common term $\frac{\mu_0 i}{4R}$:

$$B_{net} = \frac{\mu_0 i}{4R} \left(1 - \frac{4R}{2\pi R} \right) = \frac{\mu_0 i}{4R} \left(1 - \frac{2}{\pi} \right)$$

Since $\pi \approx 3.14$, $2/\pi \approx 0.637$. Thus, $1 - 2/\pi$ is a positive value. This means the net magnetic field is in the positive direction, which we defined as 'away from the page'.

Step 4: Final Answer:

The net magnetic field at point P is $\frac{\mu_0 i}{4R} \left(1 - \frac{2}{\pi}\right)$ pointed away from the page.

Quick Tip

When dealing with complex wire shapes, break them down into standard components (straight wires, circular arcs). Calculate the magnetic field (both magnitude and direction) for each component separately at the point of interest. Then, find the vector sum to get the net field. Always be careful with directions using the right-hand rule.

48. 10 resistors, each of resistance R are connected in series to a battery of emf E and negligible internal resistance. Then those are connected in parallel to the same battery, the current is increased n times. The value of n is:

- (A) 1000
- (B) 10
- (C) 100
- (D) 1

Correct Answer: (C) 100

Solution:

Step 1: Understanding the Question:

We have two scenarios with 10 identical resistors and a battery. First, they are in series, and second, they are in parallel. We are told the current in the parallel case is 'n' times the current in the series case, and we need to find 'n'.

Step 2: Key Formula or Approach:

- 1. Equivalent resistance of N resistors in series: $R_{series} = N \times R$.
- 2. Equivalent resistance of N resistors in parallel: $R_{parallel} = R/N$.
- 3. Ohm's Law: $I = V/R_{eq}$, where V is the battery emf E.

Step 3: Detailed Explanation:

Let the number of resistors be N = 10.

Case 1: Series Connection

The equivalent resistance is:

$$R_s = 10 \times R = 10R$$

The current flowing through the circuit is I_s :

$$I_s = \frac{E}{R_s} = \frac{E}{10R}$$

Case 2: Parallel Connection

The equivalent resistance is:

$$\frac{1}{R_p} = \frac{1}{R} + \frac{1}{R} + \dots (10 \text{ times}) = \frac{10}{R}$$

$$R_p = \frac{R}{10}$$

The current flowing through the circuit is I_p :

$$I_p = \frac{E}{R_p} = \frac{E}{R/10} = \frac{10E}{R}$$

Relating the Currents

The problem states that the current is increased n times, which means $I_p = n \times I_s$. Substitute the expressions for I_p and I_s :

$$\frac{10E}{R} = n \times \left(\frac{E}{10R}\right)$$

We can cancel E and R from both sides of the equation:

$$10 = n \times \frac{1}{10}$$

Solving for n:

$$n = 10 \times 10 = 100$$

Step 4: Final Answer:

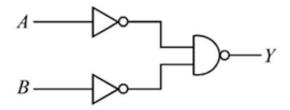
The value of n is 100.

Quick Tip

For N identical resistors, the ratio of the series resistance to parallel resistance is $R_s/R_p = (NR)/(R/N) = N^2$. Since current is inversely proportional to resistance (I = E/R), the ratio of currents will be the inverse of the resistance ratio: $I_p/I_s = R_s/R_p = N^2$. For N = 10, the current increases by $10^2 = 100$ times.

62

49. For the following logic circuit, the truth table is:



	A	В	Y		A	В	Y		A	В	Y		A	В	Y
	0	0	0		0	0	1		0	0	0		0	0	1
(A)	0	1	0	(B)	0	1	1	(C)	0	1	1	(D)	0	1	0
	1	0	0		1	0	1		1	0	1		1	0	1
	1	1	1		1	1	0		1	1	1		1	1	0

Correct Answer: (C) $\begin{bmatrix} A & B & Y \\ 0 & 0 & 0 \\ 0 & 1 & 1 \\ 1 & 0 & 1 \\ 1 & 1 & 1 \end{bmatrix}$

Solution:

Step 1: Understanding the Question:

We are given a logic circuit and need to determine its corresponding truth table. The circuit involves NOT gates and a NAND gate.

Step 2: Key Formula or Approach:

We need to find the Boolean expression for the output Y in terms of the inputs A and B. Then we can construct the truth table based on this expression. The Boolean operations are: - NOT gate: Output is the inverse of the input $(A' \text{ or } \bar{A})$. - NAND gate: Output is the inverse of the AND operation $((A \cdot B)')$. - De Morgan's Theorem: $(A \cdot B)' = A' + B'$ and $(A + B)' = A' \cdot B'$.

Step 3: Detailed Explanation:

- 1. Input A goes through a NOT gate. The output of this gate is A'.
- 2. Input B goes through a NOT gate. The output of this gate is B'.
- 3. The outputs A' and B' become the inputs for the NAND gate.
- 4. The output Y of the NAND gate is the NAND of its inputs, so $Y = (A' \cdot B')'$.
- 5. Now, we simplify this Boolean expression using De Morgan's theorem:

$$Y = (A' \cdot B')'$$

Applying the theorem $(X \cdot Y)' = X' + Y'$, where X = A' and Y = B':

$$Y = (A')' + (B')'$$

The double negation of a variable is the variable itself ((X')' = X).

$$Y = A + B$$

This is the Boolean expression for an OR gate.

6. Let's construct the truth table for Y=A+B: - When A=0, B=0: Y=0+0=0 - When A=0, B=1: Y=0+1=1 - When A=1, B=0: Y=1+0=1 - When A=1, B=1: Y=1+1=1 The resulting truth table is:

\mathbf{A}	В	Y
0	0	0
0	1	1
1	0	1
1	1	1

63

This table matches option (C).

Step 4: Final Answer:

The logic circuit is equivalent to an OR gate, and its truth table is the one shown in option (C).

Quick Tip

This specific gate configuration is known as a "bubbled AND" gate, which is logically equivalent to an OR gate according to De Morgan's laws. Recognizing these equivalences can speed up problem-solving. A NAND gate with inverted inputs behaves as an OR gate.

50. The resistance of platinum wire at 0° C is 2Ω and 6.8Ω at 80° C. The temperature coefficient of resistance of the wire is:

- (A) $3 \times 10^{-1} \, ^{\circ}\text{C}^{-1}$
- (B) $3 \times 10^{-4} \, ^{\circ}\text{C}^{-1}$
- $(C) 3 \times 10^{-3} \circ C^{-1}$
- (D) $3 \times 10^{-2} \, ^{\circ}\text{C}^{-1}$

Correct Answer: (D) 3×10^{-2} °C⁻¹

Solution:

Step 1: Understanding the Question:

We are given the resistance of a platinum wire at two different temperatures (0°C and 80°C) and asked to calculate its temperature coefficient of resistance, α .

Step 2: Key Formula or Approach:

The relationship between resistance and temperature is given by the formula:

$$R_T = R_0(1 + \alpha \Delta T)$$

where: - R_T is the resistance at temperature T. - R_0 is the resistance at the reference temperature (0°C in this case). - α is the temperature coefficient of resistance. - ΔT is the change in temperature $(T - T_0)$.

Step 3: Detailed Explanation:

From the problem statement, we have: - Resistance at $T_0 = 0$ °C is $R_0 = 2 \Omega$. - Resistance at T = 80°C is $R_T = 6.8 \Omega$. The change in temperature is:

$$\Delta T = T - T_0 = 80^{\circ} \text{C} - 0^{\circ} \text{C} = 80^{\circ} \text{C}$$

We need to find α . Let's rearrange the formula to solve for α :

$$R_T = R_0 + R_0 \alpha \Delta T$$

$$R_T - R_0 = R_0 \alpha \Delta T$$

$$\alpha = \frac{R_T - R_0}{R_0 \Delta T}$$

Now, substitute the given values into the equation:

$$\alpha = \frac{6.8 - 2}{2 \times 80}$$

$$\alpha = \frac{4.8}{160}$$

To simplify the calculation:

$$\alpha = \frac{48}{1600} = \frac{12}{400} = \frac{3}{100}$$
$$\alpha = 0.03 \, {}^{\circ}\text{C}^{-1}$$

Expressing this in scientific notation:

$$\alpha = 3 \times 10^{-2} \, ^{\circ}\mathrm{C}^{-1}$$

Step 4: Final Answer:

The temperature coefficient of resistance of the wire is 3×10^{-2} °C⁻¹.

Quick Tip

When the reference temperature is 0°C, the formula is simple: $R_T = R_0(1 + \alpha T)$. Make sure to correctly rearrange the formula to solve for α . A common mistake is to divide by R_T instead of R_0 . Remember α is defined with respect to the initial resistance R_0 .

Chemistry

- 51. In Lassaigne's extract of an organic compound, both nitrogen and sulphur are present, which gives blood red colour with Fe^{3+} due to the formation of -
- (A) $[Fe(SCN)]^{2+}$
- (B) $Fe_4[Fe(CN)_6]_3.xH_2O$
- (C) NaSCN
- (D) $[Fe(CN)_5NOS]^{4-}$

Correct Answer: $(A) [Fe(SCN)]^{2+}$

Solution:

Step 1: Understanding the Question:

This question is about the qualitative analysis of organic compounds, specifically the Lassaigne's test for the simultaneous presence of nitrogen and sulfur.

Step 2: Key Concept and Reaction:

In Lassaigne's test, an organic compound is fused with metallic sodium. If both nitrogen and sulfur are present, they react to form sodium thiocyanate (NaSCN).

$$Na + C + N + S \xrightarrow{\Delta} NaSCN$$

When the Lassaigne's extract (containing NaSCN) is treated with a neutral or slightly acidic solution of ferric chloride (FeCl₃), a blood-red coloration appears. This is due to the formation of a ferric thiocyanate complex.

Step 3: Detailed Explanation:

The reaction of thiocyanate ions (SCN⁻) from NaSCN with ferric ions (Fe³⁺) produces a complex ion, which is responsible for the characteristic blood-red color. The reaction can be represented in several ways depending on the conditions, but the simplest representation of the colored species is $[Fe(SCN)]^{2+}$.

$$Fe^{3+} + SCN^{-} \rightleftharpoons [Fe(SCN)]^{2+} (blood-red)$$

Let's analyze the other options:

- (B) Fe₄[Fe(CN)₆]₃.xH₂O is Prussian blue, which is formed in the test for nitrogen alone.
- (C) NaSCN is the substance formed during the sodium fusion, but it is colorless. The color appears only after it reacts with Fe³⁺.
- (D) $[Fe(CN)_5NOS]^{4-}$ is the complex formed in the sodium nitroprusside test for sulfur, which gives a violet color with sulfide ions.

Step 4: Final Answer:

The blood-red color is due to the formation of the ferric thiocyanate complex, represented as $[Fe(SCN)]^{2+}$. This corresponds to option (A).

Quick Tip

Remember the characteristic colors for Lassaigne's tests: - Nitrogen only: Prussian blue with $Fe^{2+} + Fe^{3+}$. - Sulfur only: Violet with sodium nitroprusside or black precipitate with lead acetate. - Nitrogen and Sulfur together: Blood-red with Fe^{3+} .

- 52. Amongst the following, the total number of species NOT having eight electrons around central atom in its outer most shell, is NH₃, AlCl₃, BeCl₂, CCl₄, PCl₅:
- (A) 1
- (B) 3
- (C) 2
- (D) 4

Correct Answer: (B) 3

Solution:

Step 1: Understanding the Question:

The question asks us to identify and count the number of molecules from the given list where the central atom does not follow the octet rule (i.e., it has fewer than or more than 8 valence electrons).

Step 2: Analyzing Each Species:

We will draw the Lewis structure for each molecule and count the valence electrons around the central atom.

- 1. NH_3 : The central atom is Nitrogen (N). It forms 3 single bonds with H and has 1 lone pair. Total electrons = $(3 \times 2) + 2 = 8$ electrons. It follows the octet rule.
- 2. AlCl₃: The central atom is Aluminum (Al). It forms 3 single bonds with Cl. Total electrons $= 3 \times 2 = 6$ electrons. This is an electron-deficient molecule with an incomplete octet.
- 3. **BeCl₂:** The central atom is Beryllium (Be). It forms 2 single bonds with Cl. Total electrons $2 \times 2 = 4$ electrons. This is an electron-deficient molecule with an incomplete octet.
- 4. **CCl₄:** The central atom is Carbon (C). It forms 4 single bonds with Cl. Total electrons = $4 \times 2 = 8$ electrons. It follows the octet rule.
- 5. PCl_5 : The central atom is Phosphorus (P). It forms 5 single bonds with Cl. Total electrons = $5 \times 2 = 10$ electrons. This molecule has an expanded octet.

Step 3: Counting the Species:

The species that do not have eight electrons around the central atom are:

- AlCl₃ (6 electrons)
- BeCl₂ (4 electrons)
- PCl₅ (10 electrons)

The total number of such species is 3.

Step 4: Final Answer:

There are 3 species (AlCl₃, BeCl₂, PCl₅) that do not have an octet around the central atom. This corresponds to option (B).

Quick Tip

The octet rule is a useful guideline but has many exceptions. Remember common examples of incomplete octets (compounds of Be, B, Al) and expanded octets (compounds of elements from the 3rd period and below, like P, S, Cl, Xe).

53. The correct order of energies of molecular orbitals of N_2 molecule, is:

(A)
$$\sigma 1s < \sigma *1s < \sigma 2s < \sigma *2s < (\pi 2p_x = \pi 2p_y) < (\pi *2p_x = \pi *2p_y) < \sigma 2p_z < \sigma *2p_z$$

(B)
$$\sigma 1s < \sigma *1s < \sigma 2s < \sigma *2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi *2p_x = \pi *2p_y) < \sigma *2p_z$$

(C)
$$\sigma 1s < \sigma *1s < \sigma 2s < \sigma *2s < \sigma 2p_z < (\pi 2p_x = \pi 2p_y) < (\pi *2p_x = \pi *2p_y) < \sigma *2p_z$$

(D)
$$\sigma 1s < \sigma *1s < \sigma 2s < \sigma *2s < \sigma 2p_z < \sigma *2p_z < (\pi 2p_x = \pi 2p_y) < (\pi *2p_x = \pi *2p_y)$$

Correct Answer: (B) $\sigma 1s < \sigma *1s < \sigma 2s < \sigma *2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi *2p_x = \pi *2p_y) < \sigma *2p_z < (\pi *2p_x = \pi *2p_y) < \sigma *2p_z$

Solution:

Step 1: Understanding the Question:

This question asks for the correct increasing order of energy for the molecular orbitals (MOs) of the dinitrogen (N_2) molecular according to Molecular Orbital Theory.

Step 2: Key Concept - MO Theory and s-p Mixing:

For diatomic molecules of the second period, the relative energies of the MOs formed from 2p atomic orbitals depend on the extent of s-p mixing.

- For lighter elements (Li₂, Be₂, B₂, C₂, N₂; i.e., up to 14 total electrons), the s-p mixing is significant. This causes the σ_{2p} orbital to be pushed to a higher energy level than the π_{2p} orbitals.
- For heavier elements (O₂, F₂, Ne₂; i.e., more than 14 electrons), the s-p mixing is less significant, and the "normal" order is observed, where σ_{2p} is lower in energy than π_{2p} .

Step 3: MO Energy Order for N_2 :

The N_2 molecule has a total of 14 electrons. Therefore, it follows the energy order with s-p mixing. The correct increasing order of energy is:

$$\sigma 1s < \sigma^* 1s < \sigma 2s < \sigma^* 2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi^* 2p_x = \pi^* 2p_y) < \sigma^* 2p_z$$

Comparing this correct sequence with the given options, we find that option (B) matches perfectly.

Option (C) shows the order for molecules like O_2 and F_2 where $\sigma 2p_z < (\pi 2p_x = \pi 2p_y)$.

Step 4: Final Answer:

The correct order of energies of molecular orbitals for the N₂ molecule is given in option (B).

Quick Tip

A simple way to remember is to check the total number of electrons. For ≤ 14 electrons (like N₂), the π bonding orbitals fill before the σ bonding orbital ($\pi < \sigma$). For > 14 electrons (like O₂), the σ bonding orbital fills first ($\sigma < \pi$). The order of antibonding orbitals ($\pi^* < \sigma^*$) is the same for all.

54. Which one is an example of heterogenous catalysis?

- (A) Combination between dinitrogen and dihydrogen to form ammonia in the presence of finely divided iron.
- (B) Oxidation of sulphur dioxide into sulphur trioxide in the presence of oxides of nitrogen.

- (C) Hydrolysis of sugar catalysed by H⁺ ions.
- (D) Decomposition of ozone in presence of nitrogen monoxide.

Correct Answer: (A) Combination between dinitrogen and dihydrogen to form ammonia in the presence of finely divided iron.

Solution:

Step 1: Understanding the Question:

The question asks to identify an example of heterogeneous catalysis from the given options.

Step 2: Defining Homogeneous and Heterogeneous Catalysis:

- Homogeneous Catalysis: The reactants and the catalyst are in the same physical phase (e.g., all are gases, or all are dissolved in the same liquid).
- **Heterogeneous Catalysis:** The reactants and the catalyst are in different physical phases (e.g., gaseous reactants and a solid catalyst).

Step 3: Analyzing the Options:

(A) Haber's process for ammonia synthesis:

$$N_2(g) + 3H_2(g) \xrightarrow{Fe(s)} 2NH_3(g)$$

Here, the reactants (N_2, H_2) are in the gaseous phase, while the catalyst (iron) is in the solid phase. Since the phases are different, this is an example of **heterogeneous catalysis**.

(B) Lead chamber process for sulfuric acid:

$$2SO_2(g) + O_2(g) \xrightarrow{NO(g)} 2SO_3(g)$$

Here, the reactants (SO_2, O_2) and the catalyst (NO) are all in the gaseous phase. This is homogeneous catalysis.

(C) Hydrolysis of sucrose:

$$C_{12}H_{22}O_{11}(aq) + H_2O(l) \xrightarrow{H^+(aq)} C_6H_{12}O_6(aq) + C_6H_{12}O_6(aq)$$

Here, the reactants and the catalyst $(H^+ ions)$ are all in the aqueous phase. This is **homogeneous catalysis**.

(D) Decomposition of ozone:

$$2O_3(g) \xrightarrow{NO(g)} 3O_2(g)$$

Here, the reactant (O_3) and the catalyst (NO) are both in the gaseous phase. This is **homogeneous catalysis**.

Step 4: Final Answer:

The only process that involves a catalyst in a different phase from the reactants is the Haber's process. This corresponds to option (A).

Quick Tip

The state symbols (g, l, s, aq) are key to distinguishing between homogeneous and heterogeneous catalysis. Look for a mismatch in the states of the reactants and the catalyst. Industrial processes often use solid catalysts for gaseous reactants because it makes separating the catalyst from the products much easier.

55. Given below are two statements:

Statement I: A unit formed by the attachment of a base to 1' position of sugar is known as nucleoside.

Statement II: When nucleoside is linked to phosphorous acid at 5'-position of sugar moiety, we get nucleotide.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (D) Statement I is true but Statement II is false.

Solution:

Step 1: Understanding the Question:

We need to evaluate the correctness of two statements related to the fundamental definitions of nucleosides and nucleotides in biochemistry.

Step 2: Analyzing Statement I:

A nucleoside is a structural subunit of nucleic acids like DNA and RNA. It is composed of a nitrogenous base (a purine or a pyrimidine) linked to a pentose sugar (ribose in RNA or de-oxyribose in DNA). The covalent bond that links the base to the sugar is called an N-glycosidic bond, and it forms at the 1' carbon of the sugar. Therefore, the definition "A unit formed by the attachment of a base to 1' position of sugar is known as nucleoside" is correct. **Statement I is true.**

Step 3: Analyzing Statement II:

A nucleotide is formed when a phosphate group is attached to a nucleoside. This attachment is a phosphoester bond, typically formed at the 5'-hydroxyl group of the sugar moiety. The statement says the nucleoside is linked to phosphorous acid (H₃PO₃). This is incorrect. The linkage is with phosphoric acid (H₃PO₄) to form a phosphate group. Therefore, "When nucleoside is linked to phosphorous acid at 5'-position of sugar moiety, we get nucleotide" is incorrect due to the use of "phosphorous acid" instead of "phosphoric acid". **Statement II is false.**

Step 4: Final Answer:

Statement I is true, and Statement II is false. This corresponds to option (D).

Quick Tip

Remember the hierarchy: Base + Sugar = Nucleoside. Nucleoside + Phosphate = Nucleotide. Pay attention to precise chemical terms. The difference between phosphoric acid (PO_4) and phosphorous acid (PO_3) is crucial here.

56. Consider the following reaction and identify the product (P). CH_3 - $CH(CH_3)$ -CH(OH)- $CH_3 \xrightarrow{HBr} Product (P)$

- (A) CH_3 - $C(CH_3)_2$ - CH_2Br
- (B) CH_3 - $C(Br)(CH_3)$ - CH_2 - CH_3
- (C) $CH_3CH=CH-CH_3$
- (D) CH_3 - $CH(CH_3)$ -CH(Br)- CH_3

Correct Answer: (B) CH₃-C(Br)(CH₃)-CH₂-CH₃

Solution:

Step 1: Understanding the Question:

The question asks for the major product of the reaction between 3-methylbutan-2-ol (a secondary alcohol) and hydrogen bromide (HBr). This is a nucleophilic substitution reaction.

Step 2: Reaction Mechanism:

Reactions of secondary alcohols with HBr typically proceed through an S_N1 mechanism, which involves the formation of a carbocation intermediate.

Step 2a: Protonation of the alcohol. The oxygen atom of the hydroxyl group gets protonated by H^+ from HBr to form a protonated alcohol, which has a good leaving group (H_2O).

$$\mathrm{CH}_3 - \mathrm{CH}(\mathrm{CH}_3) - \mathrm{CH}(\mathrm{OH}) - \mathrm{CH}_3 + \mathrm{H}^+ \rightleftharpoons \mathrm{CH}_3 - \mathrm{CH}(\mathrm{CH}_3) - \mathrm{CH}(\mathrm{OH}_2^+) - \mathrm{CH}_3$$

Step 2b: Formation of the initial carbocation. The water molecule leaves, generating a secondary carbocation.

$$\mathrm{CH}_3 - \mathrm{CH}(\mathrm{CH}_3) - \mathrm{CH}(\mathrm{OH}_2^+) - \mathrm{CH}_3 \to \mathrm{CH}_3 - \mathrm{CH}(\mathrm{CH}_3) - \mathrm{C}^+\mathrm{H} - \mathrm{CH}_3 + \mathrm{H}_2\mathrm{O}$$

This is a 2° carbocation.

Step 2c: Carbocation rearrangement. Carbocations can rearrange to form more stable carbocations. The adjacent carbon (C-3) has a hydrogen atom. If this hydride ion (H⁻) shifts from C-3 to C-2 (a 1,2-hydride shift), a more stable tertiary carbocation is formed.

$$\mathrm{CH_3} - \mathrm{CH}(\mathrm{CH_3}) - \mathrm{C^+H} - \mathrm{CH_3} \xrightarrow{1,2\text{-hydride shift}} \mathrm{CH_3} - \mathrm{C^+(CH_3)} - \mathrm{CH_2} - \mathrm{CH_3}$$

71

This is a 3° carbocation, which is more stable than the 2° carbocation. This rearrangement will be the major pathway.

Step 2d: Nucleophilic attack. The bromide ion (Br⁻) attacks the more stable tertiary carbocation to form the final product.

$$CH_3 - C^+(CH_3) - CH_2 - CH_3 + Br^- \rightarrow CH_3 - C(Br)(CH_3) - CH_2 - CH_3$$

Step 3: Identifying the Product:

The major product formed is 2-bromo-2-methylbutane. This corresponds to option (B). Option (D) would be the product formed without rearrangement, which would be the minor product.

Step 4: Final Answer:

The major product (P) is 2-bromo-2-methylbutane, formed via a carbocation rearrangement. This is option (B).

Quick Tip

57. The stability of Cu^{2+} is more than Cu^{+} salts in aqueous solution due to -

- (A) second ionisation enthalpy.
- (B) first ionisation enthalpy.
- (C) enthalpy of atomization.
- (D) hydration energy.

Correct Answer: (D) hydration energy.

Solution:

Step 1: Understanding the Question:

The question asks for the reason behind the greater stability of the cupric ion (Cu^{2+}) compared to the cuprous ion (Cu^{+}) in an aqueous medium, despite the fact that forming Cu^{2+} requires more energy (higher second ionization enthalpy).

Step 2: Key Concepts:

The stability of an ion in an aqueous solution is determined by the overall energy change when the solid salt dissolves in water. This involves lattice enthalpy and hydration enthalpy. The process can be thought of using a Born-Haber cycle. The key comparison is between the energy required to form the ion in the gaseous state (ionization enthalpy) and the energy released when the gaseous ion is dissolved in water (hydration enthalpy).

Step 3: Detailed Explanation:

- 1. **Ionization Enthalpy:** The second ionization enthalpy of copper (energy to remove the second electron, $Cu^+ \to Cu^{2+}$) is very high. Based on ionization enthalpy alone, Cu^+ should be more stable than Cu^{2+} .
- 2. **Hydration Enthalpy:** Hydration enthalpy is the energy released when one mole of gaseous ions dissolves in water to form hydrated ions. This energy depends on the charge density of the ion. Charge density is the ratio of charge to size.
- 3. Comparison: The Cu^{2+} ion has a greater positive charge (+2) and a smaller ionic radius compared to the Cu^{+} ion (+1). This results in a much higher charge density for Cu^{2+} .
- 4. Conclusion: Due to its high charge density, the Cu^{2+} ion attracts water molecules much more strongly. This leads to a very high negative hydration enthalpy (ΔH_{hyd}) . The large amount of energy released during the hydration of Cu^{2+} more than compensates for the high second ionization enthalpy required to form it. Therefore, $Cu^{2+}(aq)$ is more stable than $Cu^{+}(aq)$.

Step 4: Final Answer:

The high stability of Cu²⁺ in aqueous solution is due to its much higher hydration energy.

Quick Tip

For transition metals, the stability of different oxidation states in aqueous solution is a classic interplay between ionization enthalpy and hydration enthalpy. A higher charge on a smaller ion leads to a significantly larger (more negative) hydration enthalpy, which can often be the deciding factor for stability.

58. Taking stability as the factor, which one of the following represents correct relationship?

- (A) $TlI > TlI_3$
- (B) $TlCl_3 > TlCl$
- (C) $InI_3 > InI$
- (D) $AlCl > AlCl_3$

Correct Answer: (A) $TlI > TlI_3$

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct relationship based on the stability of compounds of Group 13 elements (Al, In, Tl). This relates to the stability of their different oxidation states.

Step 2: Key Concept - Inert Pair Effect:

The inert pair effect is the tendency of the two electrons in the outermost atomic s-orbital to

remain unshared or un-ionized in compounds of post-transition metals. This effect becomes more prominent as we move down a group. For Group 13, the common oxidation states are +3 and +1. Due to the inert pair effect, the stability of the +1 oxidation state increases down the group, while the stability of the +3 oxidation state decreases. Stability order for +3 state: $Al^{3+} > Ga^{3+} > In^{3+} > Tl^{3+}$.

Stability order for +1 state: $Al^+ < Ga^+ < In^+ < Tl^+$.

Step 3: Detailed Explanation:

Let's analyze each option based on the inert pair effect:

- (A) $TII > TII_3$: Thallium (Tl) is the heaviest element in this group. Due to the strong inert pair effect, its +1 oxidation state is much more stable than its +3 oxidation state. Therefore, TlI (Tl in +1 state) is more stable than TII_3 (implying Tl in +3 state). This statement is correct. (Note: TII_3 actually exists as $Tl^+(I_3)^-$, where Tl is still in the +1 state, but the question compares the stability of Tl compounds based on apparent oxidation states).
- (B) $TlCl_3 > TlCl$: This is incorrect. As explained above, Tl(+1) is more stable than Tl(+3). So, TlCl is more stable than $TlCl_3$.
- (C) InI₃ > InI: For Indium (In), the +3 oxidation state is more stable than the +1 oxidation state, although the +1 state is also known. So, InI₃ is generally more stable than InI. While this statement itself is correct, option (A) represents a more pronounced effect that is a key feature of the group's chemistry. However, the question asks for a correct relationship. Let's re-evaluate. The stability difference is most stark for Tl. InI is unstable and disproportionates. So InI₃ inI is a correct stability order. Why is A chosen over C? Often TlI₃ is considered as an ionic compound Tl+(I₃)- so Tl is in +1 state. So comparing TlI and TlI₃ is like comparing Tl+I- and Tl+(I₃)-. The question intends to compare the stability of +1 and +3 states. For Tl, +1 is significantly more stable.
- (D) AlCl > AlCl₃: This is incorrect. Aluminum (Al) is at the top of the group and exhibits a stable +3 oxidation state almost exclusively. AlCl₃ is very stable, while AlCl is not.

Comparing (A) and (C), the most significant and defining stability trend in this group is the high stability of Tl(+1). Therefore, the relationship $TlI > TlI_3$ (representing Tl(+1) > Tl(+3)) is the best representation of this chemical principle among the choices.

Step 4: Final Answer:

The correct relationship representing the stability based on the inert pair effect is TII > TII₃.

Quick Tip

Remember the inert pair effect: for p-block elements, as you go down the group, the stability of the lower oxidation state (Group number - 2) increases. For Group 13, this means Tl⁺ is the most stable ion among the heavier elements.

- 59. Which amongst the following molecules on polymerization produces neoprene?
- (1) $H_2C = C(CH_3) CH = CH_2$
- (2) H₂C = CH CH = CH₂

- (3) $H_2C = C(Cl) CH = CH_2$
- (4) $H_2C = CH C \equiv CH$
- (A)(1)
- (B) (2)
- (C)(3)
- (D) (4)

Correct Answer: (C) (3)

Solution:

Step 1: Understanding the Question:

The question asks to identify the monomer unit that undergoes polymerization to form neoprene.

Step 2: Key Concept - Monomers and Polymers:

Neoprene is a synthetic rubber. It is a polymer formed by the free-radical polymerization of its monomer. We need to identify the chemical structure of this monomer, which is known as chloroprene.

Step 3: Detailed Explanation:

Let's analyze the given options: 1. $H_2C = C(CH_3)$ - $CH = CH_2$: This is 2-methyl-1,3-butadiene, commonly known as **isoprene**. It is the monomer for natural rubber.

- 2. $\mathbf{H}_2\mathbf{C} = \mathbf{C}\mathbf{H} \mathbf{C}\mathbf{H} = \mathbf{C}\mathbf{H}_2$: This is **1,3-butadiene**. It is a monomer used in the production of various synthetic rubbers like SBR (styrene-butadiene rubber) and Buna-N.
- 3. $H_2C = C(Cl)$ $CH = CH_2$: This is 2-chloro-1,3-butadiene, commonly known as **chloro-prene**. The polymerization of chloroprene yields polychloroprene, which is commercially known as **neoprene**.
- 4. $\mathbf{H}_2\mathbf{C} = \mathbf{C}\mathbf{H} \mathbf{C} \equiv \mathbf{C}\mathbf{H}$: This is vinylacetylene. It is not a common monomer for producing synthetic rubber like neoprene.

The polymerization reaction for neoprene is:

$$\begin{array}{ccc} n\:({\rm H_2C{=}C(Cl)\text{-}CH{=}CH_2}) & \xrightarrow{\rm Polymerization} [{\rm -CH_2\text{-}C(Cl){=}CH{-}CH_2\text{-}}]_n \\ \\ & \text{Chloroprene} & \text{Neoprene} \: (\text{Polychloroprene}) \end{array}$$

Step 4: Final Answer:

The molecule that produces neoprene upon polymerization is 2-chloro-1,3-butadiene (chloro-prene), which corresponds to option (3).

Quick Tip

Memorizing the monomers of important polymers is crucial for competitive exams. Create a chart for common polymers like natural rubber (isoprene), neoprene (chloroprene), Buna-S (butadiene + styrene), PVC (vinyl chloride), Teflon (tetrafluoroethene), etc.

- 60. Which of the following statements are NOT correct?
- A. Hydrogen is used to reduce heavy metal oxides to metals.
- B. Heavy water is used to study reaction mechanism.
- C. Hydrogen is used to make saturated fats from oils.
- D. The H-H bond dissociation enthalpy is lowest as compared to a single bond between two atoms of any element.
- E. Hydrogen reduces oxides of metals that are more active than iron.

Choose the most appropriate answer from the options given below:

- (A) A, B, C only
- (B) B, C, D, E only
- (C) B, D only
- (D) D, E only

Correct Answer: (D) D, E only

Solution:

Step 1: Understanding the Question:

We need to evaluate five statements about hydrogen and its compounds and identify which of them are incorrect.

Step 2: Detailed Explanation of Each Statement:

- Statement A: Hydrogen is used to reduce heavy metal oxides to metals.

This is **correct**. Hydrogen is a good reducing agent and is used in metallurgy to reduce oxides of less reactive metals like copper, lead, tungsten, and molybdenum to their respective metals. For example: $CuO + H_2 \rightarrow Cu + H_2O$.

- Statement B: Heavy water is used to study reaction mechanism.

This is **correct**. Heavy water (D_2O) is used as a tracer compound to study the mechanism of chemical and biological reactions. The deuterium atom can be tracked, providing insights into reaction pathways.

- Statement C: Hydrogen is used to make saturated fats from oils.

This is **correct**. This process is called hydrogenation or "hardening" of oils. Unsaturated fats (containing C=C double bonds) in vegetable oils are reacted with hydrogen in the presence of a catalyst (like Ni, Pt, or Pd) to form saturated fats (like vanaspati ghee).

- Statement D: The H-H bond dissociation enthalpy is lowest as compared to a single bond between two atoms of any element.

This is **incorrect**. The H-H bond dissociation enthalpy is approximately 435.9 kJ/mol, which is one of the **highest** for a single bond between two atoms. The small size of the hydrogen atom allows for very effective orbital overlap, leading to a very strong covalent bond.

- Statement E: Hydrogen reduces oxides of metals that are more active than iron.

This is **incorrect**. According to the reactivity series (or Ellingham diagram), hydrogen can reduce oxides of metals that are **less** reactive (less electropositive) than iron, such as copper, tin, and lead. Oxides of highly reactive metals like sodium, potassium, calcium, and aluminum (which are more active than iron) are very stable and cannot be reduced by hydrogen.

Step 3: Identifying the Incorrect Statements:

From the analysis, statements D and E are incorrect.

Step 4: Final Answer:

The incorrect statements are D and E. Therefore, the correct option is (D).

Quick Tip

Remember the reactivity series for metallurgical processes. A reducing agent can only reduce the oxide of a metal that is less reactive than it. Hydrogen is placed between zinc/iron and copper in the reactivity series. This means it can reduce oxides of metals below it (Cu, Ag, Au) but not those above it (K, Na, Ca, Al, Zn, Fe).

61. Homoleptic complex from the following complexes is:

- (A) Triamminetriaquachromium (III) chloride
- (B) Potassium trioxalatoaluminate (III)
- (C) Diamminechloridonitrito N platinum (II)
- (D) Pentaamminecarbonatocobalt (III) chloride

Correct Answer: (B) Potassium trioxalatoaluminate (III)

Solution:

Step 1: Understanding the Question:

The question asks to identify the homoleptic complex from a given list of coordination compounds.

Step 2: Key Definitions:

- **Homoleptic Complex:** A complex in which the central metal ion is bonded to only **one type** of ligand.
- Heteroleptic Complex: A complex in which the central metal ion is bonded to more than one type of ligand.

Step 3: Detailed Explanation of Each Option:

1. **Triamminetriaquachromium (III) chloride:** The complex ion is $[Cr(NH_3)_3(H_2O)_3]^{3+}$. The ligands are ammine (NH_3) and aqua (H_2O) . Since there are two different types of ligands, this is a **heteroleptic** complex.

- 2. Potassium trioxalatoaluminate (III): The complex ion is $[Al(C_2O_4)_3]^{3-}$. The ligand is oxalato $(C_2O_4^{2-})$. Since there is only one type of ligand, this is a **homoleptic** complex. The counter ion is Potassium (K^+) .
- 3. Diamminechloridonitrito N platinum (II): The complex is $[Pt(NH_3)_2Cl(NO_2)]$. The ligands are ammine (NH_3) , chlorido (Cl^-) , and nitrito-N (NO_2^-) . Since there are three different types of ligands, this is a **heteroleptic** complex.
- 4. **Pentaamminecarbonatocobalt (III) chloride:** The complex ion is $[Co(NH_3)_5(CO_3)]^+$. The ligands are ammine (NH_3) and carbonato (CO_3^{2-}) . Since there are two different types of ligands, this is a **heteroleptic** complex.

Step 4: Final Answer:

Based on the analysis, Potassium trioxalatoaluminate (III) is the only homoleptic complex in the list.

Quick Tip

To quickly identify homoleptic vs. heteroleptic complexes, look at the name. The presence of multiple different ligand names (like 'ammine' and 'aqua' or 'ammine' and 'chlorido') immediately indicates a heteroleptic complex. A homoleptic complex will have only one type of ligand mentioned (like 'trioxalato').

- 62. The element expected to form largest ion to achieve the nearest noble gas configuration is:
- (A) Na
- (B) O
- (C) F
- (D) N

Correct Answer: (D) N

Solution:

Step 1: Understanding the Question:

We need to determine which element from the list (Na, O, F, N) will form the largest ion when it achieves a stable noble gas electron configuration.

Step 2: Key Concept - Isoelectronic Species and Ionic Radii:

Isoelectronic species are atoms or ions that have the same number of electrons. For a series of isoelectronic species, the ionic radius decreases as the nuclear charge (number of protons, Z) increases. This is because a higher nuclear charge pulls the electrons more strongly, causing the ion to shrink.

Step 3: Detailed Explanation:

Let's determine the ion each element forms to achieve a noble gas configuration: - Na (Sodium): Atomic number Z = 11. It has 11 protons and 11 electrons ([Ne] $3s^1$). It loses one electron to form Na⁺, which has 10 electrons ([Ne] configuration) and 11 protons.

- O (Oxygen): Atomic number Z = 8. It has 8 protons and 8 electrons ([He] $2s^22p^4$). It gains two electrons to form O^{2-} , which has 10 electrons ([Ne] configuration) and 8 protons.
- **F** (Fluorine): Atomic number Z = 9. It has 9 protons and 9 electrons ([He] $2s^22p^5$). It gains one electron to form F^- , which has 10 electrons ([Ne] configuration) and 9 protons.
- N (Nitrogen): Atomic number Z = 7. It has 7 protons and 7 electrons ([He] $2s^22p^3$). It gains three electrons to form N^{3-} , which has 10 electrons ([Ne] configuration) and 7 protons.

All the resulting ions $(N^{3-}, O^{2-}, F^-, Na^+)$ are isoelectronic, as they all have 10 electrons. Now we compare their sizes based on their nuclear charge (number of protons). - N^{3-} : 7 protons - O^{2-} : 8 protons - F^- : 9 protons - Na^+ : 11 protons

The ion with the fewest protons (lowest nuclear charge) will have the weakest pull on the 10 electrons, resulting in the largest ionic radius. Nitrogen has the lowest nuclear charge (Z=7). Therefore, the order of ionic radii is: $N^{3-} > O^{2-} > F^{-} > Na^{+}$.

Step 4: Final Answer:

The largest ion is N^{3-} , which is formed from the element Nitrogen (N).

Quick Tip

For isoelectronic ions, remember this simple rule: the more negative the charge (or the fewer the protons), the larger the ion. The more positive the charge (or the more the protons), the smaller the ion.

79

63. Identify the product in the following reaction:

$$\begin{array}{c} \stackrel{+}{\overset{+}{\text{N}_2}}\stackrel{-}{\text{Cl}} \\ &\stackrel{(i)}{\overset{-}{\text{Cl}}} \stackrel{(i)}{\overset{-}{\text{Cl}}} \stackrel{\text{Cu}_2\text{Br}_2/\text{HBr}}{\overset{-}{\text{(ii)}}} \stackrel{\text{Hg}}{\overset{-}{\text{Cl}}} \\ \stackrel{(ii)}{\overset{-}{\text{Hg}}} \stackrel{\text{Hg}}{\overset{-}{\text{Cl}}} \end{array} \rightarrow \text{Product}$$

(A) p-bromophenol

- (B) phenol
- (C) benzene
- (D) phenylmagnesium bromide

Correct Answer: (C) benzene

Solution:

Step 1: Understanding the Question:

The question shows a multi-step reaction starting from benzenediazonium chloride and asks for the final product. We need to analyze each step of the reaction sequence.

Step 2: Analysis of the Reaction Steps:

The reaction proceeds in three steps:

Step (i): Reaction with Cu₂Br₂/HBr

- The starting material is benzenediazonium chloride.
- This reaction is a **Sandmeyer reaction**. The diazonium group $(-N_2^+\text{Cl}^-)$ is an excellent leaving group and is replaced by a nucleophile, in this case, Br^- from HBr, using a copper(I) salt catalyst.
- The product of this step is **bromobenzene**.

$$C_6H_5N_2^+Cl^-\xrightarrow{Cu_2Br_2/HBr}C_6H_5Br+N_2$$

Step (ii): Reaction with Mg/dry ether

- Bromobenzene is treated with magnesium metal in the presence of dry ether.
- This is the standard procedure for the preparation of a **Grignard reagent**.
- The product of this step is **phenylmagnesium bromide**.

$$C_6H_5Br+Mg\xrightarrow{dry\ ether}C_6H_5MgBr$$

Step (iii): Reaction with H₂O

- The Grignard reagent, phenylmagnesium bromide, is treated with water.
- Grignard reagents are very strong bases and react readily with any source of acidic protons, such as water, alcohols, or carboxylic acids.
- The C-Mg bond is polarized ($C^{\delta-}$ -Mg^{$\delta+$}), making the phenyl group a strong nucleophile and base. It abstracts a proton (H⁺) from a water molecule.
- The final organic product is **benzene**.

$$C_6H_5MgBr + H-OH \rightarrow C_6H_6 + Mg(OH)Br$$

Step 3: Final Product:

Following the sequence of reactions, the final product is benzene.

Step 4: Final Answer:

The product of the given reaction sequence is benzene.

Quick Tip

Grignard reagents are extremely useful in synthesis but are very sensitive to moisture. Their reaction with water to form an alkane/arene is a key property. If you see a Grignard reagent followed by the addition of water (or any protic solvent), the result is almost always the protonation of the carbanionic part of the reagent.

64. Complete the following reaction:

$$(A) \longrightarrow (A) \longrightarrow (A)$$

Correct Answer: (A)

Solution:

Step 1: Understanding the Question:

This is a three-step organic synthesis problem. We need to identify the final product [C] starting from benzaldehyde.

Step 2: Analyzing the Reaction Steps:

Step I: Benzaldehyde $\xrightarrow{\text{HCN}} [A] \rightarrow [B]$

This is the nucleophilic addition of cyanide ion (from HCN) to the carbonyl group of benzaldehyde. The product [B] is a cyanohydrin, specifically mandelonitrile.

$${\rm C_6H_5CHO+HCN\rightarrow C_6H_5CH(OH)CN} \quad {\rm (Mandelonitrile,\,[B])}$$

Step II: [B]
$$\xrightarrow{\operatorname{conc. } H_2SO_4, \Delta}$$
 [C]

This step involves the acid-catalyzed hydrolysis of the nitrile group (-CN) in the cyanohydrin [B]. When a nitrile is heated with a strong acid like concentrated H_2SO_4 , it hydrolyzes to a carboxylic acid (-COOH).

$$C_6H_5CH(OH)CN + 2H_2O \xrightarrow{H^+,\Delta} C_6H_5CH(OH)COOH + NH_3$$

81

The final product [C] is mandelic acid (α -hydroxy phenylacetic acid).

Step 3: Identifying the Final Product [C]:

The structure of mandelic acid is a benzene ring attached to a carbon atom which is bonded to a hydroxyl group (-OH) and a carboxylic acid group (-COOH). This corresponds to the structure given in option (A).

Step 4: Final Answer:

The final product [C] is mandelic acid. This corresponds to option (A).

Quick Tip

Remember the reactivity of the nitrile (-CN) group. It's a versatile functional group that can be hydrolyzed to a carboxylic acid (with acid or base) or reduced to a primary amine (e.g., with LiAlH₄ or H_2/Ni). Acid hydrolysis is a key reaction to know.

65. Some tranquilizers are listed below. Which one from the following belongs to barbiturates?

- (A) Veronal
- (B) Chlordiazepoxide
- (C) Meprobamate
- (D) Valium

Correct Answer: (A) Veronal

Solution:

Step 1: Understanding the Question:

This question asks to identify which of the given tranquilizers is a member of the barbiturate class of drugs. This requires knowledge of drug classification from the topic "Chemistry in Everyday Life".

Step 2: Classifying the Given Drugs:

Let's classify each of the listed tranquilizers:

- **Veronal:** Also known as barbital, it is one of the first synthesized barbiturates. Barbiturates are derivatives of barbituric acid and act as central nervous system depressants.
- Chlordiazepoxide: This drug (marketed as Librium) is a member of the benzodiazepine class. Benzodiazepines are a different class of tranquilizers from barbiturates.
- **Meprobamate:** This is classified as a carbamate derivative. It is sometimes referred to as a non-barbiturate tranquilizer.
- Valium: The chemical name is diazepam. It is another well-known member of the benzodiazepine class of drugs.

Step 3: Identifying the Barbiturate:

Based on the classification, Veronal is the only drug in the list that belongs to the barbiturate class.

Step 4: Final Answer:

Veronal is a barbiturate. This corresponds to option (A).

Quick Tip

For the "Chemistry in Everyday Life" chapter, it's helpful to create a table classifying common drugs (e.g., tranquilizers, analgesics, antihistamines) into their major chemical classes (e.g., barbiturates, benzodiazepines, opiates) with one or two key examples for each.

66. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R:

Assertion A : In equation $\Delta G = -nFE_{cell}$, value of ΔG depends on n.

Reason R: E_{cell} is an intensive property and ΔG is an extensive property.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true and R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (B) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Analyzing Assertion A:

The assertion states that in the equation $\Delta G = -nFE_{cell}$, the value of ΔG (Gibbs Free Energy change) depends on 'n' (the number of moles of electrons transferred in the redox reaction). Looking at the equation, ΔG is directly proportional to 'n'. If we double the amount of reactants (which would double 'n'), the total energy released or absorbed (ΔG) will also double. Therefore, **Assertion A is true**.

Step 2: Analyzing Reason R:

The reason states that E_{cell} (cell potential) is an intensive property and ΔG is an extensive property.

- Intensive Property: A property that does not depend on the amount of matter. E_{cell} is an electric potential difference, which is an intensive property. The voltage of a battery is the same regardless of its size. So, this part is correct.
- Extensive Property: A property that depends on the amount of matter. Gibbs Free Energy

 (ΔG) represents the total amount of energy available to do work, which is directly proportional to the amount of substance reacting. So, ΔG is an extensive property. This part is also correct. Therefore, **Reason R is true**.

Step 3: Evaluating the Connection between A and R:

The equation is $\Delta G = -nFE_{cell}$. The reason R explains precisely why ΔG depends on 'n'. ΔG is an extensive property, meaning it must scale with the size of the system, which is represented by 'n'. On the other hand, E_{cell} is an intensive property, so its value is fixed for a given reaction under specific conditions. The factor 'n' in the equation is what scales the intensive property (E_{cell}) to produce the extensive property (ΔG) . Thus, the fact that ΔG is extensive and E_{cell} is intensive is the fundamental reason why ΔG depends on 'n'. Therefore, **R** is the correct explanation of **A**.

Step 4: Final Answer:

Both A and R are true, and R is the correct explanation for A. This corresponds to option (B).

Quick Tip

Remember the difference between intensive (e.g., temperature, pressure, density, potential) and extensive (e.g., mass, volume, energy, entropy) properties. An extensive property is the product of an intensive property and another extensive property (e.g., $Energy = Potential \times Charge$, where charge is extensive).

- 67. Intermolecular forces are forces of attraction and repulsion between interacting particles that will include :
- A. dipole dipole forces.
- B. dipole induced dipole forces.
- C. hydrogen bonding.
- D. covalent bonding.
- E. dispersion forces.

Choose the most appropriate answer from the options given below:

- (A) A, C, D, E are correct.
- (B) B, C, D, E are correct.
- (C) A, B, C, D are correct.
- (D) A, B, C, E are correct.

Correct Answer: (D) A, B, C, E are correct.

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the listed forces are classified as intermolecular forces.

Step 2: Defining Intermolecular vs. Intramolecular Forces:

- **Intermolecular forces** are the forces that exist *between* molecules. They are responsible for the physical properties of substances like boiling point and melting point. They are generally much weaker than intramolecular forces. - **Intramolecular forces** are the forces that exist *within* a molecule, holding the atoms together. Covalent, ionic, and metallic bonds are examples of intramolecular forces.

Step 3: Analyzing the Given Forces:

- A. **Dipole-dipole forces:** These are electrostatic attractions between the positive end of one polar molecule and the negative end of another. These are forces *between* molecules, so they are intermolecular.
- B. **Dipole-induced dipole forces:** These occur between a polar molecule and a nonpolar molecule, where the polar molecule induces a temporary dipole in the nonpolar one. These are forces *between* molecules, so they are intermolecular.
- C. **Hydrogen bonding:** This is a special, strong type of dipole-dipole interaction that occurs when hydrogen is bonded to a highly electronegative atom (N, O, or F). It is a force *between* molecules, so it is intermolecular.
- D. Covalent bonding: This involves the sharing of electrons between atoms to form a molecule. It is a force *within* a molecule, so it is an intramolecular force, not intermolecular.
- E. **Dispersion forces (London forces):** These are weak forces arising from temporary, induced dipoles in molecules due to the random motion of electrons. They exist between all molecules and are the only intermolecular force for nonpolar substances. They are forces *between* molecules, so they are intermolecular.

Step 4: Final Answer:

The forces A, B, C, and E are all types of intermolecular forces. Covalent bonding (D) is an intramolecular force. Therefore, the correct set is A, B, C, and E. This corresponds to option (D).

Quick Tip

Remember the main distinction: Intramolecular forces make up the molecule (like covalent bonds). Intermolecular forces make the molecules stick together to form a liquid or solid. All types of van der Waals forces (dipole-dipole, dipole-induced dipole, dispersion) and hydrogen bonds are intermolecular.

68. Match List - I with List - II:

List - I List - II

- A. Coke I. Carbon atoms are sp³ hybridised.
- B. Diamond II. Used as a dry lubricant
- C. Fullerene III. Used as a reducing agent
- D. Graphite IV. Cage like molecules

Choose the correct answer from the options given below:

(A) A-III, B-IV, C-I, D-II

- (B) A-II, B-IV, C-I, D-III
- (C) A-IV, B-I, C-II, D-III
- (D) A-III, B-I, C-IV, D-II

Correct Answer: (D) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:

We need to match the allotropes and forms of carbon in List-I with their corresponding properties or structures in List-II.

Step 2: Matching Each Item:

- A. Coke: Coke is a high-carbon content fuel derived from coal. It is widely used in metal-lurgy, particularly in blast furnaces, as a reducing agent to reduce metal oxides (like iron ore) to the metal. So, A matches with III.
- **B. Diamond:** Diamond is an allotrope of carbon where each carbon atom is covalently bonded to four other carbon atoms in a tetrahedral arrangement. This bonding involves sp³ hybridization. So, **B matches with I**.
- C. Fullerene: Fullerenes, such as C_{60} (buckminsterfullerene), are allotropes of carbon that consist of molecules forming a hollow sphere or ellipsoid. This is often described as a cage-like structure. So, C matches with IV.
- **D. Graphite:** Graphite has a layered structure. The layers are held by weak van der Waals forces and can easily slide over one another. This property makes graphite an excellent solid or dry lubricant. So, **D matches with II**.

Step 3: Compiling the Correct Match:

The correct matching is:

- $-A \rightarrow III$
- B \rightarrow I
- $C \rightarrow IV$
- D \rightarrow II

Step 4: Final Answer:

The combination A-III, B-I, C-IV, D-II is found in option (D).

Quick Tip

Associate keywords with carbon allotropes: - Diamond \to sp³, hardest, tetrahedral, insulator. - Graphite \to sp², layered, lubricant, conductor. - Fullerene \to cage-like, soccer ball (C₆₀), sp². - Coke \to impure, reducing agent, metallurgy.

69. The relation between n_m , $(n_m = \text{the number of permissible values of magnetic})$

quantum number (m)) for a given value of azimuthal quantum number (l), is

- (A) $n_m = 1 + 2$
- (B) $l = \frac{n_m 1}{2}$
- $(C) l = 2n_m + 1$
- (D) $n_m = 2l^2 + 1$

Correct Answer: (B) $l = \frac{n_m - 1}{2}$

Solution:

Step 1: Understanding the Question:

The question asks for the mathematical relationship between the azimuthal quantum number, l, and the total number of possible values for the magnetic quantum number, m_l (denoted here as n_m).

Step 2: Recalling Quantum Number Rules:

According to the rules of quantum mechanics for atomic orbitals:

- The azimuthal quantum number, l, can have integer values from 0 to n-1.
- For a given value of l, the magnetic quantum number, m_l , can have integer values from -l to +l, including 0.

Step 3: Deriving the Relationship:

Let's count the number of possible values for m_l for a given l. The values are:

$$-l, -(l-1), ..., -1, 0, +1, ..., +(l-1), +l$$

The total number of values (n_m) is the sum of:

- The number of negative values: l
- The number of positive values: l
- The value zero: 1

So,
$$n_m = l + l + 1 = 2l + 1$$
.

The direct relationship is $n_m = 2l + 1$.

Step 4: Checking the Options:

We need to find which of the given options is equivalent to $n_m = 2l + 1$.

- (A) $n_m = l + 2$. Incorrect.
- (B) $l = \frac{n_m 1}{2}$. Let's rearrange this equation to solve for n_m :

$$2l = n_m - 1$$

$$n_m = 2l + 1$$

This is the correct relationship.

- (C) $l = 2n_m + 1$. Incorrect.
- (D) $n_m = 2l^2 + 1$. Incorrect. (This is related to the number of orbitals in a subshell, but the formula is wrong).

Step 5: Final Answer:

The correct relationship, when expressed with l as the subject, is $l = \frac{n_m - 1}{2}$. This corresponds

87

to option (B).

Quick Tip

The number of orbitals in a subshell is given by 2l + 1. The number of electrons in a subshell is 2(2l + 1). Memorizing the fundamental relationship $n_m = 2l + 1$ allows you to quickly check any rearranged form of the equation.

70. For a certain reaction, the rate $= k[A]^2[B]$, when the initial concentration of A is tripled keeping concentration of B constant, the initial rate would

- (A) increase by a factor of three.
- (B) decrease by a factor of nine.
- (C) increase by a factor of six.
- (D) increase by a factor of nine.

Correct Answer: (D) increase by a factor of nine.

Solution:

Step 1: Understanding the Question:

The question asks how the initial rate of a reaction changes when the concentration of one reactant (A) is changed, given the rate law for the reaction.

Step 2: Key Formula or Approach:

The given rate law is:

$$Rate = k[A]^2[B]$$

We need to compare the initial rate with the new rate after changing the concentration of A.

Step 3: Detailed Explanation:

Let the initial rate be R_1 .

$$R_1 = k[A]^2[B]$$

Now, the concentration of A is tripled. Let the new concentration of A be [A'].

$$[A'] = 3[A]$$

The concentration of B remains constant. The new rate, R_2 , will be:

$$R_2 = k[A']^2[B]$$

Substitute the value of [A']:

$$R_2 = k(3[A])^2[B]$$

$$R_2 = k(9[A]^2)[B]$$

$$R_2 = 9 \times (k[A]^2[B])$$

Since $R_1 = k[A]^2[B]$, we can write:

$$R_2 = 9 \times R_1$$

This shows that the new rate is 9 times the initial rate.

Step 4: Final Answer:

The initial rate would increase by a factor of nine.

Quick Tip

The rate of reaction is directly proportional to the concentration of a reactant raised to the power of its order. Since the reaction is second order with respect to A, tripling its concentration will increase the rate by a factor of $3^2 = 9$.

71. Identify product (A) in the following reaction:

$$\xrightarrow{\text{Zn-Hg}}$$
 (A) + 2H₂O

- (A) Option (1)
- (B) Option (2)
- (C) Option (3)
- (D) Option (4)

Correct Answer: (B) Option (2)

Solution:

Step 1: Understanding the Question:

The question asks to identify the final product (A) of a given organic reaction. The reactant is a diketone and the reagent is Zn-Hg/conc. HCl.

Step 2: Key Formula or Approach:

The reagent system, Zinc amalgam (Zn-Hg) and concentrated hydrochloric acid (conc. HCl), is

used for the **Clemmensen reduction**. This reaction specifically reduces the carbonyl group (C=O) of aldehydes and ketones to a methylene group (CH₂).

$$\text{R-C(=O)-R'} \xrightarrow{\text{Zn-Hg, conc. HCl}} \text{R-CH}_2\text{-R'}$$

Step 3: Detailed Explanation:

The starting material is a substituted benzene ring containing two ketone functional groups (specifically, two acetyl groups, -COCH₃). The Clemmensen reduction will act on both of these carbonyl groups. Each acetyl group (-COCH₃) will be reduced to an ethyl group (-CH₂CH₃).

-C(=O)CH₃
$$\xrightarrow{\text{Zn-Hg, conc. HCl}}$$
 -CH₂CH₃

Since there are two such groups in the molecule, both will be converted. The cyclohexyl and benzene rings are unaffected by this reagent. Therefore, the product (A) will be the original molecule where both acetyl groups have been replaced by ethyl groups. Looking at the options: - Option (1) shows incorrect reduction. - Option (2) correctly shows both C=O groups reduced to CH_2 groups, forming two ethyl substituents on the benzene ring. - Option (3) shows reduction to alcohols (-CHOH), which would be the result of using a reducing agent like NaBH₄. - Option (4) shows an incorrect and incomplete reduction.

Step 4: Final Answer:

The correct product is shown in option (2), which results from the Clemmensen reduction of both ketone groups to methylene groups.

Quick Tip

Recognizing named reactions is key. Zn-Hg/conc. HCl is for Clemmensen reduction, which converts C=O to CH_2 in acidic conditions. For base-sensitive compounds, the Wolff-Kishner reduction (N_2H_4/KOH) achieves the same conversion.

72. A compound is formed by two elements A and B. The element B forms cubic close packed structure and atoms of A occupy 1/3 of tetrahedral voids. If the formula of the compound is A_xB_y , then the value of x + y is in option

- (A) 2
- (B) 5
- (C) 4
- (D) 3

Correct Answer: (B) 5

Solution:

Step 1: Understanding the Question:

This is a solid-state problem where we need to determine the empirical formula of a compound

based on the crystal lattice structure formed by its constituent elements.

Step 2: Key Formula or Approach:

In a cubic close-packed (CCP or FCC) structure: - If the number of atoms forming the lattice is N, - The number of octahedral voids is N. - The number of tetrahedral voids is 2N.

Step 3: Detailed Explanation:

Element B forms the CCP structure. Let's consider the number of atoms of B per unit cell as N. So, the effective number of B atoms is y = N.

Element A occupies 1/3 of the tetrahedral voids. The total number of tetrahedral voids is 2N. So, the effective number of A atoms is $x = \frac{1}{3} \times \text{(number of tetrahedral voids)} = \frac{1}{3} \times 2N = \frac{2N}{3}$. Now, let's find the ratio of atoms of A to B:

Ratio
$$x: y = \frac{2N}{3}: N$$

To simplify the ratio, we can divide by N:

Ratio
$$=\frac{2}{3}:1$$

To get the simplest whole number ratio, we multiply by 3:

Ratio
$$= 2:3$$

So, the formula of the compound is A_2B_3 . This means x = 2 and y = 3. The question asks for the value of x + y.

$$x + y = 2 + 3 = 5$$

Step 4: Final Answer:

The value of x + y is 5.

Quick Tip

For any close-packed arrangement (CCP/FCC or HCP), remember the ratio of lattice atoms: octahedral voids: tetrahedral voids is N:N:2N or 1:1:2. This is the foundation for solving most void-related solid-state problems.

73. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :

Assertion A: Helium is used to dilute oxygen in diving apparatus.

Reason R: Helium has high solubility in O_2 .

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.

- (C) Both A and R are true and R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (D) A is true but R is false.

Solution:

Step 1: Evaluating Assertion A:

Assertion A states that Helium is used to dilute oxygen in diving apparatus. This is a well-known fact. Deep-sea divers use a mixture of oxygen and helium (called heliox) for breathing. This is done to avoid nitrogen narcosis and decompression sickness ("the bends") which can occur when breathing compressed air (nitrogen-oxygen mixture) at great depths. Thus, **Assertion A is true**.

Step 2: Evaluating Reason R:

Reason R states that Helium has high solubility in O₂. This statement is ill-phrased as gases are miscible. The intended meaning is likely the solubility of Helium in blood. The primary reason for using helium is its **very low solubility** in blood compared to nitrogen, even under high pressure. This low solubility prevents the formation of gas bubbles in the blood and tissues during ascent, which is the cause of "the bends". Therefore, the reason provided (high solubility) is the exact opposite of the actual reason. Thus, **Reason R is false**.

Step 3: Concluding the relationship:

Since Assertion A is true and Reason R is false, we can choose the appropriate option.

Step 4: Final Answer:

Assertion A is true, but Reason R is false.

Quick Tip

Remember the cause of "the bends": nitrogen from compressed air dissolves in the blood at high pressure and forms bubbles upon decompression. Helium is used as a substitute for nitrogen because of its extremely low solubility in blood, which mitigates this dangerous condition.

- 74. Weight (g) of two moles of the organic compound, which is obtained by heating sodium ethanoate with sodium hydroxide in presence of calcium oxide is:
- (A) 18
- (B) 16
- (C) 32
- (D) 30

Correct Answer: (C) 32

Solution:

Step 1: Understanding the Question:

The question asks for the mass of two moles of a product formed from a specific chemical reaction. We first need to identify the reaction and its product.

Step 2: Key Formula or Approach:

The reaction involves heating sodium ethanoate (CH_3COONa) with sodium hydroxide (NaOH) in the presence of calcium oxide (CaO). This mixture (NaOH + CaO) is known as soda-lime, and the reaction is a **decarboxylation**. In this reaction, the carboxylate group (-COONa) is removed, and a hydrogen atom takes its place, forming an alkane with one less carbon atom.

Step 3: Detailed Explanation:

The reaction is:

$$CH_3COONa + NaOH \xrightarrow{CaO} CH_4 + Na_2CO_3$$

The organic compound obtained is methane (CH_4) .

Next, we need to calculate the weight of two moles of methane. First, find the molar mass of methane (CH₄): Molar Mass = (Atomic mass of C) + 4 × (Atomic mass of H) Molar Mass = $12.01 \text{ g/mol} + 4 \times 1.008 \text{ g/mol} \approx 16 \text{ g/mol}$.

The weight of one mole of methane is 16 g. The weight of two moles of methane is:

Weight =
$$2 \text{ moles} \times 16 \text{ g/mole} = 32 \text{ g}$$

Step 4: Final Answer:

The weight of two moles of the organic compound (methane) is 32 g.

Quick Tip

Soda-lime decarboxylation is a straightforward way to step down a carbon chain. It removes the entire -COOH group from a carboxylic acid (or -COONa from its salt) and replaces it with -H, forming an alkane. Remember that the product alkane has one carbon atom less than the parent salt.

94

75. The given compound is an example of

- (A) vinylic halide
- (B) benzylic halide
- (C) aryl halide
- (D) allylic halide

Correct Answer: (D) allylic halide

Solution:

Step 1: Understanding the Question:

The question asks to classify the given organic halide based on the position of the halogen atom (X).

Step 2: Key Definitions:

- **Aryl halide:** Halogen is directly bonded to an sp²-hybridized carbon atom of an aromatic ring. - **Vinylic halide:** Halogen is directly bonded to an sp²-hybridized carbon atom of a carbon-carbon double bond. - **Benzylic halide:** Halogen is bonded to an sp³-hybridized carbon atom which is directly attached to an aromatic ring. - **Allylic halide:** Halogen is bonded to an sp³-hybridized carbon atom which is adjacent to a carbon-carbon double bond.

Step 3: Detailed Explanation:

Let's analyze the structure of the given compound: C_6H_5 -CH=CH-CH(X)-CH₂CH₃. 1. Identify the carbon atom to which the halogen (X) is attached. It is the -CH(X)- carbon. 2. Determine the hybridization of this carbon atom. Since it forms four single bonds (to a C, another C, a H, and X), it is sp³-hybridized. 3. Look at the atom adjacent to this sp³ carbon. It is part of a carbon-carbon double bond (-CH=CH-). 4. A compound where a halogen is attached to an sp³-hybridized carbon atom that is next to a C=C double bond fits the definition of an **allylic halide**. The position itself is called the allylic position.

The compound is not vinylic (X is not on the C=C), not aryl (X is not on the benzene ring), and not benzylic (the C-X carbon is not directly attached to the benzene ring).

Step 4: Final Answer:

The given compound is an example of an allylic halide.

Quick Tip

To classify halides, always focus on the carbon atom directly bonded to the halogen. The key is its hybridization (sp^2 or sp^3) and what that carbon atom is bonded to (a double bond, an aromatic ring, etc.).

- 76. Select the correct statements from the following:
- A. Atoms of all elements are composed of two fundamental particles.
- B. The mass of the electron is 9.10939×10^{-31} kg.
- C. All the isotopes of a given element show same chemical properties.

- D. Protons and electrons are collectively known as nucleons.
- E. Dalton's atomic theory, regarded the atom as an ultimate particle of matter. Choose the correct answer from the options given below :
- (A) B, C and E only
- (B) A, B and C only
- (C) C, D and E only
- (D) A and E only

Correct Answer: (A) B, C and E only

Solution:

Step 1: Understanding the Question:

We need to evaluate the correctness of five statements related to atomic structure and theory.

Step 2: Analyzing Each Statement:

- A. Atoms of all elements are composed of two fundamental particles. This is false. Atoms are composed of three fundamental particles: protons, neutrons, and electrons. (The hydrogen-1 isotope is an exception, having no neutrons, but the statement says "all elements").
- B. The mass of the electron is 9.10939×10^{-31} kg. This is a factual statement and is true. This is the accepted value for the rest mass of an electron.
- C. All the isotopes of a given element show same chemical properties. Isotopes of an element have the same number of protons and electrons but different numbers of neutrons. Since chemical properties are primarily determined by the electron configuration (which is the same for all isotopes of an element), this statement is **true**.
- D. Protons and electrons are collectively known as nucleons. This is false. Nucleons are the particles found in the nucleus, which are protons and neutrons. Electrons orbit the nucleus.
- E. Dalton's atomic theory, regarded the atom as an ultimate particle of matter. This is a correct description of one of the main postulates of Dalton's original atomic theory. He considered atoms to be indivisible ("ultimate particles"). So, the statement is **true**.

Step 3: Identifying the Correct Set:

The correct statements are B, C, and E.

Step 4: Final Answer:

The combination of correct statements is B, C, and E. This corresponds to option (A).

Quick Tip

When evaluating historical scientific theories like Dalton's, answer based on the theory as it was originally proposed. Dalton's idea of an indivisible atom is considered correct in the context of his theory, even though we now know atoms can be divided.

77. Amongst the given options which of the following molecules / ion acts as a Lewis acid?

- $(A) OH^-$
- (B) NH₃
- $(C) H_2O$
- (D) BF₃

Correct Answer: (D) BF₃

Solution:

Step 1: Understanding the Question:

The question asks to identify the Lewis acid from the given list of species.

Step 2: Defining Lewis Acids and Bases:

- Lewis Acid: A chemical species that can accept a pair of electrons. Lewis acids are typically electron-deficient. Examples include cations, molecules with an incomplete octet on the central atom, or molecules with a central atom that can expand its octet.
- Lewis Base: A chemical species that can donate a pair of electrons. Lewis bases must have at least one lone pair of electrons. Examples include anions and neutral molecules with lone pairs.

Step 3: Analyzing the Options:

- (A) OH⁻: The hydroxide ion has lone pairs on the oxygen atom and a negative charge. It is a classic electron-pair donor, so it acts as a **Lewis base**.
- (B) NH₃: The ammonia molecule has a lone pair on the central nitrogen atom, which it can donate. It acts as a **Lewis base**.
- (C) H_2O : The water molecule has two lone pairs on the central oxygen atom, which it can donate. It acts as a **Lewis base**.
- (D) BF₃: In boron trifluoride, the central boron atom has only six electrons in its valence shell (three single bonds with fluorine). It has an incomplete octet and an empty p-orbital, making it capable of accepting a pair of electrons. Therefore, it acts as a **Lewis acid**.

Step 4: Final Answer:

 BF_3 is the only species in the list that is an electron-pair acceptor (Lewis acid). This corresponds to option (D).

Quick Tip

To quickly identify Lewis acids, look for molecules with central atoms from Group 2 (like Be) or Group 13 (like B, Al) which often have incomplete octets. To identify Lewis bases, look for anions or molecules with central atoms having lone pairs (like N, O, P, S).

78. The right option for the mass of CO_2 produced by heating 20 g of 20% pure limestone is (Atomic mass of Ca = 40)

 $\mathbf{CaCO}_3 \xrightarrow{1200 \ \mathbf{K}} \mathbf{CaO} + \mathbf{CO}_2$

- (A) 1.32 g
- (B) 1.12 g
- (C) 1.76 g
- (D) 2.64 g

Correct Answer: (C) 1.76 g

Solution:

Step 1: Understanding the Question:

This is a stoichiometry problem involving a chemical reaction with an impure reactant. We need to calculate the mass of carbon dioxide produced from a given mass of impure limestone.

Step 2: Calculating the Mass of Pure Reactant:

Limestone is primarily calcium carbonate (CaCO₃).

- Total mass of impure limestone = 20 g.
- Purity = 20%.
- Mass of pure $CaCO_3 = Total mass \times Purity percentage$

Mass of
$$CaCO_3 = 20 \text{ g} \times \frac{20}{100} = 4 \text{ g}$$

Step 3: Stoichiometric Calculation:

The balanced chemical equation is:

$$CaCO_3(s) \to CaO(s) + CO_2(g)$$

First, calculate the molar masses of $CaCO_3$ and CO_2 .

- Molar mass of $CaCO_3 = 40$ (Ca) + 12 (C) + 3 × 16 (O) = 40 + 12 + 48 = 100 g/mol.
- Molar mass of $CO_2 = 12 (C) + 2 \times 16 (O) = 12 + 32 = 44 \text{ g/mol}.$

From the stoichiometry of the reaction, 1 mole of $CaCO_3$ produces 1 mole of CO_2 .

In terms of mass, 100 g of $CaCO_3$ produces 44 g of CO_2 .

We have 4 g of pure CaCO₃. Let the mass of CO₂ produced be 'x'. We can set up a proportion:

$$\frac{\text{mass of CO}_2}{\text{mass of CaCO}_3} = \frac{44 \text{ g}}{100 \text{ g}}$$
$$\frac{x}{4 \text{ g}} = \frac{44}{100}$$
$$x = 4 \times \frac{44}{100} = \frac{176}{100} = 1.76 \text{ g}$$

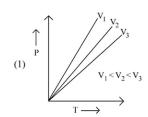
Step 4: Final Answer:

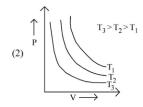
The mass of CO_2 produced is 1.76 g. This corresponds to option (C).

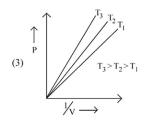
Quick Tip

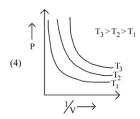
In stoichiometry problems with impure reactants, the first step is always to calculate the mass of the pure substance that will actually react. The impurities are assumed to be inert and do not participate in the reaction.

79. Which amongst the following options is correct graphical representation of Boyle's Law?









Correct Answer: (C)

Solution:

Step 1: Understanding Boyle's Law:

Boyle's Law states that for a fixed amount of gas at a constant temperature, the pressure (P) of the gas is inversely proportional to its volume (V).

99

Mathematically, $P \propto \frac{1}{V}$ or PV = k (where k is a constant).

From the ideal gas law, PV = nRT. For a fixed amount of gas (n is constant), $PV = (\text{constant}) \times T$. So, the constant k in Boyle's law is directly proportional to the absolute temperature T.

Step 2: Analyzing the Graphs:

The law states PV = k. This means a plot of P vs. V will be a hyperbola (as seen in graph 2, but with the wrong axes labels for Boyle's law).

Let's analyze the variables plotted in each graph:

- Graph (1): P vs. T at constant V. This represents Gay-Lussac's Law $(P \propto T)$, not Boyle's Law.
- Graph (2): P vs. V at constant T. The shape is hyperbolic $(P \propto 1/V)$, which is correct for Boyle's Law. The curves are called isotherms. Since PV = nRT, for a given volume, P is higher at higher T. So $T_3 > T_2 > T_1$. This graph correctly represents Boyle's Law isotherms, but the options are plots of P vs 1/V.
- Graph (3): P vs. 1/V at constant T. Boyle's Law can be written as $P = k \times \frac{1}{V}$. This is in the form of a straight line equation y = mx, where y = P, x = 1/V, and the slope m = k. Thus, a plot of P vs. 1/V should be a straight line passing through the origin. Graph (3) shows this correctly.

Furthermore, the slope m = k = nRT. This means the slope is proportional to the temperature T. A higher temperature will result in a steeper slope. In graph (3), the slope of the line for T_3 is the highest, and the slope for T_1 is the lowest. This corresponds to the condition $T_3 > T_2 > T_1$, which is consistent. Thus, graph (3) is a correct representation.

- Graph (4): P vs. 1/V at constant T. This graph shows hyperbolic curves, which is incorrect for a P vs. 1/V plot.

Step 3: Final Answer:

A plot of Pressure (P) versus the reciprocal of Volume (1/V) for an ideal gas at constant temperature (Boyle's Law) is a straight line passing through the origin, with a slope proportional to the temperature. Graph (3) correctly depicts this relationship.

Quick Tip

To analyze gas law graphs, always rearrange the relevant equation into the form y = mx + c. For Boyle's law (PV = k), plotting P vs. 1/V gives P = k(1/V), which is a straight line y = mx with slope k = nRT. This makes it easy to see the expected shape and how the slope changes with temperature.

80. Which one of the following statements is correct?

- (A) Mg plays roles in neuromuscular function and interneuronal transmission.
- (B) The daily requirement of Mg and Ca in the human body is estimated to be 0.2-0.3 g.
- (C) All enzymes that utilise ATP in phosphate transfer require Ca as the cofactor.
- (D) The bone in human body is an inert and unchanging substance.

Correct Answer: (B) The daily requirement of Mg and Ca in the human body is estimated to be 0.2-0.3 g.

Solution:

Step 1: Understanding the Question:

The question requires us to identify the correct statement among the given options related to the roles and requirements of elements in the human body. Based on the provided answer key, statement (B) is the correct one. Let's analyze all options to understand why.

Step 2: Detailed Explanation of Each Statement:

- (A) Mg plays roles in neuromuscular function and interneuronal transmission. This statement is factually correct. Magnesium is essential for nerve transmission and muscle function, acting as a physiological calcium antagonist. However, in the context of this specific question and its answer key, it is considered incorrect, possibly due to a subtle inaccuracy or being less correct than the intended answer.
- (B) The daily requirement of Mg and Ca in the human body is estimated to be 0.2-0.3 g. This statement is poorly worded but likely considered correct in the context of the exam. The daily requirement for Magnesium (Mg) is indeed around 200-300 mg (0.2-0.3 g). The requirement for Calcium (Ca) is much higher (around 1000 mg or 1 g). The statement is only accurate for Mg. Given that other options are clearly incorrect, this is likely the intended answer despite the inclusion of "and Ca".
- (C) All enzymes that utilise ATP in phosphate transfer require Ca as the cofactor. This statement is **incorrect**. The vast majority of enzymes that utilize ATP, such as kinases, require Magnesium (Mg^{2+}) as a cofactor. ATP often functions as an Mg-ATP complex.
- (D) The bone in human body is an inert and unchanging substance. This statement is **incorrect**. Bone is a dynamic, living tissue that is constantly undergoing a process of remodeling, where old bone is broken down and new bone is formed.

Step 3: Final Answer:

Statements (C) and (D) are definitively false. While statement (A) is correct, the provided answer key points to (B). This suggests we should interpret statement (B) as correct based on the value for Mg, despite the inaccuracy regarding Ca.

Quick Tip

In competitive exams, be aware that questions can sometimes be flawed or ambiguous. If three options are clearly and factually incorrect, the remaining option is likely the intended answer, even if it has some inaccuracies. In this case, the cofactor for ATP is Mg^{2+} , a very important fact to remember.

81. The conductivity of centimolar solution of KCl at 25° C is 0.0210 ohm⁻¹ cm⁻¹ and the resistance of the cell containing the solution at 25° C is 60 ohm. The value

of cell constant is -

- (A) 3.34 cm^{-1}
- (B) 1.34 cm^{-1}
- (C) 3.28 cm^{-1}
- (D) 1.26 cm^{-1}

Correct Answer: (D) 1.26 cm^{-1}

Solution:

Step 1: Understanding the Question:

We are given the conductivity (κ) and resistance (R) of an electrolyte solution in a conductivity cell. We need to calculate the cell constant (G^*) .

Step 2: Key Formula or Approach:

The relationship between conductivity (κ) , resistance (R), and cell constant (G^*) is given by:

$$\kappa = \frac{1}{R} \times G^*$$

The cell constant is defined as the ratio of the distance between the electrodes (l) to their area of cross-section (A), i.e., $G^* = \frac{l}{A}$. We can rearrange the formula to solve for the cell constant:

$$G^* = \kappa \times R$$

Step 3: Detailed Explanation:

Given values are: - Conductivity, $\kappa=0.0210\,\Omega^{-1}\,\mathrm{cm}^{-1}$ - Resistance, $R=60\,\Omega$ Now, substitute these values into the formula:

$$G^* = (0.0210 \,\Omega^{-1} \,\mathrm{cm}^{-1}) \times (60 \,\Omega)$$
$$G^* = 0.021 \times 60 \,\mathrm{cm}^{-1}$$
$$G^* = 1.26 \,\mathrm{cm}^{-1}$$

The information about "centimolar solution of KCl" is not required for this calculation as the conductivity value is already provided.

Step 4: Final Answer:

The value of the cell constant is 1.26 cm^{-1} .

Quick Tip

Remember the fundamental formula of conductivity: Conductivity = Conductance × Cell Constant. Since conductance is the reciprocal of resistance (1/R), the formula becomes $\kappa = (1/R) \times G^*$. This is a direct application problem, so knowing the formula is key.

82. The number of σ bonds, π bonds and lone pair of electrons in pyridine, respectively are:

- (A) 12, 2, 1
- (B) 11, 2, 0
- (C) 12, 3, 0
- (D) 11, 3, 1

Correct Answer: (D) 11, 3, 1

Solution:

Step 1: Understanding the Question:

We need to determine the total number of sigma (σ) bonds, pi (π) bonds, and lone pairs of electrons in the pyridine molecule.

Step 2: Structure of Pyridine:

Pyridine (C_5H_5N) is a six-membered aromatic ring similar to benzene, but with one CH group replaced by a nitrogen atom. The structure consists of 5 carbon atoms, 5 hydrogen atoms, and 1 nitrogen atom, arranged in a hexagonal ring with alternating double bonds.

Step 3: Detailed Explanation and Counting:

- 1. Sigma (σ) Bonds: These are all the single bonds within the structure.
- Bonds within the ring: There are 6 atoms in the ring, forming 6 σ bonds (5 C-C/C-N single bonds, 1 C-N single bond). In reality, it's 4 C-C bonds and 2 C-N bonds. Total = 6 σ bonds in the ring framework.
- Bonds to hydrogen: Each of the 5 carbon atoms is bonded to one hydrogen atom. This gives 5 C-H σ bonds.
- Total σ bonds = (bonds in ring) + (bonds to H) = 6 + 5 = 11 σ bonds.
- 2. Pi (π) Bonds: Pyridine is an aromatic system with 6π electrons, which corresponds to 3 double bonds (or 3 π bonds) delocalized around the ring.

3. Lone Pair of Electrons:

- Carbon atoms use all 4 valence electrons for bonding.
- Hydrogen atoms use their single electron for bonding.
- The nitrogen atom (Group 15) has 5 valence electrons. In pyridine, it forms three bonds (two σ and one π). This uses 3 valence electrons. The remaining 2 electrons form one lone pair. This lone pair resides in an sp² orbital in the plane of the ring and does not participate in the aromatic system.

So, there is 1 lone pair of electrons on the nitrogen atom.

Step 4: Final Answer:

The counts are: 11σ bonds, 3π bonds, and 1 lone pair. This corresponds to option (D).

Quick Tip

For cyclic compounds, a quick way to count σ bonds is to count the number of atoms and add the number of rings, then subtract 1 if it's a single ring structure. Or simply, count all atoms and add the number of rings - 1. For pyridine: 12 atoms, 1 ring. So 12+1-1=12 σ -bonds? No, the simple method is: Count all single bonds + number of double/triple bonds. For pyridine, count every line representing a bond once for sigma: 6 in the ring, 5 C-H bonds = 11 σ bonds. Then count the extra lines for double bonds: 3 π bonds.

83. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: A reaction can have zero activation energy.

Reason R: The minimum extra amount of energy absorbed by reactant molecules so that their energy becomes equal to threshold value, is called activation energy. In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true and R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (A) A is false but R is true.

Solution:

Step 1: Evaluating Assertion A:

Assertion A states that a reaction can have zero activation energy. In the context of many introductory chemistry curricula and competitive exams, it is considered that for any chemical transformation involving the breaking and forming of bonds, some energy is required to overcome an energy barrier. Therefore, the activation energy (Ea) must be greater than zero. From this perspective, the assertion that Ea can be zero is considered false. (Note: In advanced kinetics, some barrierless reactions, like certain radical combinations, are described as having zero or even negative activation energy, but this is typically outside the scope of these exams). Thus, for the purpose of this exam, Assertion A is false.

Step 2: Evaluating Reason R:

Reason R gives the definition of activation energy: "The minimum extra amount of energy absorbed by reactant molecules so that their energy becomes equal to threshold value, is called activation energy." This is the standard and correct definition of activation energy according to the collision theory. Thus, **Reason R is true**.

Step 3: Concluding the relationship:

Since Assertion A is considered false and Reason R is true, the correct option is (A).

Step 4: Final Answer:

A is false but R is true.

Quick Tip

In assertion-reason questions, first evaluate each statement independently for its truthfulness. Reason R provides a perfect definition, making it true. Assertion A makes an absolute claim that can be considered false in a general chemistry context, as most reactions are assumed to have an energy barrier.

- 84. Which of the following reactions will NOT give primary amine as the product?
- (A) CH₃CONH₂ $\xrightarrow{(i)LiAlH_4(ii)H_3O^+}$ Product
- (B) $CH_3CONH_2 \xrightarrow{Br_2/KOH} Product$
- (C) CH₃CN $\xrightarrow{(i)LiAlH_4(ii)H_3O^+}$ Product
- (D) CH₃NC $\xrightarrow{(i)LiAlH_4(ii)H_3O^+}$ Product

Correct Answer: (D) CH₃NC $\xrightarrow{(i)LiAlH_4(ii)H_3O^+}$ Product

Solution:

Step 1: Understanding the Question:

We need to analyze four different reactions and identify which one does not produce a primary amine (R-NH₂).

Step 2: Detailed Explanation of Each Reaction:

- (A) CH₃CONH₂ (Acetamide) with LiAlH₄: This is the reduction of a primary amide. Lithium aluminium hydride (LiAlH₄) is a strong reducing agent that reduces the carbonyl group (C=O) of an amide to a methylene group (CH₂).

$$CH_3CONH_2 \xrightarrow{LiAlH_4} CH_3CH_2NH_2$$
 (Ethylamine, a primary amine)

- (B) CH_3CONH_2 (Acetamide) with Br_2/KOH : This is the Hofmann bromamide degradation reaction. It converts a primary amide into a primary amine with one less carbon atom.

$${\rm CH_3CONH_2} \xrightarrow{{\rm Br_2/KOH}} {\rm CH_3NH_2} \quad ({\rm Methylamine, \ a \ primary \ amine})$$

- (C) CH_3CN (Acetonitrile) with $LiAlH_4$: This is the reduction of a nitrile. $LiAlH_4$ reduces the cyano group (-C \equiv N) to a primary amine group (-C H_2NH_2).

$$CH_3C \equiv N \xrightarrow{LiAlH_4} CH_3CH_2NH_2$$
 (Ethylamine, a primary amine)

- (D) CH₃NC (Methyl isocyanide) with LiAlH₄: This is the reduction of an isocyanide (or isonitrile). The reduction adds four hydrogen atoms across the $N \equiv C$ triple bond, resulting

in a secondary amine.

$$CH_3-N \equiv C \xrightarrow{LiAlH_4} CH_3-NH-CH_3$$
 (Dimethylamine, a secondary amine)

Step 3: Final Answer:

The reduction of methyl isocyanide (CH₃NC) with LiAlH₄ yields dimethylamine, which is a secondary amine, not a primary amine. Therefore, this reaction is the correct answer.

Quick Tip

Remember the key difference in the reduction of nitriles and isonitriles. Nitriles (R-CN) give primary amines (R-CH₂NH₂). Isonitriles (R-NC) give secondary amines (R-NH-CH₃). This is a common point of confusion tested in exams.

85. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :

Assertion A: Metallic sodium dissolves in liquid ammonia giving a deep blue solution, which is paramagnetic.

Reason R: The deep blue solution is due to the formation of amide.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (D) A is true but R is false.

Solution:

Step 1: Evaluating Assertion A:

Assertion A describes the result of dissolving sodium in liquid ammonia. When an alkali metal like sodium is dissolved in liquid ammonia, it ionizes to form a cation and a free electron.

$$Na(s) \xrightarrow{NH_3(l)} Na^+(am) + e^-(am)$$

The notation "(am)" signifies that the ions are ammoniated (solvated by ammonia). The presence of these ammoniated electrons is responsible for the characteristic deep blue color of the solution. Since these electrons are unpaired, they make the solution paramagnetic. Thus, **Assertion A is true**.

Step 2: Evaluating Reason R:

Reason R claims that the deep blue color is due to the formation of amide. Sodium amide

(NaNH₂) can be formed in these solutions, but it is a slow decomposition reaction that occurs over time, often requiring a catalyst.

$$2Na + 2NH_3 \rightarrow 2NaNH_2 + H_2$$

The formation of sodium amide actually leads to the fading of the blue color. The blue color itself is unequivocally due to the ammoniated electrons, which absorb energy in the visible region of the electromagnetic spectrum. Therefore, **Reason R is false**.

Step 3: Concluding the relationship:

Since Assertion A is true and Reason R is false, the correct option is (D).

Step 4: Final Answer:

A is true but R is false.

Quick Tip

The dissolution of alkali metals in liquid ammonia is a very important reaction. Remember the key species responsible for the properties:

- Blue Color: Ammoniated electrons.
- Paramagnetism: Unpaired ammoniated electrons.
- Electrical Conductivity: Both ammoniated cations and electrons.
- Reducing Nature: Ammoniated electrons.

86. Match List - I with List - II:

List - I (Oxoacids of Sulphur)

List - II (Bonds)

A. Peroxodisulphuric acid

I. Two S-OH, Four S=O, One S-O-S

B. Sulphuric acid

II. Two S-OH, One S=O

C. Pyrosulphuric acid

III. Two S-OH, Four S=O, One S-O-O-S

D. Sulphurous acid

IV. Two S-OH, Two S=O

Choose the correct answer from the options given below:

- (A) A-III, B-IV, C-II, D-I
- (B) A-I, B-III, C-II, D-IV
- (C) A-III, B-IV, C-I, D-II
- (D) A-I, B-III, C-IV, D-II

Correct Answer: (C) A-III, B-IV, C-I, D-II

Solution:

Step 1: Understanding the Question:

This question requires matching the given oxoacids of sulfur with the correct description of the bonds present in their structures.

Step 2: Analyzing the Structure of Each Acid:

- A. Peroxodisulphuric acid (H₂S₂O₈, Marshall's Acid): The structure is HO-SO₂-O-O-SO₂-OH. It contains a peroxide linkage (-O-O-). Counting the bonds: two S-OH bonds, four S=O bonds, and one S-O-O-S group. This matches with **List-II**, **III**.
- B. Sulphuric acid (H_2SO_4): The structure is HO- SO_2 -OH. Counting the bonds: two S-OH bonds and two S=O bonds. This matches with **List-II**, **IV**.
- C. Pyrosulphuric acid ($H_2S_2O_7$, Oleum): The structure is HO- SO_2 -O- SO_2 -OH. It contains an S-O-S linkage. Counting the bonds: two S-OH bonds, four S=O bonds, and one S-O-S bond. This matches with **List-II**, **I**.
- D. Sulphurous acid (H_2SO_3): The structure is HO-SO-OH. Counting the bonds: two S-OH bonds and one S=O bond. This matches with **List-II**, **II**.

Step 3: Compiling the Correct Match:

The correct matches are:

 $A \rightarrow III$

 $B \to IV$

 $\mathrm{C} \to \mathrm{I}$

 $D \rightarrow II$

This combination corresponds to option (C).

Step 4: Final Answer:

The correct match is A-III, B-IV, C-I, D-II. This corresponds to option (C).

Quick Tip

Drawing the structures of the oxoacids of sulfur is essential. Key features to remember are: pyrosulphuric acid has an S-O-S link, and peroxodisulphuric acid has an S-O-O-S link. The number of OH groups generally corresponds to the basicity of the acid.

- 87. The equilibrium concentrations of the species in the reaction $A + B \rightleftharpoons C + D$ are 2, 3, 10 and 6 mol L^{-1} , respectively at 300 K. ΔG° for the reaction is (R = 2 cal / mol K)
- (A) -13.73 cal
- (B) 1372.60 cal
- (C) -137.26 cal
- (D) -1381.80 cal

Correct Answer: (D) -1381.80 cal

Solution:

Step 1: Understanding the Question:

We are given the equilibrium concentrations for a reversible reaction and asked to calculate the

standard Gibbs free energy change, ΔG° .

Step 2: Key Formula or Approach:

The relationship between the standard Gibbs free energy change and the equilibrium constant (K) is given by:

$$\Delta G^{\circ} = -RT \ln K$$

First, we need to calculate the equilibrium constant, K_c , from the given concentrations.

Step 3: Detailed Calculation:

The reaction is $A + B \rightleftharpoons C + D$.

Given equilibrium concentrations:

Α

 $\equiv 2 \text{ mol } L^{-1}$

 $\equiv 3 \text{ mol } L^{-1}$

 $\equiv 10 \text{ mol } L^{-1}$

 $= 6 \text{ mol } L^{-1}$

Calculate K_c :

$$K_c = \frac{[C][D]}{[A][B]} = \frac{(10)(6)}{(2)(3)} = \frac{60}{6} = 10$$

Calculate ΔG° :

Given: R = 2 cal/mol K, T = 300 K.

$$\Delta G^{\circ} = -(2 \text{ cal/mol K}) \times (300 \text{ K}) \times \ln(10)$$
$$\Delta G^{\circ} = -600 \ln(10) \text{ cal/mol}$$

Using the value $ln(10) \approx 2.303$:

$$\Delta G^{\circ} = -600 \times 2.303 = -1381.8 \text{ cal/mol}$$

Step 4: Final Answer:

The value of ΔG° for the reaction is -1381.80 cal. This corresponds to option (D).

Quick Tip

Make sure to use the correct value of R based on the desired units for energy. If the answer is in Joules, use R = 8.314 J/mol K. If in calories, use $R \approx 2$ cal/mol K. Also, remember that a K value greater than 1 corresponds to a negative ΔG° (spontaneous reaction), and a K value less than 1 corresponds to a positive ΔG° .

88. Which complex compound is most stable?

- (A) $[Co(NH_3)_6]_2(SO_4)_3$
- (B) $[Co(NH_3)_4(H_2O)Br](NO_3)_2$
- (C) $[Co(NH_3)_3(NO_3)_3]$
- (D) $[CoCl_2(en)_2]NO_3$

Correct Answer: (D) $[CoCl_2(en)_2]NO_3$

Solution:

Step 1: Understanding the Question:

The question asks to identify the most stable complex compound from the given options. The stability of a coordination complex is influenced by several factors, most notably the chelate effect.

Step 2: The Chelate Effect:

The chelate effect refers to the enhanced stability of coordination complexes containing chelate rings. A chelate ring is formed when a polydentate ligand (a ligand that can bind to the central metal ion through more than one donor atom) binds to the metal ion. Complexes with chelate rings are thermodynamically more stable than analogous complexes with monodentate ligands.

Step 3: Analyzing the Ligands in Each Complex:

- (A) $[Co(NH_3)_6]_2(SO_4)_3$: The ligand is ammonia (NH₃), which is a monodentate ligand. No chelate rings are formed.
- (B) $[Co(NH_3)_4(H_2O)Br](NO_3)_2$: The ligands are ammonia (NH₃), water (H₂O), and bromide (Br⁻). All are monodentate ligands. No chelate rings are formed.
- (C) $[Co(NH_3)_3(NO_3)_3$: The ligands are ammonia (NH_3) and nitrate (NO_3^-) , both acting as monodentate ligands here. No chelate rings are formed.
- (D) [CoCl₂(en)₂]NO₃: The ligands are chloride (Cl⁻) and ethylenediamine (en). Ethylenediamine (H₂N-CH₂-CH₂-NH₂) is a bidentate ligand, meaning it binds to the cobalt ion at two points. Each 'en' ligand forms a stable five-membered chelate ring with the cobalt ion.

Step 4: Final Answer:

Due to the presence of two chelate rings formed by the bidentate ethylenediamine ligands, the complex $[\text{CoCl}_2(\text{en})_2]^+$ exhibits significant stability enhancement due to the chelate effect. Therefore, it is the most stable complex among the given options. This corresponds to option (D).

Quick Tip

When comparing the stability of coordination complexes, always look for the presence of polydentate ligands first. The chelate effect is a dominant factor in determining complex stability. Common polydentate ligands to know are ethylenediamine (en), oxalate (ox), and EDTA.

89. On balancing the given redox reaction,

a $Cr_2O_7^{2-}$ + b $SO_3^{2-}(aq)$ + c $H^+(aq)$ \rightarrow 2a $Cr^{3+}(aq)$ + b $SO_4^{2-}(aq)$ + $\frac{c}{2}H_2O(l)$ the coefficients a, b and c are found to be, respectively -

- (A) 8, 1, 3
- (B) 1, 3, 8
- (C) 3, 8, 1
- (D) 1, 8, 3

Correct Answer: (B) 1, 3, 8

Solution:

Step 1: Understanding the Question:

We need to balance the given redox reaction in an acidic medium and find the stoichiometric coefficients a, b, and c.

Step 2: Balancing using the Half-Reaction Method:

Oxidation Half-Reaction: The oxidation state of S in SO_3^{2-} is +4, and in SO_4^{2-} is +6. So, sulfite is oxidized.

$$SO_3^{2-} \rightarrow SO_4^{2-}$$

1. Balance O atoms by adding H_2O :

$$SO_3^{2-} + H_2O \rightarrow SO_4^{2-}$$

2. Balance H atoms by adding H⁺:

$$SO_3^{2-} + H_2O \rightarrow SO_4^{2-} + 2H^+$$

3. Balance charge by adding electrons (e⁻):

$$SO_3^{2-} + H_2O \rightarrow SO_4^{2-} + 2H^+ + 2e^-$$

Reduction Half-Reaction: The oxidation state of Cr in $Cr_2O_7^{2-}$ is +6, and in Cr^{3+} is +3. So, dichromate is reduced.

$$\mathrm{Cr}_2\mathrm{O}_7^{2-} \to \mathrm{Cr}^{3+}$$

1. Balance atoms other than O and H (i.e., Cr):

$$\mathrm{Cr}_2\mathrm{O}_7^{2-} \to 2\mathrm{Cr}^{3+}$$

2. Balance O atoms by adding H_2O :

$$Cr_2O_7^{2-} \to 2Cr^{3+} + 7H_2O$$

3. Balance H atoms by adding H⁺:

$$Cr_2O_7^{2-} + 14H^+ \rightarrow 2Cr^{3+} + 7H_2O$$

4. Balance charge by adding electrons:

$$Cr_2O_7^{2-} + 14H^+ + 6e^- \rightarrow 2Cr^{3+} + 7H_2O$$

Step 3: Combining the Half-Reactions:

To make the number of electrons equal in both half-reactions, multiply the oxidation half-reaction by 3.

$$3SO_3^{2-} + 3H_2O \rightarrow 3SO_4^{2-} + 6H^+ + 6e^-$$

Now, add this to the reduction half-reaction:

$$Cr_2O_7^{2-} + 14H^+ + 6e^- + 3SO_3^{2-} + 3H_2O \rightarrow 2Cr^{3+} + 7H_2O + 3SO_4^{2-} + 6H^+ + 6e^-$$

Cancel out the electrons and simplify H^+ and H_2O :

$$Cr_2O_7^{2-} + 3SO_3^{2-} + 8H^+ \rightarrow 2Cr^{3+} + 3SO_4^{2-} + 4H_2O$$

Step 4: Determining the Coefficients:

Comparing this balanced equation to the given format: a $Cr_2O_7^{2-}$ + b SO_3^{2-} + c H⁺ ... We find that a=1, b=3, and c=8.

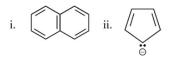
Step 5: Final Answer:

The coefficients a, b, and c are 1, 3, and 8, respectively. This corresponds to option (B).

Quick Tip

After balancing a redox reaction, always do a final check of both atom balance and charge balance on both sides of the equation to ensure it is correct. For the final equation here: Cr (2 vs 2), S (3 vs 3), O (7+9=16 vs 12+4=16), H (8 vs 8). Charge: (-2) + 3(-2) + 8(+1) = 0 vs 2(+3) + 3(-2) = 0. The equation is correctly balanced.

90. Consider the following compounds/species:



The number of compounds/species which obey Huckel's rule is

- (A) 5
- (B) 4
- (C) 6

Correct Answer: (A) 5

Solution:

Step 1: Understanding the Question and Huckel's Rule:

We need to identify which of the given species are aromatic. A species is aromatic if it satisfies Huckel's rule, which requires the species to be:

- 1. Cyclic
- 2. Planar
- 3. Fully conjugated (a p-orbital on every atom in the ring)
- 4. Contains $(4n + 2) \pi$ electrons, where n = 0, 1, 2, ...

Step 2: Analyzing Each Species:

- i. Naphthalene: It is cyclic, planar, and fully conjugated. It has 10π electrons (5 double bonds). $4n + 2 = 10 \implies n = 2$. It is aromatic.
- ii. Cyclopentadienyl anion: It is cyclic, planar, and fully conjugated. It has 6 π electrons (2 from double bonds, 2 from the lone pair). $4n + 2 = 6 \implies n = 1$. It is aromatic.
- iii. Cyclobutadiene: It is cyclic and conjugated but has 4π electrons. This is a 4n system (n=1), which makes it anti-aromatic.
- iv. Tropylium cation: It is cyclic, planar, and fully conjugated. It has 6 π electrons (3 double bonds). $4n + 2 = 6 \implies n = 1$. It is aromatic.
- v. Cyclopropenyl cation: It is cyclic, planar, and fully conjugated. It has 2π electrons (1 double bond). $4n + 2 = 2 \implies n = 0$. It is **aromatic**.
- vi. Cyclooctatetraene: It has 8 π electrons, which is a 4n system (n=2). To avoid the instability of anti-aromaticity, it adopts a non-planar, tub-like shape. Therefore, it is non-aromatic.
- vii. Anthracene: It is a polycyclic, planar, fully conjugated system. It has 14 π electrons (7 double bonds). $4n + 2 = 14 \implies n = 3$. It is aromatic.

Step 3: Counting the Aromatic Species:

The species that obey Huckel's rule and are aromatic are: i, ii, iv, v, and vii. Total count = 5.

Step 4: Final Answer:

There are 5 aromatic compounds/species in the given list. This corresponds to option (A).

Quick Tip

Remember the magic numbers for aromaticity (Huckel numbers): 2, 6, 10, 14, ... π electrons. Also, remember to distinguish between aromatic (stable, 4n+2), anti-aromatic (unstable, 4n, planar), and non-aromatic (not planar or not fully conjugated).

91. Pumice stone is an example of -

- (A) foam
- (B) sol
- (C) gel
- (D) solid sol

Correct Answer: (D) solid sol

Solution:

Step 1: Understanding the Question:

The question asks to classify pumice stone based on the type of colloidal system it represents. A colloidal system is defined by its dispersed phase and dispersion medium.

Step 2: Analyzing Pumice Stone:

Pumice stone is a type of volcanic rock formed when super-heated, highly pressurized rock is rapidly ejected from a volcano. It is extremely porous because gas bubbles are trapped in the lava as it cools and solidifies. Therefore, pumice stone consists of a gaseous phase (the bubbles) dispersed within a solid phase (the rock).

Step 3: Classifying the Colloidal System:

Let's define the options:

- Foam: Dispersed phase = Gas, Dispersion medium = Liquid.
- **Sol:** Dispersed phase = Solid, Dispersion medium = Liquid.
- Gel: Dispersed phase = Liquid, Dispersion medium = Solid.
- Solid Sol: This term can refer to a solid dispersed in a solid, or more broadly, it is used for systems where the dispersion medium is a solid. The specific name for a gas dispersed in a solid is a "solid foam". However, among the given choices, "solid sol" is the best and most common classification for this type of colloid.

Step 4: Final Answer:

Pumice stone, which is a dispersion of gas in a solid, is classified as a solid sol (or a solid foam). This corresponds to option (D).

Quick Tip

Memorizing the table of colloid types is essential for this topic. A simple mnemonic for pumice stone is to think of it as "solidified foam", which leads to the name "solid foam" or "solid sol".

92. Identify the major product obtained in the following reaction:

$$\underbrace{ \left[\operatorname{Ag}(\operatorname{NH}_{3})_{2} \right]^{+} + }_{O}$$

 $3^{-}OH \xrightarrow{\Delta} major product$

Correct Answer: (A)

Solution:

Step 1: Understanding the Question:

The question asks for the major product of the reaction of phthalaldehyde with Tollens' reagent $([Ag(NH_3)_2]^+)$ and hydroxide ions (OH^-) .

Step 2: Key Formula or Approach:

The reactant is phthalaldehyde (benzene-1,2-dicarbaldehyde). This is an aldehyde that lacks α -hydrogens. Such aldehydes, in the presence of a strong base (like OH⁻), undergo the Cannizzaro reaction. Since there are two aldehyde groups on the same molecule, it will be an **intramolecular Cannizzaro reaction**. In this reaction, one aldehyde group is oxidized to a

carboxylic acid (which will exist as its salt, carboxylate, in the basic medium), and the other aldehyde group is reduced to a primary alcohol. While Tollens' reagent is an oxidizing agent, the conditions (presence of OH⁻) are also suitable for the Cannizzaro reaction, which is a characteristic disproportionation reaction for this specific substrate. Given the options, the product of the intramolecular Cannizzaro reaction is the most plausible major product.

Step 3: Detailed Explanation:

The starting material is phthalaldehyde.

Phthalaldehyde
$$\xrightarrow{\text{OH}^-}$$
 Product

In the intramolecular Cannizzaro reaction, one of the -CHO groups undergoes oxidation to -COO⁻, and the other -CHO group undergoes reduction to -CH₂OH.

• Oxidation: $-CHO \rightarrow -COO^-$

• Reduction: $-CHO \rightarrow -CH_2OH$

So, the product formed is the 2-(hydroxymethyl)benzoate ion. The structure is a benzene ring with a -COO⁻ group and a -CH₂OH group at adjacent positions. This corresponds exactly to the structure shown in option (A). The 'OH' shown in the option is part of the alcohol group, and the 'COO' is the carboxylate group.

Step 4: Final Answer:

The major product obtained is the result of an intramolecular Cannizzaro reaction, which is the structure shown in option (A).

Quick Tip

When you see an aldehyde without any α -hydrogens (like formaldehyde or benzaldehyde) reacting with a strong base, immediately think of the Cannizzaro reaction. If two such aldehyde groups are on the same molecule, consider the intramolecular version of this reaction.

93. What fraction of one edge centred octahedral void lies in one unit cell of fcc?

- (A) $\frac{1}{12}$ (B) $\frac{1}{2}$ (C) $\frac{1}{3}$ (D) $\frac{1}{4}$

- Correct Answer: (D) $\frac{1}{4}$

Solution:

Step 1: Understanding the Question:

The question asks for the contribution of a single octahedral void located at the center of an edge to one FCC (face-centered cubic) unit cell.

Step 2: Key Concepts:

In an FCC lattice, octahedral voids are located at two positions: 1. The body center of the cube. 2. The center of each of the 12 edges of the cube. The contribution of any point within the unit cell depends on its location. A point on an edge is shared by the unit cells that meet at that edge.

Step 3: Detailed Explanation:

A cube has 12 edges. An octahedral void is present at the center of each of these edges. An edge of a unit cell in a crystal lattice is shared by four adjacent unit cells. Imagine one unit cell, and look at one of its top edges. It is shared by that unit cell, the one in front, the one to the side, and the one diagonally above and in front. Therefore, any atom or void located at the center of an edge contributes only a fraction of itself to each of the four cells it is a part of. The fraction contributed to one unit cell is $\frac{1}{4}$.

Step 4: Final Answer:

The fraction of one edge-centered octahedral void that lies within a single unit cell is $\frac{1}{4}$.

Quick Tip

Remember the contributions of different lattice points in a cubic unit cell: - Corner: 1/8 - Face Center: 1/2 - Body Center: 1 - Edge Center: 1/4 This applies to atoms, ions, and voids located at these positions.

94. Identify the final product [D] obtained in the following sequence of reactions.

$$\text{CH}_3\text{CHO} \xrightarrow{\text{i) LiAlH}_4} \left[A \right] \xrightarrow{\text{H}_2\text{SO}_4} \left[B \right]$$

$$\xrightarrow{\text{HBr}} [C] \xrightarrow{\text{Na/dry ether}} [D]$$

(1) $HC \equiv C^{\Theta} Na^{+}$

(4) C_4H_{10}

Correct Answer: (B)

Solution:

Step 1: Understanding the Question:

This is a multi-step organic synthesis problem where we need to identify the product at each stage to find the final product [D].

Step 2: Analyzing the Reaction Sequence:

Reaction I: CH₃CHO
$$\xrightarrow{i)$$
 LiAlH₄, $ii)$ H₃O⁺ [A]

Lithium aluminium hydride (LiAlH₄) is a strong reducing agent that reduces aldehydes to primary alcohols. The acidic workup (H_3O^+) protonates the intermediate alkoxide.

CH₃CHO (Acetaldehyde)
$$\xrightarrow{\text{i) LiAlH}_4, \text{ ii) H}_3\text{O}^+}$$
 CH₃CH₂OH (Ethanol, [A])

So, [A] is ethanol.

Reaction II: [A] $\xrightarrow{\mathbf{H}_2\mathbf{SO}_4,\Delta}$ [B]

This is the acid-catalyzed dehydration of an alcohol. Heating ethanol with concentrated sulfuric acid results in the elimination of a water molecule to form an alkene.

$$CH_3CH_2OH$$
 (Ethanol) $\xrightarrow{conc. H_2SO_4, \Delta}$ $CH_2 = CH_2$ (Ethene, [B])

So, [B] is ethene.

Reaction III: [B] $\xrightarrow{\text{HBr}}$ [C]

This is the electrophilic addition of hydrogen bromide (HBr) across the double bond of ethene.

$$CH_2 = CH_2$$
 (Ethene) \xrightarrow{HBr} CH_3CH_2Br (Bromoethane, [C])

So, [C] is bromoethane.

Reaction IV: [C] + $C_6H_5Br \xrightarrow{Na/dry \ ether} [D]$

The final step shows the reaction of [C] (bromoethane, an alkyl halide) with bromobenzene (an aryl halide) in the presence of sodium and dry ether. This is a Wurtz-Fittig reaction, which couples the alkyl and aryl groups.

$$CH_3CH_2Br + C_6H_5Br + 2Na \xrightarrow{dry\ ether} C_6H_5CH_2CH_3 + 2NaBr$$

The final product [D] is ethylbenzene.

Step 3: Identifying the Final Product [D]:

The structure of ethylbenzene corresponds to the one shown in option (B).

- Option (C) is biphenyl, the side product from the reaction of two bromobenzene molecules.
- Option (D) is butane, the side product from the reaction of two bromoethane molecules.

The main cross-coupled product is ethylbenzene.

Step 4: Final Answer:

The final product [D] is ethylbenzene. This corresponds to option (B).

Quick Tip

Recognize named reactions: - Step I: Reduction of aldehyde. - Step II: Dehydration of alcohol. - Step III: Electrophilic addition to alkene. - Step IV: Wurtz-Fittig reaction (coupling of alkyl halide and aryl halide). This is the key to identifying the final product.

- 95. Which of the following statements are INCORRECT?
- A. All the transition metals except scandium form MO oxides which are ionic.
- B. The highest oxidation number corresponding to the group number in transition metal oxides is attained in Sc_2O_3 to Mn_2O_7 .
- C. Basic character increases from V_2O_3 to V_2O_4 to V_2O_5 .
- D. V_2O_4 dissolves in acids to give VO_4^{3-} salts.
- E. CrO is basic but Cr_2O_3 is amphoteric.

Choose the correct answer from the options given below:

- (A) B and C only
- (B) A and E only
- (C) B and D only
- (D) C and D only

Correct Answer: (D) C and D only

Solution:

Step 1: Understanding the Question:

We need to evaluate five statements about the properties of transition metals and their oxides, and identify the combination of statements that are incorrect.

Step 2: Analyzing Each Statement:

- A. All the transition metals except scandium form MO oxides which are ionic. This statement is largely correct in the context that most first-row transition metals form a stable monoxide (MO), e.g., TiO, VO, CrO, MnO, FeO, CoO, NiO, CuO. These oxides, with the metal in a low (+2) oxidation state, are predominantly ionic in character. Scandium does not form a stable ScO, its common oxide is Sc_2O_3 . While covalent character increases across the period and with higher oxidation states, this statement is generally considered **correct** in an introductory context.
- B. The highest oxidation number corresponding to the group number in transition metal oxides is attained in Sc₂O₃ to Mn₂O₇. Sc (Group 3) shows +3 in Sc₂O₃. Ti (Group 4) shows +4 in TiO₂. V (Group 5) shows +5 in V₂O₅. Cr (Group 6) shows +6 in CrO₃. Mn (Group 7) shows +7 in Mn₂O₇. This trend holds true from Scandium to Manganese. Therefore, this statement is **correct**.
- C. Basic character increases from V_2O_3 to V_2O_4 to V_2O_5 . For oxides of a given element, as the oxidation state increases, the covalent character increases and the acidic character increases (conversely, the basic character decreases). V_2O_3 (V^{3+}) is basic. V_2O_4 (V^{4+}) is amphoteric. V_2O_5 (V^{5+}) is acidic. The statement claims that basic character increases, which is the opposite of the actual trend. Therefore, this statement is **incorrect**.
- **D.** V_2O_4 dissolves in acids to give $VO_4{}^{3-}$ salts. V_2O_4 contains Vanadium in the +4 oxidation state. When it dissolves in acid, it forms the vanadyl ion, which is VO^{2+} . The vanadate ion, $VO_4{}^{3-}$, has Vanadium in the +5 oxidation state and is formed from the acidic oxide V_2O_5 in basic solutions. Therefore, this statement is **incorrect**.
- E. CrO is basic but Cr_2O_3 is amphoteric. This follows the same trend as the vanadium oxides. CrO (Cr^{2+}) is basic. Cr_2O_3 (Cr^{3+}) is amphoteric. CrO_3 (Cr^{6+}) is acidic. The statement is **correct**.

Step 3: Identifying the Incorrect Statements:

The statements that are incorrect are C and D.

Step 4: Final Answer:

The option that lists only the incorrect statements is C and D. This corresponds to option (D).

A key trend for metal oxides is that their acidic/basic character depends on the metal's oxidation state. - Low oxidation state \rightarrow Basic oxide (e.g., FeO) - Intermediate oxidation state \rightarrow Amphoteric oxide (e.g., Al₂O₃, Cr₂O₃) - High oxidation state \rightarrow Acidic oxide (e.g., Mn₂O₇, V₂O₅)

96. Consider the following reaction:

$$CH_2-O$$
 HI $A+B$

Identify products A and B.

(1)
$$A = \bigvee$$
 CH₃ and $B = \bigvee$ I

(2)
$$A = \langle CH_3 \text{ and } B = \langle CH_3 \text{ OH } CH_3 \text{ of } B \rangle$$

(3)
$$A = \langle CH_2OH \text{ and } B = \langle DH_2OH \text{ and } B \rangle$$

(4)
$$A = \langle A \rangle$$
 CH₂I and $B = \langle A \rangle$ OH

Correct Answer: (D)

Solution:

Step 1: Understanding the Question:

The question asks for the products of the reaction of an unsymmetrical ether (benzyl phenyl ether) with hydrogen iodide (HI). This reaction is known as the cleavage of ethers.

Step 2: Reaction Mechanism (Cleavage of Ethers with HI):

The reaction proceeds in two steps:

1. **Protonation of the ether oxygen:** The oxygen atom of the ether gets protonated by the H⁺ from HI to form a protonated ether intermediate.

121

$$C_6H_5CH_2-O-C_6H_5 + HI \rightleftharpoons C_6H_5CH_2-O^+(H)-C_6H_5 + I^-$$

- 2. Nucleophilic attack by iodide ion (I⁻): The iodide ion then attacks one of the carbon atoms attached to the oxygen, displacing the other part of the molecule. The attack occurs at the carbon atom which is more susceptible to nucleophilic attack.
- Attack at phenyl carbon (C of C_6H_5): Attack at an sp^2 hybridized carbon of a benzene ring is very difficult and does not occur via S_N2 or S_N1 mechanisms.
- Attack at benzyl carbon (C of CH₂): This carbon is benzylic. The cleavage can proceed via an S_N 2-like mechanism where I^- attacks the less hindered benzylic carbon. Alternatively, it can proceed via an S_N 1-like mechanism because the benzyl carbocation ($C_6H_5CH_2^+$) is highly stabilized by resonance. In either case, the C-O bond at the benzylic position is the one that breaks.

$$I^- + C_6H_5CH_2-O^+(H)-C_6H_5 \to C_6H_5CH_2I + C_6H_5OH$$

Step 3: Identifying the Products A and B:

The products formed are benzyl iodide ($C_6H_5CH_2I$) and phenol (C_6H_5OH).

Comparing this with the given options, we can identify:

- -A = Benzyl iodide
- -B = Phenol

This matches the structures given in option (D).

Step 4: Final Answer:

The products A and B are benzyl iodide and phenol, respectively. This corresponds to option (D).

Quick Tip

When cleaving an ether with HX, if one of the groups is tertiary, benzylic, or allylic, the cleavage usually proceeds via an S_N1 mechanism, and the halide attaches to that group. If both groups are primary or secondary, it's an S_N2 mechanism, and the halide attaches to the less hindered group. A phenyl group attached to the ether oxygen will always form a phenol because the C(phenyl)-O bond is strong and resistant to cleavage.

- 97. The reaction that does NOT take place in a blast furnace between 900 K to 1500 K temperature range during extraction of iron is:
- (A) $CaO + SiO_2 \rightarrow CaSiO_3$
- (B) $Fe_2O_3 + CO \rightarrow 2FeO + CO_2$
- (C) $FeO + CO \rightarrow Fe + CO_2$
- (D) $C + CO_2 \rightarrow 2CO$

Correct Answer: (B) $Fe_2O_3 + CO \rightarrow 2FeO + CO_2$

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the given chemical reactions does not occur in the specified temperature range (900 K - 1500 K) within a blast furnace used for iron extraction.

Step 2: Temperature Zones in a Blast Furnace:

A blast furnace has different temperature zones, and specific reactions occur in each zone.

- Lower Zone (Zone of Combustion, ~ 2000 K): Coke burns to produce CO_2 , which is then reduced to CO. $C + O_2 \rightarrow CO_2$; $CO_2 + C \rightarrow 2CO$.
- Middle Zone (Zone of Slag Formation and Main Reduction, 900 K 1500 K): This is the high-temperature zone in the middle of the furnace. Here, limestone decomposes, slag is formed, and the final reduction of iron oxide to iron occurs.
- Upper Zone (Zone of Reduction, 500 K 800 K): This is the cooler zone at the top. Here, the incoming iron ore is preheated and undergoes initial reduction.

Step 3: Analyzing the Reactions based on Temperature Zones:

- (A) CaO + SiO₂ \rightarrow CaSiO₃: This is the formation of slag (calcium silicate). It occurs at high temperatures (around 1200 K) after the decomposition of limestone (CaCO₃ \rightarrow CaO + CO₂). This reaction **does occur** in the 900 K 1500 K range.
- (B) $\text{Fe}_2\text{O}_3 + \text{CO} \rightarrow 2\text{FeO} + \text{CO}_2$: This is the initial reduction of hematite (Fe₂O₃) to iron(II) oxide (FeO). This reaction takes place in the upper, cooler part of the furnace at temperatures around 500 K 800 K. It **does NOT occur** in the higher temperature range of 900 K 1500 K.
- (C) FeO + CO → Fe + CO₂: This is the final reduction of iron(II) oxide to molten iron. This is the main reduction step and requires higher temperatures, occurring within the 900 K
 1500 K range. This reaction does occur in this range.
- (D) C + CO₂ \rightarrow 2CO: This is the Boudouard reaction, where carbon dioxide is reduced by coke to form carbon monoxide, the primary reducing agent. This reaction is endothermic and is favored at high temperatures (above 1000 K). This reaction **does occur** in the 900 K 1500 K range.

Step 4: Final Answer:

The reduction of Fe_2O_3 to FeO occurs at lower temperatures than the 900 K - 1500 K range. Therefore, this is the correct answer. This corresponds to option (B).

Quick Tip

Remember the general sequence of reduction in a blast furnace from top to bottom (cooler to hotter): $Fe_2O_3 \rightarrow Fe_3O_4 \rightarrow FeO \rightarrow Fe$. The initial reduction steps occur at lower temperatures, while the final reduction to molten iron happens at higher temperatures.

98. Which amongst the following will be most readily dehydrated under acidic conditions?

(1)
$$\begin{array}{c} NO_2 \\ OH \\ NO_2 \\ OH \\ CH_3 \\ OH \\ OH \\ OH \\ OH \\ OH \\ \end{array}$$

Correct Answer: (C)

Solution:

Step 1: Understanding the Question:

The question asks which of the given alcohols will undergo acid-catalyzed dehydration most readily.

Step 2: Key Concept - Mechanism of Dehydration:

The acid-catalyzed dehydration of secondary and tertiary alcohols typically proceeds through an E1 elimination mechanism. The rate-determining step of this mechanism is the formation of a carbocation intermediate after the loss of a water molecule from the protonated alcohol. The readiness of an alcohol to dehydrate via the E1 mechanism is directly related to the stability of the carbocation it forms.

Carbocation Stability Order: $3^{\circ} > 2^{\circ} > 1^{\circ}$. Stability is increased by electron-donating groups (like alkyl groups, +I effect, hyperconjugation) and decreased by electron-withdrawing groups (like -NO₂, -I effect).

Step 3: Analyzing the Carbocation Formed from Each Alcohol:

Let's consider the most stable carbocation that can be formed from each starting material:

- (A) 4-nitropentan-2-ol: Dehydration will form a secondary (2°) carbocation at C-2. This carbocation is destabilized by the strong electron-withdrawing inductive effect (-I) of the nitro group (-NO₂) at C-4.
- (B) 4-nitropentane-2,3-diol: Dehydration of either hydroxyl group would lead to a carbocation that is destabilized by the electron-withdrawing -NO₂ group.
- (C) 4-methylpentane-2,3-diol: Dehydration of the hydroxyl group at C-3 would form a secondary (2°) carbocation. This carbocation is stabilized by the electron-donating inductive effect (+I) and hyperconjugation from the surrounding alkyl groups (a methyl group and an isopropyl group). There are no electron-withdrawing groups to destabilize it.
- (D) 1-nitropentane-2,3-diol: Dehydration of either hydroxyl group will form a carbocation that is significantly destabilized by the strong electron-withdrawing -NO₂ group, especially if the carbocation is at C-2 (adjacent to the -NO₂ group).

Step 4: Comparing Stability and Reactivity:

The carbocation formed from compound (C) is the most stable among all the options because it is stabilized by electron-donating alkyl groups and lacks any destabilizing electron-withdrawing groups. Since the rate of dehydration depends on the stability of the carbocation intermediate, compound (C) will dehydrate the most readily.

Step 5: Final Answer:

4-methylpentane-2,3-diol will be most readily dehydrated as it forms the most stable carbocation intermediate. This corresponds to option (C).

Quick Tip

When predicting the rate of reactions that proceed through carbocation intermediates (like S_N1 and E1), always analyze the stability of the potential carbocations. Look for factors that stabilize (alkyl groups, resonance) or destabilize (electron-withdrawing groups) the positive charge.

- 99. Which amongst the following options is the correct relation between change in enthalpy and change in internal energy?
- $(A) \Delta H + \Delta U = \Delta n R$
- (B) $\Delta H = \Delta U \Delta n_g RT$
- (C) $\Delta H = \Delta U + \Delta n_g RT$
- (D) $\Delta H \Delta U = -\Delta nRT$

Correct Answer: (C) $\Delta H = \Delta U + \Delta n_g RT$

Solution:

Step 1: Understanding the Question:

The question asks for the fundamental thermodynamic relationship between enthalpy change (ΔH) and internal energy change (ΔU) .

Step 2: Derivation of the Relationship:

The definition of enthalpy (H) is given by:

$$H = U + PV$$

where U is internal energy, P is pressure, and V is volume. For a change in the system, the change in enthalpy (ΔH) is:

$$\Delta H = \Delta U + \Delta (PV)$$

If the reaction occurs at constant pressure, this becomes:

$$\Delta H = \Delta U + P\Delta V$$

This equation relates ΔH and ΔU for any process at constant pressure. For chemical reactions involving gases, we can express the $P\Delta V$ term differently.

Assuming the gases behave ideally, the ideal gas law states PV = nRT.

The term $P\Delta V$ can be related to the change in the number of moles of gas (Δn_q) .

Let V_i and V_f be the initial and final volumes of the gaseous components, and n_i and n_f be the initial and final number of moles of gas.

$$PV_f = n_f RT$$

$$PV_i = n_i RT$$

Subtracting these equations gives:

$$P(V_f - V_i) = (n_f - n_i)RT$$
$$P\Delta V = \Delta n_q RT$$

where $\Delta n_g =$ (moles of gaseous products) - (moles of gaseous reactants). Substituting $P\Delta V = \Delta n_g RT$ into the enthalpy equation gives the desired relationship:

$$\Delta H = \Delta U + \Delta n_a RT$$

Step 3: Comparing with Options:

The derived equation, $\Delta H = \Delta U + \Delta n_q RT$, exactly matches the expression given in option (C).

Step 4: Final Answer:

The correct relation is $\Delta H = \Delta U + \Delta n_g RT$. This corresponds to option (C).

Quick Tip

Remember that H comes after U alphabetically, and H is generally larger than U (for reactions with expanding gases). This can help you remember the plus sign in the equation $\Delta H = \Delta U + \Delta n_q RT$.

100. Given below are two statements:

Statement I: The nutrient deficient water bodies lead to eutrophication.

Statement II: Eutrophication leads to decrease in the level of oxygen in the water bodies.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is incorrect but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is correct but Statement II is false.

Correct Answer: (A) Statement I is incorrect but Statement II is true.

Solution:

Step 1: Understanding Eutrophication:

Eutrophication is a process of ecological imbalance in water bodies. It is crucial to understand its cause and effect to evaluate the given statements.

Step 2: Analyzing Statement I:

"The nutrient deficient water bodies lead to eutrophication."

Eutrophication is defined as the enrichment of water bodies with nutrients, especially nitrogen and phosphorus. These nutrients act as fertilizers and stimulate excessive growth of algae and other aquatic plants (a phenomenon known as an algal bloom). Therefore, eutrophication is caused by an excess of nutrients, not a deficiency. Water bodies that are naturally low in nutrients are called oligotrophic. Thus, Statement I is **incorrect**.

Step 3: Analyzing Statement II:

"Eutrophication leads to decrease in the level of oxygen in the water bodies."

The massive algal blooms caused by eutrophication eventually die and sink to the bottom of the water body. Aerobic bacteria then decompose this large amount of dead organic matter. This decomposition process consumes a significant amount of the dissolved oxygen in the water. This depletion of oxygen (hypoxia or anoxia) can lead to the death of fish and other aquatic animals. Thus, Statement II is **correct**.

Step 4: Final Answer:

Based on the analysis, Statement I is incorrect, and Statement II is true. This corresponds to option (A).

Remember the cause-and-effect chain of eutrophication: 1. Cause: Excess nutrients (like phosphates from detergents, nitrates from fertilizers).

- 2. Immediate Effect: Algal bloom (excessive plant growth).
- 3. Consequence: Death and decomposition of algae by bacteria.
- 4. Final Result: Depletion of dissolved oxygen, leading to the death of aquatic life.

Botany

101. Identify the pair of heterosporous pteridophytes among the following:

- (A) Equisetum and Salvinia
- (B) Lycopodium and Selaginella
- (C) Selaginella and Salvinia
- (D) Psilotum and Salvinia

Correct Answer: (C) Selaginella and Salvinia

Solution:

Step 1: Understanding the Question:

The question asks to identify a pair of pteridophytes that are heterosporous. Heterosporous plants produce two different types of spores: smaller microspores (male) and larger megaspores (female). This is in contrast to homosporous plants, which produce only one type of spore.

Step 2: Detailed Explanation:

Let's analyze the options:

- 1. **Equisetum:** It is a homosporous pteridophyte.
- 2. **Lycopodium:** Most species are homosporous (some are heterosporous, but generally it's considered homosporous in this context).
- 3. **Selaginella:** It is a classic example of a heterosporous pteridophyte.
- 4. Salvinia: It is an aquatic fern and is also heterosporous.
- 5. **Psilotum:** It is a homosporous pteridophyte.

Based on this analysis:

- (A) Equisetum (homosporous) and Salvinia (heterosporous) Incorrect pair.
- (B) Lycopodium (homosporous) and Selaginella (heterosporous) Incorrect pair.
- (C) Selaginella (heterosporous) and Salvinia (heterosporous) Correct pair.
- (D) Psilotum (homosporous) and Salvinia (heterosporous) Incorrect pair.

Step 3: Final Answer:

The correct pair of heterosporous pteridophytes is Selaginella and Salvinia.

Remember the key examples for heterospory in pteridophytes: Selaginella, Salvinia, Marsilea, and Azolla. Most other common pteridophytes like Pteris, Dryopteris, Equisetum, and Lycopodium are homosporous. This distinction is a precursor to the seed habit seen in higher plants.

102. Cellulose does not form blue colour with Iodine because

- (A) It breakes down when iodine reacts with it.
- (B) It is a disaccharide.
- (C) It is a helical molecule.
- (D) It does not contain complex helices and hence cannot hold iodine molecules.

Correct Answer: (D) It does not contain complex helices and hence cannot hold iodine molecules.

Solution:

Step 1: Understanding the Question:

The question asks for the reason why cellulose does not give a positive iodine test (blue-black color), unlike starch.

Step 2: Detailed Explanation:

The iodine test is specific for the presence of starch. Starch is a polymer of α -glucose and consists of two components: amylose and amylopectin. Amylose has a helical secondary structure. When iodine is added, the iodine molecules (I_2) get trapped inside these helices, forming a starch-iodine complex which appears blue-black in color.

Cellulose, on the other hand, is a linear polymer of β -glucose units linked by β -1,4 glycosidic bonds. It does not form a helical structure. Instead, the long, straight chains of cellulose are arranged parallel to each other, forming strong hydrogen bonds between them. This rigid, linear structure does not have the helical cavities necessary to trap iodine molecules. Therefore, cellulose cannot form the colored complex with iodine.

Let's evaluate the options:

- (A) It does not break down in the presence of iodine.
- (B) It is a polysaccharide, not a disaccharide.
- (C) It is a linear molecule, not a helical one.
- (D) This is the correct explanation. The absence of complex helices prevents the trapping of iodine molecules.

Step 3: Final Answer:

Cellulose does not give a blue color with iodine because its linear structure lacks the complex

helices required to hold iodine molecules.

Quick Tip

Associate the starch-iodine test with the helical structure of amylose. Visualize the iodine molecules fitting into the coils of the helix. Since cellulose is a straight chain, there are no coils to trap the iodine.

103. Frequency of recombination between gene pairs on same chromosome as a measure of the distance between genes to map their position on chromosome, was used for the first time by

- (A) Henking
- (B) Thomas Hunt Morgan
- (C) Sutton and Boveri
- (D) Alfred Sturtevant

Correct Answer: (D) Alfred Sturtevant

Solution:

Step 1: Understanding the Question:

The question asks to identify the scientist who first used the frequency of genetic recombination to map the positions of genes on a chromosome.

Step 2: Detailed Explanation:

Let's review the contributions of the scientists listed:

- Henking (1891): Discovered the X chromosome while studying insect spermatogenesis, referring to it as the 'X-body'.
- Thomas Hunt Morgan (early 1900s): Through his work on *Drosophila melanogaster*, he provided experimental evidence for the chromosomal theory of inheritance, discovered linkage, and showed that genes are located on chromosomes.
- Sutton and Boveri (1902): Independently proposed the Chromosomal Theory of Inheritance, which states that genes are found at specific locations on chromosomes.
- Alfred Sturtevant (1913): He was a student in T.H. Morgan's lab. He hypothesized that the frequency of recombination between two linked genes is proportional to the physical distance between them on the chromosome. Using this principle, he constructed the first-ever genetic map for the X chromosome of *Drosophila*.

Therefore, Alfred Sturtevant was the first to use recombination frequency for gene mapping.

Step 3: Final Answer:

Alfred Sturtevant was the scientist who first utilized recombination frequencies to create a ge-

netic map.

Quick Tip

Remember the hierarchy of discovery: Sutton & Boveri proposed the theory, Morgan provided the experimental proof with fruit flies, and Sturtevant (Morgan's student) took it a step further by creating the first gene map based on recombination frequencies.

104. In tissue culture experiments, leaf mesophyll cells are put in a culture medium to form callus. This phenomenon may be called as:

- (A) Senescence
- (B) Differentiation
- (C) Dedifferentiation
- (D) Development

Correct Answer: (C) Dedifferentiation

Solution:

Step 1: Understanding the Question:

The question describes a process in plant tissue culture where specialized, mature cells (leaf mesophyll) are induced to form an undifferentiated mass of cells called a callus. We need to identify the correct biological term for this process.

Step 2: Detailed Explanation:

Let's define the terms in the options:

- **Senescence:** The process of aging in living organisms.
- **Differentiation:** The process by which cells become specialized in structure and function. For example, a meristematic cell differentiating into a mesophyll cell.
- **Dedifferentiation:** The process by which mature, differentiated cells revert to a meristematic state, losing their specialization and regaining the capacity for cell division. This is exactly what happens when mesophyll cells form a callus.
- **Development:** The sum of all changes that an organism goes through during its life cycle, from germination to senescence. It is a very broad term.

The phenomenon described, where differentiated mesophyll cells form an undifferentiated callus, is correctly termed dedifferentiation.

Step 3: Final Answer:

The process of forming a callus from differentiated leaf mesophyll cells is called dedifferentiation.

Remember the sequence in plant tissue culture: 1. **Differentiation**: Forms the initial plant part (explant). 2. **Dedifferentiation**: The explant cells form a callus (undifferentiated mass). 3. **Redifferentiation**: The callus cells differentiate to form new organs like roots and shoots.

105. Among 'The Evil Quartet', which one is considered the most important cause driving extinction of species?

- (A) Co-extinctions
- (B) Habitat loss and fragmentation
- (C) Over exploitation for economic gain
- (D) Alien species invasions

Correct Answer: (B) Habitat loss and fragmentation

Solution:

Step 1: Understanding the Question:

The question asks to identify the primary cause of species extinction from the four major causes known collectively as 'The Evil Quartet'.

Step 2: Detailed Explanation:

'The Evil Quartet' refers to the four major causes of biodiversity loss:

- 1. Habitat loss and fragmentation: This involves the destruction or division of natural habitats due to activities like deforestation, urbanization, and agriculture. It is considered the single most important driver of extinction because it directly removes the places where species live, find food, and reproduce.
- 2. Over-exploitation: This is the harvesting of species from the wild at rates faster than natural populations can recover (e.g., overfishing, overhunting).
- 3. Alien species invasions: The introduction of non-native species into an ecosystem can disrupt the local food web, outcompete native species for resources, and introduce diseases, leading to extinctions.
- 4. **Co-extinctions:** This occurs when the extinction of one species causes the extinction of another species that depended on it (e.g., a host and its specific parasite).

Among these four, scientists widely agree that habitat loss and fragmentation is the most significant cause of species extinction worldwide.

Step 3: Final Answer:

Habitat loss and fragmentation is considered the most important cause driving the extinction of species.

While all four components of 'The Evil Quartet' are significant threats, always remember that habitat loss is the leading cause. If a species has nowhere to live, it cannot survive. This is the foundational threat upon which others can act.

106. Upon exposure to UV radiation, DNA stained with ethidium bromide will show

- (A) Bright orange colour
- (B) Bright red colour
- (C) Bright blue colour
- (D) Bright yellow colour

Correct Answer: (A) Bright orange colour

Solution:

Step 1: Understanding the Question:

The question asks about the appearance of DNA stained with ethidium bromide when viewed under UV light. This is a standard technique used in molecular biology.

Step 2: Detailed Explanation:

Agarose gel electrophoresis is a common method used to separate DNA fragments based on their size. However, DNA is colorless and not visible to the naked eye. To visualize the DNA bands in the gel, a fluorescent dye is used.

The most common dye is Ethidium Bromide (EtBr). EtBr works by intercalating, or inserting itself, between the base pairs of the DNA double helix. When the agarose gel containing the DNA and EtBr is exposed to ultraviolet (UV) radiation, the EtBr molecules absorb the UV light and fluoresce, emitting light in the visible spectrum. This fluorescence appears as a bright orange color. This allows researchers to see the separated DNA fragments as distinct orange bands.

Step 3: Final Answer:

DNA stained with ethidium bromide fluoresces with a bright orange color under UV radiation.

Quick Tip

This is a fundamental fact in molecular biology labs. Remember the combination: DNA + Ethidium Bromide (EtBr) + UV light = Bright Orange Bands. This is a frequently asked question in exams.

107. How many ATP and NADPH₂ are required for the synthesis of one molecule of Glucose during Calvin cycle?

- (A) 18 ATP and 16 NADPH₂
- (B) 12 ATP and 12 NADPH₂
- (C) 18 ATP and 12 NADPH₂
- (D) 12 ATP and 16 NADPH $_2$

Correct Answer: (C) 18 ATP and 12 NADPH₂

Solution:

Step 1: Understanding the Question:

The question asks for the total number of ATP and NADPH₂ molecules required to synthesize one molecule of glucose via the Calvin cycle.

Step 2: Key Formula or Approach:

The synthesis of one molecule of glucose $(C_6H_{12}O_6)$ requires the fixation of 6 molecules of carbon dioxide (CO_2) . The Calvin cycle must therefore turn 6 times to produce one glucose molecule. We need to determine the ATP and NADPH₂ consumption for a single turn and then multiply by 6.

Step 3: Detailed Explanation:

The Calvin cycle has three main stages for each molecule of CO₂ fixed:

- 1. **Carboxylation:** One molecule of CO₂ is fixed to one molecule of RuBP. No ATP or NADPH₂ is used here.
- 2. **Reduction:** The two molecules of 3-PGA formed are reduced to two molecules of glyceraldehyde-3-phosphate (G3P). This step uses **2 ATP** and **2 NADPH**₂ per CO₂ molecule.
- 3. **Regeneration:** One molecule of RuBP is regenerated from G3P. This step uses **1 ATP** per CO₂ molecule.

So, for one turn of the Calvin cycle (fixing one CO_2):

- Total ATP used = 2 ATP (in reduction) + 1 ATP (in regeneration) = 3 ATP.
- Total NADPH₂ used = 2 NADPH_2 (in reduction).

To synthesize one molecule of glucose (C_6) , the cycle must turn 6 times.

- Total ATP required = 6 turns \times 3 ATP/turn = 18 ATP.
- Total NADPH₂ required = 6 turns \times 2 NADPH₂/turn = **12 NADPH**₂.

Step 4: Final Answer:

The synthesis of one molecule of glucose requires 18 ATP and 12 NADPH₂.

To avoid confusion, always calculate the energy requirement for a single CO₂ fixation first (3 ATP, 2 NADPH₂) and then multiply by the number of carbons in the final product (e.g., 6 for glucose). The options are often designed to confuse, so be precise with your calculation.

108. Large, colourful, fragrant flowers with nectar are seen in:

- (A) wind pollinated plants
- (B) insect pollinated plants
- (C) bird pollinated plants
- (D) bat pollinated plants

Correct Answer: (B) insect pollinated plants

Solution:

Step 1: Understanding the Question:

The question asks to identify the type of pollination associated with flowers that are large, colorful, fragrant, and produce nectar.

Step 2: Detailed Explanation:

These characteristics are adaptations to attract pollinators. Let's analyze the pollination types:

- Wind pollinated plants (Anemophily): These plants do not need to attract pollinators. Their flowers are typically small, inconspicuous, not colorful, lack nectar and fragrance. They produce large amounts of lightweight pollen. Examples include grasses and maize.
- Insect pollinated plants (Entomophily): These plants rely on insects for pollination. To attract insects, their flowers have evolved to be large, brightly colored (especially in the blue-violet range visible to bees), fragrant, and they produce nectar as a food reward for the visiting insect.
- Bird pollinated plants (Ornithophily): These flowers are often large, brightly colored (typically red or orange), but usually lack a strong fragrance because birds have a poor sense of smell. They produce copious amounts of dilute nectar.
- Bat pollinated plants (Chiropterophily): These flowers are typically large, pale or white, open at night, and have a strong, musty, or fruity odor. They also produce abundant nectar.

The combination of all four traits - large, colorful, fragrant, and nectar-rich - is the classic suite of characteristics for insect-pollinated plants.

Step 3: Final Answer:

Large, colourful, fragrant flowers with nectar are characteristic features of insect pollinated plants.

Create a mental table linking pollinating agents to floral characteristics. For example: Wind -; small, dull, no nectar/scent; Insects -; colorful, fragrant, nectar; Birds -; bright (often red), no scent, lots of nectar; Bats -; pale, nocturnal, musty scent.

109. Which of the following stages of meiosis involves division of centromere?

- (A) Telophase
- (B) Metaphase I
- (C) Metaphase II
- (D) Anaphase II

Correct Answer: (D) Anaphase II

Solution:

Step 1: Understanding the Question:

The question asks to identify the specific stage of meiosis where the centromere splits, allowing sister chromatids to separate.

Step 2: Detailed Explanation:

Let's review the key events of the meiotic stages:

- Meiosis I: This is the reductional division. Metaphase I: Homologous chromosome pairs (bivalents) align at the metaphase plate. The centromeres do not divide. Anaphase I: Homologous chromosomes separate and move to opposite poles. Sister chromatids remain attached at their centromeres.
- Meiosis II: This is the equational division, very similar to mitosis. Metaphase II: Individual chromosomes (each with two sister chromatids) align at the metaphase plate. Anaphase II: The centromeres of each chromosome finally divide. This allows the sister chromatids to separate and move to opposite poles. Once separated, they are considered individual chromosomes. Telophase I II: Chromosomes decondense and nuclear envelopes reform. Division of centromeres does not occur in this stage.

Therefore, the division of the centromere occurs during Anaphase II of meiosis.

Step 3: Final Answer:

The division of the centromere occurs during Anaphase II.

A key difference between Meiosis I and Meiosis II is the behavior of centromeres. Remember: Anaphase I separates homologous chromosomes, but sister chromatids stay together (no centromere split). Anaphase II separates sister chromatids (centromere splits). Anaphase II is analogous to Anaphase of mitosis.

110. During the purification process for recombinant DNA technology, addition of chilled ethanol precipitates out

- (A) Polysaccharides
- (B) RNA
- (C) DNA
- (D) Histones

Correct Answer: (C) DNA

Solution:

Step 1: Understanding the Question:

The question concerns a specific step in the isolation and purification of DNA, which is a fundamental procedure in recombinant DNA technology. It asks what macromolecule is precipitated by adding chilled ethanol.

Step 2: Detailed Explanation:

The process of isolating DNA from cells involves several steps:

- 1. Lysis: Breaking open the cells to release their contents, including DNA.
- 2. **Removal of contaminants:** The cell lysate contains DNA, RNA, proteins, lipids, and polysaccharides. Enzymes are used to digest these contaminants. For example, Ribonuclease (RNase) is used to remove RNA, and Protease is used to remove proteins (like histones).
- 3. **DNA Precipitation:** After removing other macromolecules, the goal is to isolate the pure DNA. DNA is soluble in the aqueous solution, but it is insoluble in alcohol (like ethanol or isopropanol). When chilled ethanol is added to the solution, the DNA precipitates out of the solution, forming a visible mass of fine, white threads. This process is called ethanol precipitation. The low temperature reduces the solubility further. The precipitated DNA can then be physically separated from the solution, for example, by spooling it onto a glass rod.

Step 3: Final Answer:

The addition of chilled ethanol causes the DNA to precipitate out of the solution.

Remember the key phrase "chilled ethanol precipitation". It is a standard and very specific method for isolating DNA. The alcohol makes the DNA come out of solution so it can be collected.

- 111. Family Fabaceae differs from Solanaceae and Liliaceae. With respect to the stamens, pick out the characteristics specific to family Fabaceae but not found in Solanaceae or Liliaceae.
- (A) Epiphyllous and Dithecous anthers
- (B) Diadelphous and Dithecous anthers
- (C) Polyadelphous and epipetalous stamens
- (D) Monoadelphous and Monothecous anthers

Correct Answer: (B) Diadelphous and Dithecous anthers

Solution:

Step 1: Understanding the Question:

The question asks to identify a characteristic of the stamens (androecium) that is specific to the family Fabaceae and not found in Solanaceae or Liliaceae.

Step 2: Detailed Explanation:

Let's analyze the characteristics of stamens in these three families:

- **Fabaceae** (**Pea family**): The stamens are typically ten in number. A key feature is that they are often **diadelphous**, meaning they are fused into two bundles. The common arrangement is (9)+1, where nine stamens are fused to form a tube, and one is free. The anthers are **dithecous** (having two lobes).
- Solanaceae (Potato family): The stamens are typically five in number. They are epipetalous, meaning they are attached to the petals. The anthers are dithecous. Fusion of stamens (adelphy) is not a characteristic feature.
- Liliaceae (Lily family): The stamens are typically six, arranged in two whorls of three. They are often **epiphyllous** (also called epitepalous), meaning they are attached to the tepals (undifferentiated petals and sepals). The anthers are dithecous.

Now let's evaluate the options based on this information:

- (A) Epiphyllous and Dithecous anthers: Epiphyllous condition is characteristic of Liliaceae, not Fabaceae.
- (B) Diadelphous and Dithecous anthers: The diadelphous condition ((9)+1) is a hallmark of many species in Fabaceae. This condition is not found in Solanaceae or Liliaceae. Dithecous anthers are common to all three, but the diadelphous condition is specific.
- (C) Polyadelphous and epipetalous stamens: Polyadelphous (fused into many bundles) is seen in families like Rutaceae (Citrus). Epipetalous condition is seen in Solanaceae, not Fabaceae.
- (D) Monoadelphous and Monothecous anthers: Monoadelphous (fused into one bundle) is seen

in Malvaceae (China rose). Monothecous anthers (one lobe) are also seen in Malvaceae, not Fabaceae.

Step 3: Final Answer:

The combination of diadelphous stamens and dithecous anthers is specific to Fabaceae among the given choices.

Quick Tip

For floral formulas, remember these key stamen characteristics: Fabaceae - \dot{i} Diadelphous ((9)+1); Solanaceae - \dot{i} Epipetalous; Liliaceae - \dot{i} Epiphyllous/Epitepalous; Malvaceae - \dot{i} Monoadelphous and Monothecous. These are high-yield points for family-based questions.

112. Axile placentation is observed in

- (A) China rose, Petunia and Lemon
- (B) Mustard, Cucumber and Primrose
- (C) China rose, Beans and Lupin
- (D) Tomato, Dianthus and Pea

Correct Answer: (A) China rose, Petunia and Lemon

Solution:

Step 1: Understanding the Question:

The question asks to identify the group of plants that all exhibit axile placentation. Placentation refers to the arrangement of ovules within the ovary.

Step 2: Detailed Explanation:

In **axile placentation**, the ovary is partitioned into two or more chambers (locules) by septa. The placenta, where the ovules are attached, is located on the central axis where the septa meet. A cross-section of an ovary with axile placentation looks like the segments of an orange or lemon.

Let's examine the placentation types for the plants in each option:

- (A) China rose, Petunia and Lemon: China rose (Hibiscus, family Malvaceae): Shows axile placentation. Petunia (family Solanaceae): Shows axile placentation. Lemon (Citrus, family Rutaceae): Shows axile placentation. This group is consistent.
- **(B) Mustard, Cucumber and Primrose: Mustard** (Brassica, family Brassicaceae): Shows parietal placentation. **Cucumber** (Cucurbita, family Cucurbitaceae): Shows parietal placentation. **Primrose** (Primula, family Primulaceae): Shows free-central placentation. This group is inconsistent.
- (C) China rose, Beans and Lupin: China rose: Shows axile placentation. Beans and Lupin (family Fabaceae): Show marginal placentation. This group is inconsistent.

- **(D) Tomato, Dianthus and Pea:** - **Tomato** (Lycopersicon, family Solanaceae): Shows axile placentation. - **Dianthus** (family Caryophyllaceae): Shows free-central placentation. - **Pea** (Pisum, family Fabaceae): Shows marginal placentation. This group is inconsistent.

Step 3: Final Answer:

The correct group of plants, all of which exhibit axile placentation, is China rose, Petunia, and Lemon.

Quick Tip

Memorize at least one clear example for each type of placentation: - Marginal: Pea (Fabaceae) - Axile: Tomato, China rose, Lemon - Parietal: Mustard, Argemone, Cucumber - Free-central: Dianthus, Primrose - Basal: Sunflower, Marigold (Asteraceae) Visualizing a cross-section of these fruits/ovaries helps in remembering.

113. What is the function of tassels in the corn cob?

- (A) To protect seeds
- (B) To attract insects
- (C) To trap pollen grains
- (D) To disperse pollen grains

Correct Answer: (C) To trap pollen grains

Solution:

Step 1: Understanding the Question:

The question asks for the function of the "tassels" in a corn cob. It's important to correctly identify the parts. The corn plant has a male inflorescence (the tassel at the top) and a female inflorescence (the ear or cob, which develops into the fruit). The question seems to confuse "tassels" with the "silks" emerging from the corn cob. The tassel is the pollen-producing structure at the top of the plant. The silks are the structures emerging from the tip of the ear/cob. Given the options, the question is most likely referring to the silks of the corn cob, not the tassel at the top of the plant.

Step 2: Detailed Explanation:

Let's clarify the terminology and function: - **Tassel:** This is the male inflorescence located at the apex of the corn plant. Its function is to produce and **disperse pollen grains** (Option D). - **Corn Cob (Ear):** This is the female inflorescence. - **Silks:** These are the long, thread-like styles that emerge from the tip of the cob. Each silk is connected to a potential kernel (ovule). The ends of the silks are feathery and sticky. Their function is to **trap the airborne pollen grains** (Option C) released from the tassels.

Since the question asks about the "tassels in the corn cob" and provides "To trap pollen grains" as an option, it is using incorrect terminology but referring to the function of the silks. The silks are the part of the female flower (on the cob) responsible for trapping pollen. If the question

strictly meant the tassel (male flower), the answer would be (D) to disperse pollen. However, the context "in the corn cob" points towards the silks. The answer key confirms the intended question is about the silks.

Therefore, the function of the silks (referred to as tassels in the question) is to trap pollen grains.

Step 3: Final Answer:

The structures emerging from the corn cob (the silks, which the question calls tassels) function to trap pollen grains.

Quick Tip

Be aware of potential ambiguities in questions. In corn (maize), the Tassel is the male flower at the top (disperses pollen), and the Silk is the female style on the cob (traps pollen). Maize is wind-pollinated (anemophilous), so the silks are long and feathery to effectively catch wind-borne pollen.

114. The process of appearance of recombination nodules occurs at which sub stage of prophase I in meiosis?

- (A) Diakinesis
- (B) Zygotene
- (C) Pachytene
- (D) Diplotene

Correct Answer: (C) Pachytene

Solution:

Step 1: Understanding the Question:

The question asks to identify the substage of Prophase I of meiosis where recombination nodules are observed. Recombination nodules are the sites where crossing over occurs.

Step 2: Detailed Explanation:

Prophase I is the longest phase of meiosis and is divided into five substages:

- 1. **Leptotene:** Chromosomes start to condense and become visible.
- 2. **Zygotene:** Homologous chromosomes pair up in a process called synapsis, forming bivalents. The synaptonemal complex begins to form.
- 3. **Pachytene:** Synapsis is complete. The paired chromosomes (bivalents) are clearly visible as tetrads (four chromatids). This is the stage where **crossing over** (exchange of genetic material between non-sister chromatids of homologous chromosomes) occurs. The sites of crossing over are marked by the appearance of **recombination nodules**.
- 4. **Diplotene:** The synaptonemal complex dissolves, and the homologous chromosomes start to separate from each other, except at the points of crossing over. These X-shaped structures are called chiasmata.

5. **Diakinesis:** Chromosomes become fully condensed. The chiasmata terminalize (move towards the ends of the chromatids), and the nuclear envelope breaks down.

Therefore, recombination nodules, which mediate crossing over, appear during the pachytene stage.

Step 3: Final Answer:

The appearance of recombination nodules occurs during the pachytene substage of prophase I.

Quick Tip

Use a mnemonic to remember the order of Prophase I stages: "Lazy Zebra Painted Dotted Diagrams" (Leptotene, Zygotene, Pachytene, Diplotene, Diakinesis). Associate Pachytene with "packing" together and "patching" (crossing over).

115. The thickness of ozone in a column of air in the atmosphere is measured in terms of :

- (A) Kilobase
- (B) Dobson units
- (C) Decibels
- (D) Decameter

Correct Answer: (B) Dobson units

Solution:

Step 1: Understanding the Question:

The question asks for the unit of measurement for the thickness of the ozone layer in the atmosphere.

Step 2: Detailed Explanation:

Let's analyze the units given in the options:

- Kilobase (kb): A unit of length for DNA or RNA molecules, equal to 1000 base pairs. Not used for atmospheric measurements.
- **Dobson Units (DU):** This is the standard unit used to measure the total amount of ozone in a vertical column of the atmosphere. One Dobson Unit is the number of molecules of ozone that would be required to create a layer of pure ozone 0.01 millimeters thick at a temperature of 0 degrees Celsius and a pressure of 1 atmosphere.
- **Decibels** (dB): A logarithmic unit used to measure sound level or the power level of an electrical signal.
- **Decameter (dam):** A unit of length equal to 10 meters. Not used for measuring atmospheric gas concentration.

The correct unit for measuring the thickness of the ozone layer is the Dobson Unit, named after Gordon Dobson, who was a pioneer in ozone research.

Step 3: Final Answer:

The thickness of the ozone layer is measured in Dobson units (DU).

Quick Tip

Associate "Ozone" with "Dobson". The thinning of the ozone layer over Antarctica is often described as the "ozone hole," which is defined as an area where the ozone column thickness is less than 220 Dobson Units.

116. The reaction centre in PS II has an absorption maxima at

- (A) 780 nm
- (B) 680 nm
- (C) 700 nm
- (D) 660 nm

Correct Answer: (B) 680 nm

Solution:

Step 1: Understanding the Question:

The question asks for the specific wavelength of light at which the reaction center of Photosystem II (PS II) shows its maximum absorption.

Step 2: Detailed Explanation:

In photosynthesis, there are two photosystems, PS I and PS II, which work in series. Each photosystem consists of a reaction center surrounded by light-harvesting complexes (antenna molecules).

- **Photosystem I (PS I):** The reaction center of PS I is a special chlorophyll *a* molecule called **P700**, which has an absorption peak at 700 nm (far-red light).
- **Photosystem II (PS II):** The reaction center of PS II is a special chlorophyll *a* molecule called **P680**, which shows maximum absorption at a wavelength of 680 nm (red light).

The names P680 and P700 directly correspond to their absorption maxima in nanometers.

Step 3: Final Answer:

The reaction centre in PS II has an absorption maximum at 680 nm.

Remember that PS II comes before PS I in the electron transport chain (Z-scheme), even though its number is higher. Associate PS II with the lower wavelength (680 nm) and PS I with the higher wavelength (700 nm). A simple trick: II comes before I in the alphabet if you read backwards, just like 680 comes before 700.

117. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: The first stage of gametophyte in the life cycle of moss is protonema stage.

Reason R: Protonema develops directly from spores produced in capsule.

In the light of the above statements, choose the most appropriate answer from the options given below:

- (A) A is not correct but R is correct.
- (B) Both A and R are correct and R is the correct explanation of A.
- (C) Both A and R are correct but R is NOT the correct explanation of A.
- (D) A is correct but R is not correct.

Correct Answer: (B) Both A and R are correct and R is the correct explanation of A.

Solution:

Step 1: Understanding the Question:

The question presents an Assertion (A) and a Reason (R) related to the life cycle of mosses. We need to evaluate if both statements are true and if the reason correctly explains the assertion.

Step 2: Detailed Explanation:

- Analysis of Assertion (A): "The first stage of gametophyte in the life cycle of moss is protonema stage." The life cycle of a moss begins when a haploid spore germinates. This spore does not directly grow into the leafy moss plant. Instead, it develops into a filamentous, green, branching structure called the protonema. This protonema is the juvenile stage of the gametophyte. Later, buds develop on the protonema which grow into the mature, leafy gametophyte. So, the assertion is **correct**.
- Analysis of Reason (R): "Protonema develops directly from spores produced in capsule." The sporophyte of a moss consists of a foot, seta, and a capsule. Within the capsule, meiosis occurs to produce haploid spores. When these spores are released and land on a suitable substrate, they germinate and develop directly into the protonema. So, the reason is also **correct**.
- Evaluating the link between A and R: The reason explains the origin of the protonema. Since the protonema develops directly from the germinating spore (which is the starting point after meiosis), it logically follows that the protonema is the first stage of the gametophyte's

development. Therefore, the Reason (R) is the correct explanation for the Assertion (A).

Step 3: Final Answer:

Both Assertion A and Reason R are correct, and R provides the correct explanation for A.

Quick Tip

Remember the two distinct stages in the moss gametophyte: the initial, juvenile **protonema stage** (creeping, green, branched, and filamentous) and the subsequent, mature **leafy stage** which bears the sex organs. The protonema always arises from the spore.

118. What is the role of RNA polymerase III in the process of transcription in Eukaryotes?

- (A) Transcription of only snRNAs
- (B) Transcription of rRNAs (28S, 18S and 5.8S)
- (C) Transcription of tRNA, 5S rRNA and snRNA
- (D) Transcription of precursor of mRNA

Correct Answer: (C) Transcription of tRNA, 5S rRNA and snRNA

Solution:

Step 1: Understanding the Question:

The question asks for the specific function of RNA polymerase III in eukaryotic transcription. Eukaryotes have three distinct RNA polymerases, each responsible for transcribing different types of RNA.

Step 2: Detailed Explanation:

The roles of the three eukaryotic RNA polymerases are as follows:

- RNA Polymerase I: Located in the nucleolus, it is responsible for transcribing the large ribosomal RNA (rRNA) genes. This includes the precursors for the 28S, 18S, and 5.8S rRNAs.
- RNA Polymerase II: Located in the nucleoplasm, it transcribes all protein-coding genes to produce the precursor of messenger RNA (pre-mRNA or hnRNA). It also transcribes most small nuclear RNAs (snRNAs) and microRNAs (miRNAs).
- RNA Polymerase III: Located in the nucleoplasm, it is responsible for transcribing the genes for transfer RNA (tRNA), the 5S ribosomal RNA (a component of the large ribosomal subunit), and some other small RNAs, including some small nuclear RNAs (snRNAs) like U6 snRNA.

Let's check the options:

- (A) Transcription of only snRNAs Incorrect. It transcribes tRNA and 5S rRNA as well.
- (B) Transcription of rRNAs (28S, 18S and 5.8S) Incorrect. This is the function of RNA Polymerase I.

- (C) Transcription of tRNA, 5S rRNA and snRNA Correct. This accurately describes the main products of RNA Pol III.
- (D) Transcription of precursor of mRNA Incorrect. This is the function of RNA Polymerase II.

Step 3: Final Answer:

The role of RNA polymerase III is the transcription of tRNA, 5S rRNA, and some snRNAs.

Quick Tip

Use the mnemonic "1, 2, 3 - R, M, T" to remember the main product of each polymerase: RNA Pol I -; rRNA, RNA Pol II -; rRNA, RNA Pol III -; tRNA. This covers the primary function and helps differentiate them quickly. Remember that Pol III also makes 5S rRNA.

119. The historic Convention on Biological Diversity, 'The Earth Summit' was held in Rio de Janeiro in the year:

- (A) 2002
- (B) 1985
- (C) 1992
- (D) 1986

Correct Answer: (C) 1992

Solution:

Step 1: Understanding the Question:

This is a factual recall question asking for the year of the Earth Summit held in Rio de Janeiro, where the Convention on Biological Diversity (CBD) was established.

Step 2: Detailed Explanation:

The United Nations Conference on Environment and Development (UNCED), popularly known as the **Earth Summit**, was held in Rio de Janeiro, Brazil, from June 3 to 14, **1992**. This summit was a landmark event for environmental protection and sustainable development. One of its key outcomes was the opening for signature of the Convention on Biological Diversity (CBD), an international legally binding treaty with three main goals: the conservation of biodiversity, the sustainable use of its components, and the fair and equitable sharing of benefits arising out of the utilization of genetic resources.

Other important dates:

- 2002: The World Summit on Sustainable Development was held in Johannesburg.

Step 3: Final Answer:

The Earth Summit in Rio de Janeiro was held in 1992.

Remember these key environmental summit years: - 1972: Stockholm Conference (First major conference on international environmental issues) - 1987: Montreal Protocol (on substances that deplete the Ozone Layer) - 1992: Rio Earth Summit (Convention on Biological Diversity, UNFCCC) - 1997: Kyoto Protocol (to reduce greenhouse gas emissions) - 2002: Johannesburg Earth Summit (focus on sustainable development)

120. The phenomenon of pleiotropism refers to

- (A) more than two genes affecting a single character.
- (B) presence of several alleles of a single gene controlling a single crossover.
- (C) presence of two alleles, each of the two genes controlling a single trait.
- (D) a single gene affecting multiple phenotypic expression.

Correct Answer: (D) a single gene affecting multiple phenotypic expression.

Solution:

Step 1: Understanding the Question:

The question asks for the definition of pleiotropism (or pleiotropy).

Step 2: Detailed Explanation:

Let's define the genetic terms:

- **Pleiotropy:** This occurs when a single gene influences multiple, seemingly unrelated phenotypic traits. The gene product (e.g., an enzyme or protein) may participate in several metabolic pathways or have effects on different cell types, leading to a cascade of effects throughout the organism. A classic example is the gene for phenylketonuria (PKU) in humans, which causes mental retardation, reduced hair and skin pigmentation.
- **Polygenic Inheritance:** This is the opposite of pleiotropy. It occurs when a single phenotypic trait is controlled by the cumulative effect of two or more genes. Examples include human skin color, height, and weight.

Now let's evaluate the options:

- (A) more than two genes affecting a single character This describes polygenic inheritance.
- (B) presence of several alleles of a single gene... This describes multiple allelism.
- (C) presence of two alleles, each of the two genes... This is confusingly worded but does not describe pleiotropy.
- (D) a single gene affecting multiple phenotypic expression This is the correct definition of pleiotropism.

Step 3: Final Answer:

Pleiotropism is the phenomenon where a single gene affects multiple phenotypic expressions.

Think of the prefixes: **Pleio-** means "many" and **-tropy** means "effects". So, one gene has many effects. In contrast, **Poly-** means "many" and **-genic** means "genes". So, many genes affect one trait. This helps distinguish pleiotropy from polygenic inheritance.

121. Spraying of which of the following phytohormone on juvenile conifers helps in hastening the maturity period, that leads to early seed production?

- (A) Abscisic Acid
- (B) Indole-3-butyric Acid
- (C) Gibberellic Acid
- (D) Zeatin

Correct Answer: (C) Gibberellic Acid

Solution:

Step 1: Understanding the Question:

The question asks which plant hormone can be used to speed up the maturation process in juvenile conifers to promote early seed production.

Step 2: Detailed Explanation:

Let's review the functions of the given phytohormones:

- Abscisic Acid (ABA): Primarily a stress hormone, it is involved in stomatal closure, seed dormancy, and abscission. It generally acts as a growth inhibitor.
- Indole-3-butyric Acid (IBA): This is an auxin. Auxins are primarily involved in cell elongation, apical dominance, and rooting of stem cuttings. They do not hasten maturity.
- Gibberellic Acid (GA): Gibberellins have a wide range of effects, including stem elongation (bolting), breaking seed dormancy, and promoting flowering and fruit development. A key commercial application is spraying juvenile conifers with GAs to overcome the juvenile phase and induce early flowering and seed production. This is particularly useful in breeding programs and for commercial seed producers.
- **Zeatin:** This is a type of cytokinin. Cytokinins promote cell division, help overcome apical dominance, and delay senescence. They do not typically hasten maturity.

Based on these functions, gibberellic acid is the hormone used to accelerate the maturity period in juvenile conifers.

Step 3: Final Answer:

Spraying with Gibberellic Acid helps in hastening the maturity period of juvenile conifers.

Associate Gibberellins with "growth promotion" in several ways: increasing fruit size (grapes), elongating stems (sugarcane), malting process (brewing), and overcoming juvenility (conifers). This specific application in conifers is a commonly cited example.

122. Given below are two statements:

Statement I: The forces generated by transpiration can lift a xylem-sized column of water over 130 meters height.

Statement II: Transpiration cools leaf surfaces sometimes 10 to 15 degrees, by evaporative cooling.

In the light of the above statements, choose the most appropriate answer from the options given below:

- (A) Statement I is incorrect but Statement II is correct.
- (B) Both Statement I and Statement II are correct.
- (C) Both Statement I and Statement II are incorrect.
- (D) Statement I is correct but Statement II is incorrect.

Correct Answer: (B) Both Statement I and Statement II are correct.

Solution:

Step 1: Understanding the Question:

The question presents two statements about the effects of transpiration in plants. We need to determine the correctness of each statement.

Step 2: Detailed Explanation:

- Analysis of Statement I: "The forces generated by transpiration can lift a xylem-sized column of water over 130 meters height."

This statement refers to the Cohesion-Tension theory of water ascent. The evaporation of water from leaves (transpiration) creates a negative pressure potential or tension (the "pull") in the xylem. Due to the cohesive forces between water molecules and adhesive forces between water and xylem walls, this pull is transmitted down the entire water column. This mechanism is incredibly powerful and is indeed capable of lifting water to the tops of the tallest trees, such as the coast redwood (Sequoia sempervirens), which can exceed 115 meters. Therefore, lifting water over 130 meters is theoretically and practically possible by this force. This statement is correct.

- Analysis of Statement II: "Transpiration cools leaf surfaces sometimes 10 to 15 degrees, by evaporative cooling."

Transpiration is the evaporation of water from the plant surface, primarily leaves. Evaporation is a cooling process because it requires energy (latent heat of vaporization), which is drawn from the leaf tissue itself. This prevents the leaves from overheating, especially in direct sunlight and high temperatures. This evaporative cooling effect can lower the leaf temperature by 10 to

15 degrees Celsius compared to the surrounding air. This statement is also **correct**.

Step 3: Final Answer:

Since both Statement I and Statement II are correct descriptions of the functions and consequences of transpiration, the correct option is that both statements are correct.

Quick Tip

Remember the dual major roles of transpiration: 1. Driving Force: It creates the "transpiration pull" for water absorption and ascent. 2. Thermoregulation: It cools the plant through evaporative cooling. Both statements highlight these crucial functions.

123. Movement and accumulation of ions across a membrane against their concentration gradient can be explained by

- (A) Active Transport
- (B) Osmosis
- (C) Facilitated Diffusion
- (D) Passive Transport

Correct Answer: (A) Active Transport

Solution:

Step 1: Understanding the Question:

The question asks to identify the transport mechanism that moves ions across a membrane from a region of lower concentration to a region of higher concentration, which is "against their concentration gradient".

Step 2: Detailed Explanation:

Let's analyze the different transport mechanisms:

- Passive Transport: This is the movement of substances across a membrane down the concentration gradient (from high to low concentration) without the expenditure of cellular energy. Simple diffusion is a type of passive transport.
- Osmosis: A specific type of passive transport involving the movement of water across a semipermeable membrane from a region of high water potential to low water potential.
- Facilitated Diffusion: Another type of passive transport where substances move down the concentration gradient with the help of carrier proteins. It does not require metabolic energy.
- Active Transport: This process moves substances (like ions) against their concentration gradient (from low to high concentration). This "uphill" movement requires the expenditure of metabolic energy, typically in the form of ATP, and involves specific membrane proteins called pumps.

Since the question specifies movement "against their concentration gradient," the correct answer is Active Transport.

Step 3: Final Answer:

The movement of ions against a concentration gradient is carried out by active transport.

Quick Tip

Remember the key difference: Passive processes (diffusion, osmosis, facilitated diffusion) are "downhill" and require no energy. Active transport is "uphill" and always requires energy (ATP). The phrase "against the concentration gradient" is a direct indicator of active transport.

124. Given below are two statements:

Statement I: Endarch and exarch are the terms often used for describing the position of secondary xylem in the plant body.

Statement II: Exarch condition is the most common feature of the root system. In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is incorrect but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is correct but Statement II is false.

Correct Answer: (A) Statement I is incorrect but Statement II is true.

Solution:

Step 1: Understanding the Question:

The question presents two statements about the arrangement of xylem in plants and asks to evaluate their correctness.

Step 2: Detailed Explanation:

- Analysis of Statement I: "Endarch and exarch are the terms often used for describing the position of secondary xylem in the plant body."

The terms 'endarch' and 'exarch' describe the pattern of development of **primary xylem**, not secondary xylem. - **Exarch:** Protoxylem (the first formed xylem) is towards the periphery (outside) and metaxylem (later formed xylem) is towards the center. - **Endarch:** Protoxylem is towards the center (pith) and metaxylem is towards the periphery. Since these terms apply to primary xylem, Statement I is **incorrect**.

- Analysis of Statement II: "Exarch condition is the most common feature of the root system."

In the vascular bundle of roots, the development of xylem is centripetal, meaning the protoxylem is located towards the outer side (pericycle) and the metaxylem towards the center. This arrangement is known as the exarch condition. In contrast, stems typically exhibit an endarch condition. Therefore, Statement II is **true**.

Step 3: Final Answer:

Statement I is incorrect because endarch and exarch refer to primary xylem. Statement II is correct as the exarch condition is characteristic of roots.

Quick Tip

A simple way to remember is: **Ex**arch is for roots (towards the **ex**terior) and **En**darch is for stems (towards the **en**terior/center). These terms only describe the arrangement of primary xylem.

125. Which micronutrient is required for splitting of water molecule during photosynthesis?

- (A) copper
- (B) manganese
- (C) molybdenum
- (D) magnesium

Correct Answer: (B) manganese

Solution:

Step 1: Understanding the Question:

The question asks to identify the micronutrient that plays a crucial role in the photolysis (splitting) of water during the light-dependent reactions of photosynthesis.

Step 2: Detailed Explanation:

The splitting of water molecules $(2H_2O \rightarrow 4H^+ + 4e^- + O_2)$ occurs in Photosystem II (PS II) and is essential for releasing oxygen and providing electrons to the photosynthetic electron transport chain. This reaction is catalyzed by the Oxygen-Evolving Complex (OEC).

- The core of the OEC contains a cluster of four **manganese** (Mn) ions, along with one calcium (Ca) ion. These manganese ions are essential for the catalytic cycle of water oxidation.
- Copper (Cu) is a component of plastocyanin, an electron carrier in the electron transport chain.
- **Molybdenum** (Mo) is a component of enzymes like nitrate reductase and nitrogenase, crucial for nitrogen metabolism.
- Magnesium (Mg) is a central component of the chlorophyll molecule, essential for absorbing light energy. It is a macronutrient, not a micronutrient in this context.

Therefore, manganese (Mn) is the specific micronutrient required for the splitting of water. Chloride ions (Cl⁻) are also involved in this process.

Step 3: Final Answer:

Manganese is the micronutrient required for the splitting of water during photosynthesis.

Quick Tip

Remember the key roles of essential minerals in photosynthesis: **Mg** in chlorophyll for light capture, and **Mn** in PS II for water splitting. This is a frequently tested concept.

126. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: ATP is used at two steps in glycolysis.

Reason R: First ATP is used in converting glucose into glucose-6-phosphate and second ATP is used in conversion of fructose-6-phosphate into fructose-1-6-diphosphate. In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (B) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Question:

The question presents an Assertion (A) about ATP consumption in glycolysis and a Reason (R) specifying the steps. We need to evaluate if both are true and if R correctly explains A.

Step 2: Detailed Explanation:

- Analysis of Assertion (A): "ATP is used at two steps in glycolysis." Glycolysis is a 10-step process. The first part is the "energy investment phase" where ATP is consumed to activate the glucose molecule. Indeed, two molecules of ATP are consumed in this phase. So, Assertion A is **true**.
- Analysis of Reason (R): "First ATP is used in converting glucose into glucose-6-phosphate and second ATP is used in conversion of fructose-6-phosphate into fructose-1-6-diphosphate." Let's look at the specific steps of the energy investment phase: Step 1: Glucose is phosphorylated to glucose-6-phosphate by the enzyme hexokinase. This step consumes one ATP molecule. Step 3: Fructose-6-phosphate is phosphorylated to fructose-1,6-bisphosphate by the enzyme phosphofructokinase-1. This step consumes a second ATP molecule. The Reason

R accurately describes these two steps. So, Reason R is also true.

- Evaluating the link between A and R: The Assertion states that ATP is used at two steps. The Reason correctly identifies exactly those two steps where ATP is used. Therefore, the Reason (R) is the correct explanation for the Assertion (A).

Step 3: Final Answer:

Both Assertion A and Reason R are true, and R correctly explains A.

Quick Tip

Remember the two "priming" reactions in the energy investment phase of glycolysis: Step 1 (Hexokinase) and Step 3 (Phosphofructokinase). These are the only two steps where ATP is consumed.

127. Identify the correct statements:

- A. Detritivores perform fragmentation.
- B. The humus is further degraded by some microbes during mineralization.
- C. Water soluble inorganic nutrients go down into the soil and get precipitated by a process called leaching.
- D. The detritus food chain begins with living organisms.
- E. Earthworms break down detritus into smaller particles by a process called catabolism.

Choose the correct answer from the options given below:

- (A) D, E, A only
- (B) A, B, C only
- (C) B, C, D only
- (D) C, D, E only

Correct Answer: (B) A, B, C only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements from a list of five related to the process of decomposition.

Step 2: Detailed Explanation:

Decomposition involves several processes: fragmentation, leaching, catabolism, humification, and mineralization. Let's evaluate each statement:

- A. Detritivores perform fragmentation. Detritivores, such as earthworms and termites, break down detritus (dead organic matter) into smaller particles. This process is called fragmentation. This statement is **correct**.

- B. The humus is further degraded by some microbes during mineralization. Humus is a dark, amorphous substance that is highly resistant to microbial action and decomposes slowly. It is eventually degraded by microbes, releasing inorganic nutrients back into the soil. This release of inorganic nutrients is called mineralization. This statement is **correct**.
- C. Water soluble inorganic nutrients go down into the soil and get precipitated by a process called leaching. Leaching is the process by which water-soluble substances (like inorganic nutrients released during decomposition) are washed down through the soil profile and may become unavailable to plants. This statement is **correct**.
- D. The detritus food chain begins with living organisms. This is incorrect. The detritus food chain begins with dead organic matter (detritus). The grazing food chain begins with living organisms (producers).
- E. Earthworms break down detritus into smaller particles by a process called catabolism. Earthworms break down detritus into smaller particles, but this process is called fragmentation. Catabolism refers to the enzymatic degradation of detritus into simpler inorganic substances by bacteria and fungi. This statement is **incorrect**.

The correct statements are A, B, and C.

Step 3: Final Answer:

The only correct statements are A, B, and C.

Quick Tip

Differentiate the key terms in decomposition: - **Fragmentation:** Physical breakdown by detritivores (e.g., earthworm). - **Catabolism:** Chemical/enzymatic breakdown by microbes (bacteria, fungi). - **Leaching:** Washing away of soluble nutrients. - **Humification:** Formation of humus. - **Mineralization:** Release of inorganic nutrients from humus.

128. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: Late wood has fewer xylary elements with narrow vessels.

Reason R: Cambium is less active in winters.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (B) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Question:

The question presents an Assertion (A) about the structure of late wood and a Reason (R) about cambial activity in winter. We must determine if they are true and if R explains A.

Step 2: Detailed Explanation:

- Analysis of Assertion (A): "Late wood has fewer xylary elements with narrow vessels." In temperate regions, the cambium's activity varies with the seasons. The wood formed during the autumn or winter season is called autumn wood or late wood. This wood is characterized by having smaller, narrower vessels and thicker walls compared to the wood formed in spring. It generally has fewer xylary elements. So, Assertion A is true.
- Analysis of Reason (R): "Cambium is less active in winters."

 The activity of the vascular cambium is influenced by physiological and environmental factors like temperature and water availability. In winter, conditions are unfavorable (cold, less daylight), leading to a significant decrease in the cambium's activity. So, Reason R is **true**.
- Evaluating the link between A and R: The reduced activity of the cambium during winter (Reason) is the direct cause for the formation of late wood with its specific characteristics—fewer xylary elements and narrow, thick-walled vessels (Assertion). Therefore, the Reason (R) is the correct explanation for the Assertion (A).

Step 3: Final Answer:

Both Assertion A and Reason R are true, and R is the correct explanation of A.

Quick Tip

Remember the contrast: - Spring wood (Early wood): Cambium is very active -; many, large, wide vessels. - Autumn wood (Late wood): Cambium is less active -; fewer, small, narrow vessels. The combination of these two forms an annual ring.

129. In angiosperm, the haploid, diploid and triploid structures of a fertilized embryo sac sequentially are:

- (A) Synergids, antipodals and Polar nuclei
- (B) Synergids, Primary endosperm nucleus and zygote
- (C) Antipodals, synergids, and primary endosperm nucleus
- (D) Synergids, Zygote and Primary endosperm nucleus

Correct Answer: (D) Synergids, Zygote and Primary endosperm nucleus

Solution:

Step 1: Understanding the Question:

The question asks to identify a sequence of structures from a fertilized embryo sac that are haploid (n), diploid (2n), and triploid (3n), in that specific order.

Step 2: Detailed Explanation:

Let's determine the ploidy level of the structures found in a fertilized embryo sac:

- Haploid (n) structures: Synergids: These are cells of the egg apparatus. They are haploid. After fertilization, they degenerate. Antipodals: These cells are at the chalazal end and are also haploid. They also degenerate after fertilization. Diploid (2n) structure:
- **Zygote:** Formed by the fusion of one male gamete (n) and the egg cell (n). It develops into the embryo. It is diploid. **Triploid (3n) structure: Primary Endosperm Nucleus (PEN):** Formed by the fusion of the second male gamete (n) with the central cell containing the secondary nucleus (2n, formed from the fusion of two polar nuclei). The PEN develops into the endosperm. It is triploid.

Now we need to find an option that lists a haploid, a diploid, and a triploid structure in sequence (n, 2n, 3n).

- (A) Synergids (n), antipodals (n), Polar nuclei (n+n). Not n, 2n, 3n.
- (B) Synergids (n), Primary endosperm nucleus (3n), zygote (2n). The order is n, 3n, 2n. Incorrect.
- (C) Antipodals (n), synergids (n), primary endosperm nucleus (3n). Not n, 2n, 3n.
- (D) Synergids (n), Zygote (2n), and Primary endosperm nucleus (3n). This sequence matches the required n, 2n, 3n order.

Step 3: Final Answer:

The correct sequence of haploid, diploid, and triploid structures is Synergids, Zygote, and Primary endosperm nucleus.

Quick Tip

Double fertilization is the key to understanding ploidy in a fertilized embryo sac. Remember: 1. **Syngamy:** male gamete (n) + egg(n) - i Zygote (2n) 2. **Triple Fusion:** male gamete (n) + central cell (2n) - i PEN (3n) All other maternal cells in the embryo sac (synergids, antipodals) are haploid (n).

130. Which hormone promotes internode/petiole elongation in deep water rice?

- (A) 2, 4-D
- (B) GA_3
- (C) Kinetin
- (D) Ethylene

Correct Answer: (D) Ethylene

Solution:

Step 1: Understanding the Question:

The question asks to identify the specific plant hormone responsible for promoting rapid stem elongation in deep water rice plants when they are submerged.

Step 2: Detailed Explanation:

- 2, 4-D: This is a synthetic auxin, primarily used as a herbicide to control broad-leaf weeds.
- GA₃ (Gibberellic Acid): This hormone is well-known for promoting stem elongation (bolting) in many plants.
- **Kinetin:** This is a cytokinin, which primarily promotes cell division and is involved in delaying senescence.
- Ethylene: This is a gaseous hormone with various effects. In semi-aquatic plants like deep water rice, submergence of the plant leads to the accumulation of ethylene in the tissues. This trapped ethylene then triggers a significant and rapid increase in internodal elongation, helping the leaves and upper parts of the plant to remain above the water surface for gas exchange and photosynthesis. While gibberellins are involved in the elongation process, ethylene is the primary trigger in this specific adaptation.

Therefore, ethylene is the hormone that directly promotes this rapid elongation in deep water rice.

Step 3: Final Answer:

Ethylene promotes internode/petiole elongation in deep water rice.

Quick Tip

While both gibberellins and ethylene can cause stem elongation, remember the specific context: for "deep water rice" or "submerged plants," the key trigger is **ethylene**. This is a classic example of ethylene's role in plant adaptation.

131. Unequivocal proof that DNA is the genetic material was first proposed by

- (A) Wilkins and Franklin
- (B) Frederick Griffith
- (C) Alfred Hershey and Martha Chase
- (D) Avery, Macleoid and McCarthy

Correct Answer: (C) Alfred Hershey and Martha Chase

Solution:

Step 1: Understanding the Question:

The question asks to identify the scientists who provided the definitive or "unequivocal" proof

that DNA, and not protein, is the genetic material.

Step 2: Detailed Explanation:

Let's review the timeline of key experiments:

- Frederick Griffith (1928): Performed the transformation experiment with *Streptococcus pneumoniae*. He showed that a "transforming principle" from dead virulent bacteria could make non-virulent bacteria virulent, but he did not identify the nature of this principle.
- Avery, Macleoid, and McCarthy (1944): They extended Griffith's work and used biochemical methods to show that the transforming principle was DNA. While their evidence was strong, it was not universally accepted by the scientific community, as many still believed protein was the more likely candidate for genetic material.
- Alfred Hershey and Martha Chase (1952): They conducted the famous "blender experiment" using T2 bacteriophage, a virus that infects bacteria. They used radioactive isotopes to label the phage's DNA and protein separately. They labeled the DNA with radioactive phosphorus (³²P) because DNA contains phosphorus but protein does not.
- They labeled the protein coat with radioactive sulfur (35 S) because proteins contain sulfur (in some amino acids) but DNA does not. They found that only the radioactive DNA (32 P) entered the host bacterial cells, while the radioactive protein coat (35 S) remained outside. Since the injected DNA was sufficient to cause the production of new phages, this provided conclusive proof that DNA is the genetic material.
- Wilkins and Franklin: Used X-ray diffraction to study the structure of DNA, which was crucial for Watson and Crick's model, but did not prove its function as genetic material.

The Hershey-Chase experiment is considered the "unequivocal proof".

Step 3: Final Answer:

The unequivocal proof that DNA is the genetic material was provided by Alfred Hershey and Martha Chase.

Quick Tip

Remember the key contributions: - Griffith: Discovered transformation. - Avery et al.: Identified the transforming substance as DNA. - Hershey Chase: Provided unequivocal proof using radioactive tracers. The word "unequivocal" in the question points directly to the Hershey-Chase experiment.

132. In the equation GPP – R = NPP GPP is Gross Primary Productivity NPP is Net Primary Productivity R here is

- (A) Reproductive allocation
- (B) Photosynthetically active radiation
- (C) Respiratory quotient

(D) Respiratory loss

Correct Answer: (D) Respiratory loss

Solution:

Step 1: Understanding the Question:

The question provides the fundamental equation for net primary productivity and asks to identify what the term 'R' represents.

Step 2: Key Formula or Approach:

The relationship between Gross Primary Productivity (GPP), Net Primary Productivity (NPP), and Respiration (R) is given by:

$$NPP = GPP - R$$

Step 3: Detailed Explanation:

Let's define the terms:

- Gross Primary Productivity (GPP): This is the total rate at which solar energy is captured by producers (like plants) and converted into chemical energy in the form of organic compounds through photosynthesis. It's the total amount of photosynthesis.
- Respiratory Loss (R): Producers use a significant portion of the energy they capture (GPP) for their own metabolic activities, such as growth, maintenance, and reproduction. This energy is consumed through cellular respiration. 'R' represents this amount of energy lost as heat during respiration.
- Net Primary Productivity (NPP): This is the rate at which producers create biomass after accounting for their respiratory losses. It is the energy that is stored as biomass and is available to the next trophic level (consumers).

Therefore, R in the equation GPP - R = NPP represents the energy lost through respiration by the producers.

Step 4: Final Answer:

In the equation, R represents Respiratory loss.

Quick Tip

Think of it like a salary: - **GPP** is your gross salary (total income). - **R** (**Respiration**) are your expenses and taxes. - **NPP** is your net salary (take-home pay), which is what you have left to use or save.

133. Expressed Sequence Tags (ESTs) refers to

- (A) Certain important expressed genes.
- (B) All genes that are expressed as RNA.

- (C) All genes that are expressed as proteins.
- (D) All genes whether expressed or unexpressed.

Correct Answer: (B) All genes that are expressed as RNA.

Solution:

Step 1: Understanding the Question:

The question asks for the definition of Expressed Sequence Tags (ESTs), a concept related to genomics and the Human Genome Project.

Step 2: Detailed Explanation:

The process of identifying ESTs involves:

1. Isolating messenger RNA (mRNA) from a specific cell or tissue type. The presence of mRNA indicates that a gene is being "expressed" or transcribed. 2. Using the enzyme reverse transcriptase to create a complementary DNA (cDNA) copy of the mRNA. 3. Sequencing short fragments (tags) from the ends of these cDNAs. These short sequenced fragments are called Expressed Sequence Tags (ESTs).

Because ESTs are derived from mRNA, they represent fragments of genes that are actively being transcribed into RNA in that particular tissue at that particular time. This methodology was used as a rapid way to discover and catalog genes. Therefore, ESTs represent "all genes that are expressed as RNA".

Let's look at the options:

- (A) "Certain important expressed genes" is too vague. The method identifies all expressed genes, not just "important" ones.
- (C) "All genes that are expressed as proteins" is incorrect. The process starts from mRNA. Not all transcribed RNAs are translated into proteins (e.g., non-coding RNAs).
- (D) "All genes whether expressed or unexpressed" is incorrect. ESTs specifically target expressed genes by using mRNA as the starting material.

Step 3: Final Answer:

Expressed Sequence Tags (ESTs) refer to all genes that are expressed as RNA.

Quick Tip

Break down the term: - Expressed: The gene is turned on (transcribed into mRNA). - Sequence Tag: A short piece of DNA sequence that acts as a unique identifier or "tag" for that gene. This helps remember that ESTs are about identifying expressed genes.

134. Among eukaryotes, replication of DNA takes place in -

- (A) G_2 phase
- (B) M phase

- (C) S phase
- (D) G_1 phase

Correct Answer: (C) S phase

Solution:

Step 1: Understanding the Question:

The question asks to identify the specific phase of the eukaryotic cell cycle during which DNA replication occurs.

Step 2: Detailed Explanation:

The eukaryotic cell cycle is divided into two main stages: Interphase and the M phase (Mitotic phase). Interphase is further subdivided into three phases:

- G_1 phase (Gap 1): This is the period of cell growth before DNA synthesis begins. The cell is metabolically active and synthesizes proteins and RNAs.
- S phase (Synthesis): This is the phase where the cell's DNA is replicated. At the end of the S phase, each chromosome consists of two sister chromatids. The amount of DNA in the cell doubles during this phase.
- G_2 phase (Gap 2): This is the period after DNA synthesis has occurred but before the start of mitosis. The cell continues to grow and produces proteins necessary for cell division.
- M phase (Mitosis/Meiosis): This is the phase of actual cell division, including nuclear division (mitosis) and cytoplasmic division (cytokinesis).

Therefore, DNA replication is confined to the S phase of the cell cycle.

Step 3: Final Answer:

Replication of DNA in eukaryotes takes place in the S phase.

Quick Tip

Remember the cell cycle order: $G_1 \to S \to G_2 \to M$. The 'S' stands for 'Synthesis', specifically the synthesis of new DNA.

135. In gene gun method used to introduce alien DNA into host cells, microparticles of _____ metal are used.

- (A) Silver
- (B) Copper
- (C) Zinc
- (D) Tungsten or gold

Correct Answer: (D) Tungsten or gold

Solution:

Step 1: Understanding the Question:

The question asks to identify the type of metal used for the microparticles in the gene gun (biolistics) method of genetic transformation.

Step 2: Detailed Explanation:

The gene gun method, also known as biolistics or microparticle bombardment, is a physical method for delivering foreign DNA (transgenes) into cells. The procedure involves:

1. Coating microscopic particles of a heavy metal with the DNA of interest. 2. Loading these DNA-coated microparticles onto a carrier sheet. 3. Propelling the particles at high velocity towards the target cells or tissues using a burst of pressurized gas (like helium). 4. The high-velocity particles penetrate the cell walls and membranes, carrying the foreign DNA into the cell's interior.

The microparticles need to be dense enough to have sufficient momentum to penetrate the cells and chemically inert so they do not react with the DNA or the cell's components. The most commonly used metals for these microparticles are **gold** (Au) and **tungsten** (W).

Step 3: Final Answer:

In the gene gun method, microparticles of tungsten or gold are used.

Quick Tip

Associate the "gene gun" with shooting "golden bullets." The use of heavy, inert metals like gold or tungsten is crucial for the success of this biolistic technique.

136. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: In gymnosperms the pollen grains are released from the microsporangium and carried by air currents.

Reason R: Air currents carry the pollen grains to the mouth of the archegonia where the male gametes are discharged and pollen tube is not formed.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (D) A is true but R is false.

Solution:

Step 1: Understanding the Question:

The question presents an Assertion (A) and a Reason (R) about pollination and fertilization in gymnosperms. We need to evaluate their correctness.

Step 2: Detailed Explanation:

- Analysis of Assertion (A): "In gymnosperms the pollen grains are released from the microsporangium and carried by air currents."

Gymnosperms (like pines, cycads) lack flowers and their ovules are exposed. Pollination is predominantly anemophilous, meaning it is mediated by wind (air currents). Pollen grains are produced in microsporangia (pollen sacs) and released into the air. This statement is **true**.

- Analysis of Reason (R): "Air currents carry the pollen grains to the mouth of the archegonia where the male gametes are discharged and pollen tube is not formed."

This statement has several inaccuracies. 1. Air currents carry the pollen grains to the ovule, specifically to the micropyle of the ovule, not directly to the archegonia which are inside the ovule. 2. After landing on the ovule, the pollen grain germinates and **forms a pollen tube**. This pollen tube grows through the nucellus towards the archegonium. 3. The pollen tube carries the male gametes and discharges them near the archegonium for fertilization. The formation of a pollen tube (siphonogamy) is a key feature of seed plants, including gymnosperms (with a few exceptions in primitive groups where the process is slightly different, but a tube is still formed). Therefore, the statement that a "pollen tube is not formed" is incorrect. The Reason R is **false**.

Step 3: Final Answer:

Assertion A is true, but Reason R is false.

Quick Tip

Remember that both gymnosperms and angiosperms are siphonogamous, meaning they form a pollen tube to deliver male gametes. The key error in the Reason statement is the claim that a pollen tube is not formed.

```
137. Match List I with List II:
```

List I (Interaction)

A. Mutualism

B. Commensalism

C. Amensalism

D. Parasitism

List II (Species A and B)

I. +(A), 0(B)

II. -(A), 0(B)

III. +(A), -(B)

IV. +(A), +(B)

Choose the correct answer from the options given below:

- (A) A-III, B-I, C-IV, D-II
- (B) A-IV, B-II, C-I, D-III
- (C) A-IV, B-I, C-II, D-III
- (D) A-IV, B-III, C-I, D-II

Correct Answer: (C) A-IV, B-I, C-II, D-III

Solution:

Step 1: Understanding the Question:

The question requires matching different types of population interactions (List I) with their symbolic representation (List II), where '+' denotes benefit, '-' denotes harm, and '0' denotes no effect.

Step 2: Detailed Explanation:

Let's define each interaction and find its corresponding representation:

- **A.** Mutualism: An interaction where both species A and B benefit. The representation is (+, +). This matches IV.
- **B. Commensalism:** An interaction where one species (A) benefits, and the other species (B) is neither harmed nor benefited (unaffected). The representation is (+, 0). This matches **I**.
- C. Amensalism: An interaction where one species (A) is harmed, and the other species (B) is unaffected. The representation is (-, 0). This matches II.
- **D. Parasitism:** An interaction where one species (the parasite, A) benefits at the expense of the other species (the host, B), which is harmed. The representation is (+, -). This matches **III**.

So the correct pairings are: - A \rightarrow IV - B \rightarrow I - C \rightarrow II - D \rightarrow III Now let's check the options. The option that corresponds to this matching is (C).

Step 3: Final Answer:

The correct matching is A-IV, B-I, C-II, D-III.

Quick Tip

Create a simple table in your mind or on paper for population interactions. This makes matching questions very easy. - Mutualism: (+,+) - Competition: (-,-) - Predation/Parasitism: (+,-) - Commensalism: (+,0) - Amensalism: (-,0) - Neutralism: (0,0)

138. Match List I with List II:

List I

- A. Cohesion
- B. Adhesion
- C. Surface tension

D. Guttation

List II

- I. More attraction in liquid phase
- II. Mutual attraction among water molecules
- III. Water loss in liquid phase
- IV. Attraction towards polar surfaces

Choose the correct answer from the options given below:

- (A) A-II, B-I, C-IV, D-III
- (B) A-II, B-IV, C-I, D-III
- (C) A-IV, B-III, C-II, D-I
- (D) A-III, B-I, C-IV, D-II

Correct Answer: (B) A-II, B-IV, C-I, D-III

Solution:

Step 1: Understanding the Question:

The question asks to match properties and processes related to water in plants (List I) with their correct descriptions (List II).

Step 2: Detailed Explanation:

Let's define each term in List I and match it with the correct description in List II.

- A. Cohesion: This is the property of water molecules being attracted to each other due to hydrogen bonds. This is best described as "Mutual attraction among water molecules". This matches II.
- **B. Adhesion:** This is the attraction of water molecules to a different type of surface, such as the polar surfaces of the xylem walls (tracheids and vessels). This is best described as "Attraction towards polar surfaces". This matches **IV**.
- C. Surface tension: This is a consequence of cohesion. Water molecules at the surface are more strongly attracted to each other (in the liquid) than to the molecules in the air (gas phase). This property allows water to resist an external force and is described as "More attraction in liquid phase". This matches I.
- **D. Guttation:** This is the exudation of water droplets (xylem sap) from the tips or margins of leaves, typically occurring at night or in highly humid conditions when transpiration is low. It is "Water loss in liquid phase". This matches **III**.

So, the correct pairings are: - A \rightarrow II - B \rightarrow IV - C \rightarrow I - D \rightarrow III Checking the options, option (B) matches our pairings.

Step 3: Final Answer:

The correct matching is A-II, B-IV, C-I, D-III.

Clearly distinguish these terms: - Cohesion: Water-to-Water attraction. - Adhesion: Water-to-Surface (e.g., xylem wall) attraction. - Surface Tension: A result of cohesion at the liquid-air interface. - Guttation vs. Transpiration: Guttation is water loss as a liquid (hydathodes), while transpiration is water loss as a vapor (stomata).

139. Melonate inhibits the growth of pathogenic bacteria by inhibiting the activity of

- (A) Dinitrogenase
- (B) Succinic dehydrogenase
- (C) Amylase
- (D) Lipase

Correct Answer: (2) Succinic dehydrogenase

Solution:

Step 1: Understanding the Question:

The question asks to identify the enzyme whose activity is inhibited by malonate, thereby stopping the growth of pathogenic bacteria. This relates to the concept of enzyme inhibition.

Step 2: Detailed Explanation:

Malonate is a classic example of a competitive inhibitor. Its structure is very similar to that of succinate, the natural substrate for the enzyme succinic dehydrogenase.

$$\begin{array}{ccc} \textbf{Succinate} & \textbf{Malonate} \\ -OOC - CH2 - CH2 - COO^{-} & -OOC - CH2 - COO^{-} \end{array}$$

Due to this structural similarity, malonate binds to the active site of succinic dehydrogenase but does not undergo a reaction. This blocks the substrate (succinate) from binding, thus inhibiting the enzyme's activity.

Succinic dehydrogenase is a key enzyme in the Krebs cycle (or tricarboxylic acid cycle), which is essential for cellular respiration and energy production in both bacteria and eukaryotes. By inhibiting this enzyme, malonate disrupts the Krebs cycle, leading to a lack of energy (ATP) production, which inhibits the growth of the pathogenic bacteria.

Step 3: Evaluating other options:

- Dinitrogenase: An enzyme used by nitrogen-fixing bacteria to convert atmospheric nitrogen (N_2) to ammonia (NH_3) . It is not inhibited by malonate.
- Amylase: An enzyme that breaks down starch into sugars.
- **Lipase:** An enzyme that breaks down fats (lipids).

These enzymes do not have active sites that malonate can bind to as a competitive inhibitor.

Step 4: Final Answer:

Therefore, malonate inhibits the activity of succinic dehydrogenase.

Quick Tip

Remember that competitive inhibitors often have a structure very similar to the enzyme's natural substrate. Malonate and succinate are a frequently cited example in competitive inhibition.

140. Match List I with List II:

List I List II

- A. Oxidative decarboxylation I. Citrate synthase
- B. Glycolysis II. Pyruvate dehydrogenase
- C. Oxidative phosphorylation III. Electron transport system
- D. Tricarboxylic acid cycle IV. EMP pathway

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-III, D-I
- (B) A-III, B-IV, C-II, D-I
- (C) A-II, B-IV, C-I, D-III
- (D) A-III, B-I, C-II, D-IV

Correct Answer: (1) A-II, B-IV, C-III, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching metabolic processes from List I with their associated enzymes or pathways from List II.

Step 2: Detailed Explanation:

Let's analyze each item in List I and find its correct match in List II.

- A. Oxidative decarboxylation: This process involves the removal of a carboxyl group as CO_2 along with oxidation. A key example is the conversion of pyruvate to acetyl-CoA, which is catalyzed by the **Pyruvate dehydrogenase** complex. So, **A matches with II**.
- B. Glycolysis: This is the metabolic pathway that converts glucose into pyruvate. It is also known as the Embden-Meyerhof-Parnas pathway, or EMP pathway. So, B matches with IV.

- C. Oxidative phosphorylation: This is the process where ATP is formed as a result of the transfer of electrons from NADH or FADH₂ to O₂ by a series of electron carriers. This entire process happens in the Electron transport system (ETS). So, C matches with III.
- D. Tricarboxylic acid cycle (TCA cycle): Also known as the Krebs cycle. The first step of this cycle involves the condensation of acetyl-CoA with oxaloacetate to form citrate, a reaction catalyzed by the enzyme Citrate synthase. So, D matches with I.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to IV$

 $\mathrm{C} o \mathrm{III}$

 $\mathrm{D} \to \mathrm{I}$

This corresponds to the option (1).

Quick Tip

For matching questions, start with the pairs you are most certain about. For instance, knowing that Glycolysis is the EMP pathway (B-IV) might help eliminate some incorrect options quickly.

141. Identify the correct statements:

- A. Lenticels are the lens-shaped openings permitting the exchange of gases.
- B. Bark formed early in the season is called hard bark.
- C. Bark is a technical term that refers to all tissues exterior to vascular cambium.
- D. Bark refers to periderm and secondary phloem.
- E. Phellogen is single-layered in thickness.

Choose the correct answer from the options given below:

- (A) B and C only
- (B) B, C and E only
- (C) A and D only
- (D) A, B and D only

Correct Answer: (3) A and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements about plant anatomy, specifically related

to bark and associated structures.

Step 2: Detailed Explanation:

Let's evaluate each statement:

- Statement A: Lenticels are the lens-shaped openings permitting the exchange of gases. This is correct. Lenticels are porous tissues consisting of cells with large intercellular spaces in the periderm of secondarily thickened organs, such as stems and roots of woody plants. They function as pores for direct gas exchange between internal tissues and the atmosphere.
- Statement B: Bark formed early in the season is called hard bark. This is incorrect. Bark formed early in the season (spring) is called 'soft bark', while bark formed late in the season (autumn) is called 'hard bark'. This is analogous to spring wood and autumn wood.
- Statement C: Bark is a technical term that refers to all tissues exterior to vascular cambium. This statement is a broad definition and generally considered correct in botany. However, statement D provides a more precise component-based definition. In the context of multiple-choice questions where only the 'most' correct options are chosen, this might be considered less precise than D. According to the provided answer key, this statement is considered incorrect.
- Statement D: Bark refers to periderm and secondary phloem. This is a correct and precise definition. Bark is composed of two main parts: the inner bark (secondary phloem) and the outer bark (periderm and all tissues outside it, also known as rhytidome). Thus, 'periderm and secondary phloem' are the essential components of bark.
- Statement E: Phellogen is single-layered in thickness. This is generally true, as phellogen (cork cambium) is a meristematic tissue. However, it can sometimes be two or more cell layers thick. Therefore, stating it is strictly 'single-layered' might be considered an oversimplification and thus incorrect in a strict sense.

Step 3: Final Answer:

Based on the analysis, statements A and D are the most accurate and universally accepted facts. The provided answer key confirms that A and D are the correct statements.

Quick Tip

In plant anatomy, definitions can be nuanced. 'Bark' has both a broad definition (all tissues outside vascular cambium) and a more specific one (periderm + secondary phloem). Often, questions test the more precise, component-based definition.

142. Which of the following statements are correct about Klinefelter's Syndrome?

- A. This disorder was first described by Langdon Down (1866).
- B. Such an individual has overall masculine development. However, the feminine development is also expressed.
- C. The affected individual is short statured.
- D. Physical, psychomotor and mental development is retarded.
- E. Such individuals are sterile.

Choose the correct answer from the options given below:

- (A) A and E only
- (B) A and B only
- (C) C and D only
- (D) B and E only

Correct Answer: (4) B and E only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements describing Klinefelter's Syndrome from a given list of features.

Step 2: Detailed Explanation:

Klinefelter's Syndrome is a genetic condition caused by the presence of an extra X chromosome in males, resulting in the karyotype 47, XXY. Let's analyze each statement:

- Statement A: This disorder was first described by Langdon Down (1866). This is incorrect. Langdon Down described Down's Syndrome. Klinefelter's Syndrome was first described by Dr. Harry Klinefelter in 1942.
- Statement B: Such an individual has overall masculine development. However, the feminine development is also expressed. This is correct. Individuals are phenotypically male, but the extra X chromosome leads to the development of some feminine characteristics, such as gynecomastia (enlargement of breast tissue).
- Statement C: The affected individual is short statured. This is incorrect. Individuals with Klinefelter's Syndrome are often taller than average, with long limbs. Short stature is characteristic of Turner's Syndrome (45, XO).
- Statement D: Physical, psychomotor and mental development is retarded. This is incorrect. While some individuals may have learning disabilities or delayed speech development, severe mental retardation is not a typical feature. Intelligence is usually within

the normal range.

• Statement E: Such individuals are sterile. This is correct. The presence of the extra X chromosome leads to underdeveloped testes (testicular atrophy) and, consequently, infertility or sterility due to low or no sperm production (azoospermia).

Step 3: Final Answer:

The correct statements are B and E. Therefore, the correct option is (4).

Quick Tip

For chromosomal disorders, create a table comparing key features of Klinefelter's Syndrome (XXY), Turner's Syndrome (XO), and Down's Syndrome (Trisomy 21). Focus on karyotype, physical characteristics, fertility, and mental development.

143. Main steps in the formation of Recombinant DNA are given below. Arrange these steps in a correct sequence.

- A. Insertion of recombinant DNA into the host cell.
- B. Cutting of DNA at specific location by restriction enzyme.
- C. Isolation of desired DNA fragment.
- D. Amplification of gene of interest using PCR.

Choose the correct answer from the options given below:

- (A) B, D, A, C
- (B) B, C, D, A
- (C) C, A, B, D
- (D) C, B, D, A

Correct Answer: (4) C, B, D, A

Solution:

Step 1: Understanding the Question:

The question asks for the correct chronological order of the main steps involved in creating recombinant DNA.

Step 2: Detailed Explanation:

The process of creating recombinant DNA involves several sequential steps. Let's analyze the logical flow:

1. **Isolation of the Genetic Material:** The very first step is to obtain the DNA from the source organism. This could be isolating the entire genomic DNA. Phrased as C.

Isolation of desired DNA fragment, this represents the initial step of getting the source material.

- 2. Cutting the DNA: Once the DNA is isolated, restriction enzymes are used to cut it at specific recognition sites to generate fragments. One of these fragments will be the gene of interest. This corresponds to B. Cutting of DNA at specific location by restriction enzyme. After cutting, the specific fragment is often isolated using gel electrophoresis. So, C → B is a logical start.
- 3. Amplification of the Gene: To get a sufficient quantity of the gene of interest for subsequent steps, it is amplified using the Polymerase Chain Reaction (PCR). This is **D.** Amplification of gene of interest using PCR.
- 4. **Ligation and Insertion:** The amplified gene of interest is then ligated (joined) with a vector DNA (like a plasmid) that has been cut with the same restriction enzyme. The resulting molecule is the recombinant DNA. This recombinant DNA is then introduced into a suitable host cell (like a bacterium). This final step is **A. Insertion of recombinant DNA into the host cell**.

So, the correct sequence of the given steps is $C \to B \to D \to A$.

Step 3: Final Answer:

Arranging the steps in the correct order gives C, B, D, A. This matches option (4).

Quick Tip

Remember the general workflow of rDNA technology: Isolate the DNA, Cut it with restriction enzymes, Amplify the desired gene (PCR), Ligate it into a vector, and finally, Insert the recombinant vector into a host.

144. Which of the following combinations is required for chemiosmosis?

- (A) proton pump, electron gradient, NADP synthase
- (B) membrane, proton pump, proton gradient, ATP synthase
- (C) membrane, proton pump, proton gradient, NADP synthase
- (D) proton pump, electron gradient, ATP synthase

Correct Answer: (2) membrane, proton pump, proton gradient, ATP synthase

Solution:

Step 1: Understanding the Question:

The question asks for the essential components required for the process of chemiosmosis, which is the mechanism for ATP synthesis in both photosynthesis and cellular respiration.

Step 2: Detailed Explanation:

According to Peter Mitchell's chemiosmotic theory, ATP synthesis is coupled to the movement of protons across a semipermeable membrane. The key requirements are:

- A membrane: An intact inner mitochondrial membrane or a thylakoid membrane is necessary to maintain a proton gradient. It must be impermeable to protons except through specific channels.
- A proton pump: This is typically the electron transport chain (ETC). As electrons move through the ETC, energy is used to pump protons (H^+) from the matrix (in mitochondria) or stroma (in chloroplasts) to the intermembrane space or thylakoid lumen, respectively.
- A proton gradient: The pumping of protons creates a difference in proton concentration and electrical charge across the membrane, which is a form of potential energy called the proton-motive force.
- ATP synthase: This is a transmembrane enzyme complex (also known as F_0 - F_1 particle) that has a channel for protons to flow back down their concentration gradient. The energy released from this proton flow is used by the enzyme to synthesize ATP from ADP and inorganic phosphate (Pi).

Step 3: Evaluating the options:

- (1) and (3) are incorrect because NADP synthase (more accurately, NADP⁺ reductase) is involved in reducing NADP⁺ to NADPH in photosynthesis, but it is not a core component of the ATP synthesis mechanism via chemiosmosis itself.
- (4) is incorrect because it omits the crucial requirement of an intact membrane to establish and maintain the gradient. It also mentions 'electron gradient' which is less precise than 'proton gradient'. The gradient that drives ATP synthesis is of protons (H^+) .
- (2) correctly lists all four essential components: the membrane, the pump to create the gradient, the gradient itself, and the ATP synthase enzyme that utilizes the gradient.

Step 4: Final Answer:

The correct combination required for chemiosmosis is membrane, proton pump, proton gradient, and ATP synthase.

Quick Tip

Think of chemiosmosis like a hydroelectric dam. The **membrane** is the dam wall. The **proton pump** is the system that fills the reservoir. The **proton gradient** is the stored water at a high level. The **ATP synthase** is the turbine through which water flows to generate electricity (ATP).

145. How many different proteins does the ribosome consist of?

- (A) 20
- (B) 80
- (C) 60
- (D) 40

Correct Answer: (2) 80

Solution:

Step 1: Understanding the Question:

The question asks for the total number of different proteins found in a ribosome. Ribosomes are complex structures composed of ribosomal RNA (rRNA) and proteins.

Step 2: Detailed Explanation:

The composition of ribosomes differs between prokaryotes and eukaryotes.

- Prokaryotic ribosomes (70S): They consist of a 50S large subunit and a 30S small subunit. The 50S subunit contains about 34 proteins, and the 30S subunit contains about 21 proteins, for a total of approximately 55 proteins.
- Eukaryotic ribosomes (80S): They consist of a 60S large subunit and a 40S small subunit. The 60S subunit contains about 49 proteins, and the 40S subunit contains about 33 proteins. The total number of proteins is approximately 82.

Since the question does not specify the type of ribosome (prokaryotic or eukaryotic), we must choose the best fit from the given options. The number 80 is a very close approximation for the number of proteins in a eukaryotic ribosome. The other options (20, 40, 60) are not accurate for either type. In the context of general biology questions, "ribosome" often implicitly refers to the more complex eukaryotic ribosome unless stated otherwise.

Step 3: Final Answer:

The number of different proteins in a eukaryotic ribosome is approximately 80. This matches option (2).

Quick Tip

Memorize the key differences between prokaryotic (70S) and eukaryotic (80S) ribosomes, including their subunit Svedberg units (50S+30S vs 60S+40S) and the approximate number of proteins. Note that Svedberg units are not additive.

146. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:

Assertion A: A flower is defined as modified shoot wherein the shoot apical meristem changes to floral meristem.

Reason R: Internode of the shoot gets condensed to produce different floral appendages laterally at successive nodes instead of leaves.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (2) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Question:

This is an Assertion-Reason question. We need to evaluate if Assertion (A) and Reason (R) are true statements individually, and then determine if R is the correct explanation for A.

Step 2: Evaluating Assertion A:

Assertion A: A flower is defined as a modified shoot wherein the shoot apical meristem changes to floral meristem.

This statement is **true**. Botanically, a flower is considered a highly modified and condensed reproductive shoot. The transition from vegetative growth to reproductive growth involves the transformation of the shoot apical meristem into a floral meristem, which then gives rise to the flower.

Step 3: Evaluating Reason R:

Reason R: Internode of the shoot gets condensed to produce different floral appendages laterally at successive nodes instead of leaves.

This statement is also **true**. The modification of the shoot involves a dramatic condensation of the axis, bringing the nodes very close together. The appendages produced at these nodes are modified leaves, which become the floral parts (sepals, petals, stamens, and carpels) arranged in whorls.

Step 4: Linking Assertion and Reason:

Reason R explains *how* the shoot is modified to become a flower. It details the process of internode condensation and the production of floral appendages instead of vegetative leaves. This directly explains why a flower is called a modified shoot. Therefore, R is the correct explanation for A.

Step 5: Final Answer:

Since both A and R are true, and R correctly explains A, the correct option is (2).

When tackling Assertion-Reason questions, use a two-step process: 1. Check the validity of each statement independently. 2. If both are true, check if the Reason logically explains the Assertion by asking "Why?" or "How?" after reading the Assertion. If the Reason answers that question, it's the correct explanation.

147. Match List I with List II:

List I List II

A. M Phase I. Proteins are synthesized

B. G₂ Phase II. Inactive phase

C. Quiescent stage III. Interval between mitosis and initiation of DNA replication

D. G₁ Phase IV. Equational division

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-I, D-III
- (B) A-III, B-II, C-IV, D-I
- (C) A-IV, B-II, C-I, D-III
- (D) A-IV, B-I, C-II, D-III

Correct Answer: (4) A-IV, B-I, C-II, D-III

Solution:

Step 1: Understanding the Question:

The question requires matching the phases of the cell cycle (List I) with their correct descriptions or events (List II).

Step 2: Detailed Explanation:

Let's analyze each phase in List I and find its corresponding description in List II.

- A. M Phase: This is the mitotic phase, where the cell divides. Mitosis is known as equational division because the number of chromosomes in the parent and daughter cells remains the same. So, A matches with IV.
- B. G₂ Phase: This is the gap 2 phase, which occurs after DNA synthesis (S phase) and before mitosis (M phase). During this phase, the cell continues to grow, and **proteins** are synthesized in preparation for mitosis, such as tubulin for spindle formation. So, B matches with I.
- C. Quiescent stage (G_0) : This is a period in the cell cycle where cells exist in a non-dividing state. They are metabolically active but have exited the cell cycle. This is

considered an inactive phase with respect to proliferation. So, C matches with II.

• D. G₁ Phase: This is the gap 1 phase, the first phase of interphase. It is the interval between mitosis (the previous M phase) and the initiation of DNA replication (the next S phase). The cell grows and carries out its normal metabolic functions during this phase. So, D matches with III.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow IV$

 $\mathrm{B} \to \mathrm{I}$

 $\mathrm{C} \to \mathrm{II}$

 $D \to III$

This corresponds to the option (4).

Quick Tip

Draw a diagram of the cell cycle $(G_1 \to S \to G_2 \to M)$ and write down the key event for each phase. Also, remember that the G_0 phase is an exit from the G_1 phase for non-dividing cells.

148. Given below are two statements:

Statement I: Gause's 'Competitive Exclusion Principle' states that two closely related species competing for the same resources cannot co-exist indefinitely and competitively inferior one will be eliminated eventually.

Statement II: In general, carnivores are more adversely affected by competition than herbivores.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is incorrect but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is correct but Statement II is false.

Correct Answer: (4) Statement I is correct but Statement II is false.

Solution:

Step 1: Understanding the Question:

The question presents two statements related to ecological principles of competition and asks

to evaluate their correctness.

Step 2: Evaluating Statement I:

Statement I provides a definition of Gause's 'Competitive Exclusion Principle'. It states that two species competing for the exact same limited resources cannot coexist in the same place at the same time; one species will have an advantage that will eventually lead to the elimination of the other. This is the precise and correct definition of the principle. Thus, **Statement I** is **correct**.

Step 3: Evaluating Statement II:

Statement II claims that carnivores are more adversely affected by competition than herbivores. This is generally false. Competition is typically more intense at lower trophic levels. Herbivores often compete for specific, stationary plant resources, which can lead to intense competition. Carnivores, on the other hand, often have larger territories, are more mobile, and may have a broader range of prey, which can reduce the intensity of direct competition. Furthermore, population sizes are smaller at higher trophic levels, which can also influence the dynamics of competition. Thus, Statement II is false.

Step 4: Final Answer:

Since Statement I is correct and Statement II is false, the correct option is (4).

Quick Tip

Remember Gause's principle with the example of Paramecium aurelia and Paramecium caudatum. When grown together, P. aurelia outcompetes P. caudatum for the limited food resources. Also, recall that competition intensity is generally greater at lower trophic levels.

149. Which one of the following statements is NOT correct?

- (A) The amount of some toxic substances of industrial waste water increases in the organisms at successive trophic levels.
- (B) The micro-organisms involved in biodegradation of organic matter in a sewage polluted water body consume a lot of oxygen causing the death of aquatic organisms.
- (C) Algal blooms caused by excess of organic matter in water improve water quality and promote fisheries.
- (D) Water hyacinth grows abundantly in eutrophic water bodies and leads to an imbalance in the ecosystem dynamics of the water body.

Correct Answer: (3) Algal blooms caused by excess of organic matter in water improve water quality and promote fisheries.

Solution:

Step 1: Understanding the Question:

The question asks to identify the incorrect statement among the given options related to water pollution and its ecological consequences.

Step 2: Detailed Explanation:

Let's analyze each statement:

- Statement (1): This describes the process of biomagnification, where the concentration of non-biodegradable toxic substances (like DDT or mercury) increases at each successive trophic level in a food chain. This statement is **correct**.
- Statement (2): This describes the effect of sewage pollution. When sewage (rich in organic matter) enters a water body, decomposer microorganisms multiply rapidly. Their respiratory activity consumes a large amount of dissolved oxygen, increasing the Biological Oxygen Demand (BOD). The depletion of oxygen can lead to the death of fish and other aquatic organisms. This statement is **correct**.
- Statement (3): Algal blooms are caused by an excess of nutrients (eutrophication), particularly nitrates and phosphates, not primarily organic matter itself (though decomposition of organic matter releases these nutrients). These blooms severely degrade water quality. They block sunlight from reaching submerged plants, and when the algae die, their decomposition by bacteria consumes vast amounts of dissolved oxygen, leading to hypoxic or anoxic conditions that kill fish. Therefore, they are detrimental to fisheries, not promotional. This statement is incorrect.
- Statement (4): Water hyacinth (Eichhornia crassipes) is a notorious invasive aquatic weed that thrives in nutrient-rich (eutrophic) water bodies. Its rapid growth covers the water surface, blocking light, reducing oxygen levels, and disrupting the entire aquatic ecosystem. This statement is **correct**.

Step 3: Final Answer:

The statement that is not correct is (3).

Quick Tip

Associate algal blooms with eutrophication and negative consequences: high BOD upon decay, oxygen depletion, and fish kills. They are a sign of poor water quality, not improvement.

150. Match List I with List II:

List I List II

A. Iron I. Synthesis of auxin

B. Zinc II. Component of nitrate reductase

C. Boron III. Activator of catalase

D. Molybdenum IV. Cell elongation and differentiation

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-I, D-III
- (B) A-III, B-II, C-I, D-IV
- (C) A-II, B-III, C-IV, D-I
- (D) A-III, B-I, C-IV, D-II

Correct Answer: (4) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:

The question requires matching the micronutrients in List I with their specific functions in plants from List II.

Step 2: Detailed Explanation:

Let's determine the function of each micronutrient:

- A. Iron (Fe): Iron is an essential component of proteins involved in electron transport, such as cytochromes and ferredoxin. It is also required for the formation of chlorophyll and acts as an activator for the catalase enzyme. So, A matches with III.
- B. Zinc (Zn): Zinc is required for the activity of various enzymes, especially carboxylases. It is also essential for the **synthesis of auxin**, a key plant growth hormone. So, B matches with I.
- C. Boron (B): Boron is required for the uptake and utilization of Ca²⁺, membrane functioning, pollen germination, cell elongation, and cell differentiation. So, C matches with IV.
- D. Molybdenum (Mo): Molybdenum is a component of several enzymes, including nitrogenase (involved in nitrogen fixation) and nitrate reductase, which is critical for converting nitrate to nitrite during nitrogen assimilation. So, D matches with II.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow III$

 $\mathrm{B} \to \mathrm{I}$

 $C \to IV$

 $D \rightarrow II$

This corresponds to the option (4).

Quick Tip

Create flashcards for essential plant nutrients. For each nutrient, list its key function(s) and a major deficiency symptom. For example: $Zn \to Auxin$ synthesis \to Deficiency causes 'little leaf' disease. Mo \to Nitrate reductase \to Deficiency affects nitrogen metabolism.

151. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R.

Assertion A: Amniocentesis for sex determination is one of the strategies of Reproductive and Child Health Care Programme.

Reason R: Ban on amniocentesis checks increasing menace of female foeticide. In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true and R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (1) A is false but R is true.

Solution:

Step 1: Understanding the Question:

The question presents an Assertion (A) and a Reason (R) related to the medical procedure amniocentesis. We need to evaluate the truthfulness of both statements and determine if R is the correct explanation for A.

Step 2: Evaluating Assertion A:

Assertion A states that amniocentesis for sex determination is a strategy of the Reproductive and Child Health Care (RCH) Programme. This is **false**. Amniocentesis is a prenatal diagnostic technique used to detect genetic abnormalities in the fetus. However, its use for sex determination is legally banned in India and is not a part of any government healthcare strategy. The RCH programme focuses on improving reproductive and child health, and promoting sex determination goes against its objectives of preventing female foeticide.

Step 3: Evaluating Reason R:

Reason R states that a ban on amniocentesis checks the increasing menace of female foeticide.

This is **true**. The misuse of amniocentesis to determine the sex of the fetus, followed by abortion if the fetus is female, led to a significant increase in female foeticide. To curb this social evil, the government of India enacted the Pre-conception and Pre-natal Diagnostic Techniques (PCPNDT) Act, 1994, which bans the use of this technique for sex determination.

Step 4: Final Answer:

Since Assertion A is false and Reason R is true, the correct option is (1).

Quick Tip

Remember that while amniocentesis itself is a valid medical procedure for detecting chromosomal abnormalities (like Down's syndrome), its misuse for sex determination is illegal and is the primary reason for the statutory ban. Government programs like RCH promote health, not practices that lead to social evils like female foeticide.

152. Vital capacity of lung is _____

- (A) IRV + ERV + TV
- (B) IRV + ERV
- (C) IRV + ERV + TV + RV
- (D) IRV + ERV + TV RV

Correct Answer: (1) IRV + ERV + TV

Solution:

Step 1: Understanding the Question:

The question asks for the correct formula representing the Vital Capacity (VC) of the lungs.

Step 2: Key Formula or Approach:

Vital Capacity is defined as the maximum volume of air a person can breathe out after a forced inspiration. It is the sum of three respiratory volumes:

- Tidal Volume (TV): Volume of air inspired or expired during a normal respiration (approx. 500 mL).
- Inspiratory Reserve Volume (IRV): Additional volume of air a person can inspire by a forcible inspiration (approx. 2500-3000 mL).
- Expiratory Reserve Volume (ERV): Additional volume of air a person can expire by a forcible expiration (approx. 1000-1100 mL).

The formula for Vital Capacity is:

$$VC = IRV + ERV + TV$$

Step 3: Detailed Explanation:

Let's analyze the components:

- \bullet A forced inspiration fills the lungs with TV + IRV.
- A forced expiration from this point empties the lungs of IRV + TV + ERV.
- This total exhaled volume is the Vital Capacity.

The other options are:

- (2) IRV + ERV: This sum is part of VC but omits the Tidal Volume.
- (3) IRV + ERV + TV + RV: This sum represents the Total Lung Capacity (TLC), which includes the Residual Volume (RV).
- (4) IRV + ERV + TV RV: This is not a standard respiratory capacity.

Step 4: Final Answer:

The correct formula for the vital capacity of the lung is IRV + ERV + TV.

Quick Tip

Remember the key respiratory capacities:

- Vital Capacity (VC) = TV + IRV + ERV (Maximum exchangeable air)
- Total Lung Capacity (TLC) = VC + RV (All air in the lungs)
- Inspiratory Capacity (IC) = TV + IRV
- Functional Residual Capacity (FRC) = ERV + RV

153. Match List I with List II.

List I

List II

- A. P-wave I. Beginning of systole
- B. Q-wave II. Repolarisation of ventricles
- C. QRS complex III. Depolarisation of atria
- D. T-wave IV. Depolarisation of ventricles

Choose the correct answer from the options given below:

- (A) A-I, B-II, C-III, D-IV
- (B) A-III, B-I, C-IV, D-II
- (C) A-IV, B-III, C-II, D-I
- (D) A-II, B-IV, C-I, D-III

Correct Answer: (2) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:

The question requires matching the different waves of an Electrocardiogram (ECG) in List I with the cardiac events they represent in List II.

Step 2: Detailed Explanation:

Let's analyze each component of the ECG from List I:

- A. P-wave: This small upward wave represents the electrical excitation, or depolarisation of the atria, which leads to the contraction of both atria. So, A matches with III.
- C. QRS complex: This complex starts shortly after the P-wave and represents the depolarisation of the ventricles, which initiates ventricular contraction (systole). So, C matches with IV.
- B. Q-wave: The Q-wave is the initial downward deflection of the QRS complex. Since the QRS complex as a whole initiates ventricular systole, the Q-wave marks the very beginning of systole. So, B matches with I.
- D. T-wave: This wave represents the return of the ventricles from the excited to the normal state, which is called **repolarisation of the ventricles**. The end of the T-wave marks the end of systole. So, D matches with II.

Step 3: Final Matching:

The correct pairings are:

 $A \rightarrow III$

 $\mathrm{B} \to \mathrm{I}$

 $\mathrm{C} \to \mathrm{IV}$

 $D \to II$

This set of matches corresponds to option (2).

Quick Tip

Remember the sequence: Atrial Depolarization (P-wave) \rightarrow Ventricular Depolarization (QRS complex) \rightarrow Ventricular Repolarization (T-wave). Depolarization leads to contraction (systole), and repolarization leads to relaxation (diastole).

154. Given below are two statements:

Statement I: A protein is imagined as a line, the left end represented by first amino acid (C-terminal) and the right end represented by last amino acid (N-terminal) Statement II: Adult human haemoglobin, consists of 4 subunits (two subunits of α type and two subunits of β type.)

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (1) Statement I is false but Statement II is true.

Solution:

Step 1: Understanding the Question:

The question presents two statements, one about the representation of protein structure and the other about the composition of hemoglobin. We need to evaluate the correctness of each statement.

Step 2: Evaluating Statement I:

Statement I describes the convention for representing a polypeptide chain. It claims the left end is the C-terminal and the right end is the N-terminal. This is **false**. By universal biochemical convention, a protein or polypeptide chain is written starting with the amino-terminal (N-terminal) end on the left and ending with the carboxyl-terminal (C-terminal) end on the right. The first amino acid is the N-terminal residue, and the last is the C-terminal residue.

Step 3: Evaluating Statement II:

Statement II describes the structure of adult human hemoglobin (HbA). It states that it consists of 4 subunits: two of α type and two of β type. This is **true**. Adult hemoglobin is a tetrameric protein with a quaternary structure, composed of two α -globin chains and two β -globin chains ($\alpha_2\beta_2$).

Step 4: Final Answer:

Since Statement I is false and Statement II is true, the correct option is (1).

Quick Tip

Remember the N to C convention for proteins. The N-terminus has a free amino group $(-NH_2)$ and is considered the beginning. The C-terminus has a free carboxyl group (-COOH) and is the end. Think of it as reading from "N" to "C".

155. Which of the following functions is carried out by cytoskeleton in a cell?

- (A) Transportation
- (B) Nuclear division
- (C) Protein synthesis

(D) Motility

Correct Answer: (4) Motility

Solution:

Step 1: Understanding the Question:

The question asks to identify a function performed by the cytoskeleton from the given options.

Step 2: Detailed Explanation:

The cytoskeleton is a network of protein filaments and tubules in the cytoplasm of many living cells, giving them shape and coherence. Its primary functions include:

- Mechanical Support: Maintaining the shape of the cell.
- Motility: This includes the movement of the entire cell (e.g., amoeboid movement, crawling) and the movement of structures within the cell. Cilia and flagella, which are responsible for cell movement, are made of microtubules (a component of the cytoskeleton).
- Intracellular Transport: It acts as a track for motor proteins to move vesicles and organelles around the cell.
- Cell Division: The formation of the mitotic spindle, which separates chromosomes during nuclear division (mitosis and meiosis), is a critical function of microtubules.

Step 3: Evaluating the Options:

- (1) Transportation: This is a function (intracellular transport), but "Motility" is a more direct and encompassing term for movement.
- (2) Nuclear division: The cytoskeleton is essential for this (spindle formation).
- (3) Protein synthesis: This is the function of ribosomes, not the cytoskeleton.
- (4) Motility: This is a major and well-recognized function of the cytoskeleton.

Both (1), (2), and (4) are functions of the cytoskeleton. However, in multiple-choice questions, we often need to choose the most prominent or direct function. Motility is a classic, defining feature of the cytoskeleton's role. Many textbooks highlight mechanical support and motility as the key functions. Therefore, Motility is the best answer among the choices.

Step 4: Final Answer:

Among the given options, motility is a key function carried out by the cytoskeleton.

Quick Tip

Associate the cytoskeleton with "Structure and Movement." It's the cell's skeleton and muscle system combined. Remember its three main components: microfilaments (cell shape, muscle contraction), intermediate filaments (anchorage), and microtubules (cell division, motility via cilia/flagella).

156. Select the correct group/set of Australian Marsupials exhibiting adaptive radiation.

- (A) Lemur, Anteater, Wolf
- (B) Tasmanian wolf, Bobcat, Marsupial mole
- (C) Numbat, Spotted cuscus, Flying phalanger
- (D) Mole, Flying squirrel, Tasmanian tiger cat

Correct Answer: (3) Numbat, Spotted cuscus, Flying phalanger

Solution:

Step 1: Understanding the Question:

The question asks to identify the option that contains only Australian marsupials, which are a classic example of adaptive radiation.

Step 2: Detailed Explanation:

Adaptive radiation is the evolution of different species from a common ancestor, each adapted to a different ecological niche. Australian marsupials are a prime example, having evolved from a common ancestor to fill niches occupied by placental mammals elsewhere in the world. We need to find the group that consists exclusively of these marsupials.

Step 3: Evaluating the Options:

- (1) Lemur, Anteater, Wolf: Lemurs are primates. Anteaters and wolves are placental mammals. This group is incorrect.
- (2) Tasmanian wolf, Bobcat, Marsupial mole: Tasmanian wolf (Thylacine) and Marsupial mole are marsupials. However, the Bobcat is a placental mammal (a feline). This group is incorrect.
- (3) Numbat, Spotted cuscus, Flying phalanger: The Numbat (or marsupial anteater), the Spotted cuscus (a type of possum), and the Flying phalanger (a glider) are all native Australian marsupials. This group is correct.
- (4) Mole, Flying squirrel, Tasmanian tiger cat: The Mole and the Flying squirrel are placental mammals that show convergent evolution with the Marsupial mole and Flying phalanger, respectively. The Tasmanian tiger cat (thylacine) is a marsupial. This group contains placental mammals and is incorrect.

Step 4: Final Answer:

The only set containing entirely Australian marsupials is Numbat, Spotted cuscus, and Flying phalanger.

To answer questions on adaptive radiation, be familiar with the key examples: Darwin's finches in the Galapagos, and Australian marsupials. Differentiate between marsupials and their placental mammal counterparts (e.g., Marsupial Mole vs. Placental Mole; Sugar Glider vs. Flying Squirrel).

157. Given below are two statements:

Statement I: RNA mutates at a faster rate.

Statement II: Viruses having RNA genome and shorter life span mutate and evolve faster.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (2) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:

The question asks us to evaluate two statements about the mutation and evolution rates of RNA and RNA viruses.

Step 2: Evaluating Statement I:

Statement I: RNA mutates at a faster rate. This statement is true. RNA is chemically less stable than DNA (due to the 2'-hydroxyl group on the ribose sugar). More importantly, the enzymes that replicate RNA (RNA-dependent RNA polymerases) lack the proofreading capabilities that DNA polymerases have. This lack of a "spell-check" function results in a much higher error rate during replication, leading to a faster mutation rate.

Step 3: Evaluating Statement II:

Statement II: Viruses having RNA genome and shorter life span mutate and evolve faster. This statement is also true. As established in Statement I, RNA genomes have a high mutation rate, which provides the raw material for natural selection. Combined with a very short generation time (life span), this allows for rapid evolutionary change. New variants can arise and be selected for in a very short period, as seen with viruses like influenza and coronaviruses.

Step 4: Final Answer:

Both statements are scientifically correct. Therefore, the correct option is (2).

Remember the link: Unstable RNA + No proofreading enzyme \rightarrow High mutation rate. High mutation rate + Short generation time \rightarrow Rapid evolution. This is why developing long-lasting vaccines against RNA viruses is challenging.

158. Once the undigested and unabsorbed substances enter the caecum, their backflow is prevented by-

- (A) Pyloric sphincter
- (B) Sphincter of Oddi
- (C) Ileo caecal valve
- (D) Gastro oesophageal sphincter

Correct Answer: (3) Ileo - caecal valve

Solution:

Step 1: Understanding the Question:

The question asks to identify the anatomical structure that prevents the backflow of material from the caecum (the beginning of the large intestine) into the small intestine.

Step 2: Detailed Explanation:

The digestive tract is equipped with several sphincters or valves that regulate the one-way flow of food and waste. Let's examine the location and function of each option:

- (1) Pyloric sphincter: Located between the stomach and the duodenum (the first part of the small intestine). It controls the passage of chyme from the stomach into the small intestine.
- (2) Sphincter of Oddi: Located at the junction of the common bile duct/pancreatic duct and the duodenum. It controls the flow of bile and pancreatic juices into the small intestine.
- (3) Ileo-caecal valve: Located at the junction of the ileum (the last part of the small intestine) and the caecum (the first part of the large intestine). Its primary function is to allow the passage of digested food from the ileum to the caecum and to prevent the backflow of fecal matter into the ileum.
- (4) Gastro-oesophageal sphincter: Also known as the cardiac sphincter, it is located between the oesophagus and the stomach. It prevents the acidic contents of the stomach from moving back up into the oesophagus.

Step 3: Final Answer:

The structure that prevents backflow from the caecum is the Ileo-caecal valve.

To remember the locations of digestive valves, trace the path of food: Oesophagus \rightarrow [Gastro-oesophageal sphincter] \rightarrow Stomach \rightarrow [Pyloric sphincter] \rightarrow Small Intestine (Ileum) \rightarrow [Ileo-caecal valve] \rightarrow Large Intestine (Caecum).

159. Which one of the following techniques does not serve the purpose of early diagnosis of a disease for its early treatment?

- (A) Enzyme Linked Immuno-Sorbent Assay (ELISA) technique
- (B) Recombinant DNA Technology
- (C) Serum and Urine analysis
- (D) Polymerase Chain Reaction (PCR) technique

Correct Answer: (3) Serum and Urine analysis

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the listed diagnostic techniques is generally not used for the *early* diagnosis of a disease, especially compared to the other options.

Step 2: Detailed Explanation:

Let's evaluate the sensitivity of each technique for early detection:

- (1) ELISA: This technique is based on antigen-antibody interactions and is highly sensitive. It can detect very low concentrations of antigens (like viral proteins) or antibodies produced by the body in response to an infection, often before symptoms are severe. It is used for the early diagnosis of diseases like HIV.
- (2) Recombinant DNA Technology: This technology allows the creation of DNA/RNA probes that can bind to the genetic material of a pathogen or a mutated gene. This allows for detection at the molecular level, which is a very early stage.
- (3) Serum and Urine analysis: These are conventional methods. They typically measure physiological or biochemical parameters (e.g., glucose in urine, liver enzymes in serum). Often, significant changes in these parameters occur only after the disease has progressed to a point where the pathogen concentration is high or organ damage has begun. Therefore, they are generally less sensitive for very early diagnosis compared to molecular techniques.
- (4) PCR: This is a molecular technique that can amplify a minute amount of DNA or RNA by millions of times. This makes it possible to detect the presence of a pathogen's genetic material even when it is present in very low numbers, long before the body mounts a detectable immune response or shows clinical symptoms. It is a cornerstone of early diagnosis.

Step 3: Final Answer:

Compared to the highly sensitive molecular techniques of ELISA, PCR, and Recombinant DNA technology, conventional serum and urine analysis is generally less effective for the *early* diagnosis of many diseases.

Quick Tip

Remember that modern molecular techniques like PCR and ELISA are designed for high sensitivity and specificity, allowing detection of pathogens or biomarkers at very low concentrations. This is the key to "early diagnosis." Conventional tests often require a higher concentration or a more advanced disease state to yield a positive result.

160. Given below are statements: one is labelled as Assertion A and the other is labelled as Reason R.

Assertion A: Nephrons are of two types: Cortical & Juxta medullary, based on their relative position in cortex and medulla.

Reason R: Juxta medullary nephrons have short loop of Henle whereas, cortical nephrons have longer loop of Henle.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (4) A is true but R is false.

Solution:

Step 1: Understanding the Question:

This is an Assertion-Reason question about the types of nephrons in the kidney. We need to assess if Assertion (A) and Reason (R) are individually correct.

Step 2: Evaluating Assertion A:

Assertion A states that nephrons are categorized into two types, Cortical and Juxtamedullary, based on their position in the cortex and medulla. This statement is **true**. The majority of nephrons are cortical, with their glomeruli in the outer cortex. Juxtamedullary nephrons have their glomeruli close to the junction of the cortex and medulla.

Step 3: Evaluating Reason R:

Reason R states that Juxtamedullary nephrons have short loops of Henle, while cortical nephrons have long loops of Henle. This statement is **false**. The opposite is true.

- Cortical nephrons have short loops of Henle that extend only a little way into the medulla.
- Juxtamedullary nephrons have very long loops of Henle that extend deep into the medulla. This feature is crucial for creating the concentration gradient in the medulla, which allows for the production of concentrated urine.

Step 4: Final Answer:

Since Assertion A is true and Reason R is false, the correct option is (4).

Quick Tip

Associate "Juxtamedullary" with "juxtaposed to the medulla" and "long loop". These long loops are essential for the counter-current mechanism and concentrating urine. Cortical nephrons, being mostly in the cortex, have shorter loops.

161. Which one of the following common sexually transmitted diseases is completely curable when detected early and treated properly?

- (A) HIV Infection
- (B) Genital herpes
- (C) Gonorrhoea
- (D) Hepatitis-B

Correct Answer: (3) Gonorrhoea

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the given sexually transmitted diseases (STDs) is completely curable, especially with early detection and treatment. The key distinction is between curable and manageable diseases.

Step 2: Detailed Explanation:

Let's analyze the treatability of each disease:

- (1) HIV Infection: Caused by the Human Immunodeficiency Virus (HIV). It is a viral disease that currently has no cure. However, it can be effectively managed with Antiretroviral Therapy (ART).
- (2) Genital herpes: Caused by the Herpes Simplex Virus (HSV). It is a viral infection that is not curable. The virus remains dormant in the body and can cause recurrent outbreaks. Antiviral medications can manage symptoms and reduce the frequency of outbreaks.

- (3) Gonorrhoea: Caused by the bacterium *Neisseria gonorrhoeae*. As a bacterial infection, it is **completely curable** with a course of appropriate antibiotics. Early treatment is crucial to prevent complications like pelvic inflammatory disease (PID) and infertility.
- (4) Hepatitis-B: Caused by the Hepatitis B Virus (HBV). Acute infection may resolve on its own, but it can become chronic, for which there is no cure. The chronic infection can be managed with antiviral drugs. A vaccine is available for prevention.

Step 3: Final Answer:

Among the given options, only Gonorrhoea, being a bacterial infection, is completely curable with proper antibiotic treatment.

Quick Tip

A general rule for STDs: bacterial infections (like Gonorrhoea, Syphilis, Chlamydia) are generally curable with antibiotics. Viral infections (like HIV, Herpes, Hepatitis-B, HPV) are generally not curable, but can often be managed or prevented with vaccines.

162. Which of the following statements are correct regarding female reproductive cycle?

- A. In non-primate mammals cyclical changes during reproduction are called oestrus cycle.
- B. First menstrual cycle begins at puberty and is called menopause.
- C. Lack of menstruation may be indicative of pregnancy.
- D. Cyclic menstruation extends between menarche and menopause.

Choose the most appropriate answer from the options given below:

- (A) A, C and D only
- (B) A and D only
- (C) A and B only
- (D) A, B and C only

Correct Answer: (1) A, C and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements about the female reproductive cycle from a list of four statements.

Step 2: Detailed Explanation:

Let's evaluate each statement:

• Statement A: In non-primate mammals cyclical changes during reproduction are called oestrus cycle. This is correct. Mammals like cows, sheep, dogs, etc., have

an oestrus cycle, characterized by a period of "heat" or oestrus when the female is sexually receptive. Primates (monkeys, apes, humans) have a menstrual cycle.

- Statement B: First menstrual cycle begins at puberty and is called menopause. This is incorrect. The first menstrual cycle is called menarche. Menopause is the permanent cessation of the menstrual cycle, which occurs later in life.
- Statement C: Lack of menstruation may be indicative of pregnancy. This is correct. Amenorrhea, or the absence of menstruation, is one of the primary and earliest signs of pregnancy. (Though it can also be caused by stress, poor health, etc.)
- Statement D: Cyclic menstruation extends between menarche and menopause. This is **correct**. The reproductive phase in human females is marked by the menstrual cycle, which starts at menarche and ends at menopause.

Step 3: Final Answer:

The correct statements are A, C, and D. This corresponds to option (1).

Quick Tip

Memorize the key terms for the female reproductive life stages: **Menarche** = Beginning of menstruation. **Menopause** = Cessation of menstruation. Differentiate clearly between the menstrual cycle (in primates) and the oestrus cycle (in non-primates).

163. Given below are two statements:

Statement I: Low temperature preserves the enzyme in a temporarily inactive state whereas high temperature destroys enzymatic activity because proteins are denatured by heat.

Statement II: When the inhibitor closely resembles the substrate in its molecular structure and inhibits the activity of the enzyme, it is known as competitive inhibitor.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (2) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:

The question presents two statements about enzyme kinetics and asks to evaluate their correctness.

Step 2: Evaluating Statement I:

Statement I describes the effect of temperature on enzyme activity. It states that low temperatures cause temporary inactivation, while high temperatures cause permanent destruction (denaturation). This is **true**. At low temperatures, enzymes become temporarily inactive because molecules have less kinetic energy, but their structure is preserved. Upon warming, their activity is restored. At high temperatures, the thermal energy breaks the weak bonds (like hydrogen bonds) that maintain the protein's specific three-dimensional structure. This irreversible change is called denaturation, and it destroys the enzyme's catalytic activity.

Step 3: Evaluating Statement II:

Statement II defines a competitive inhibitor. It states that a competitive inhibitor closely resembles the substrate and inhibits the enzyme. This is the correct definition of **competitive inhibition**. The inhibitor molecule competes with the normal substrate for binding to the active site of the enzyme.

Step 4: Final Answer:

Both Statement I and Statement II are correct and fundamental principles of enzymology. Therefore, the correct option is (2).

Quick Tip

For enzyme kinetics, remember:

- **Temperature:** Low temp = inactive, High temp = denatured.
- pH: Each enzyme has an optimal pH; extreme pH causes denaturation.
- Inhibition: Competitive inhibitors mimic the substrate and bind to the active site. Non-competitive inhibitors bind elsewhere (allosteric site) and change the enzyme's shape.

164. Match List I with List II.

List I List II

- A. Taenia I. Nephridia
- B. Paramoecium II. Contractile vacuole
- C. Periplaneta III. Flame cells
- D. Pheretima IV. Urecose gland

Choose the correct answer from the options give below:

- (A) A-II, B-I, C-IV, D-III
- (B) A-I, B-II, C-III, D-IV
- (C) A-I, B-II, C-IV, D-III
- (D) A-III, B-II, C-IV, D-I

Correct Answer: (4) A-III, B-II, C-IV, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching organisms from List I with their corresponding excretory or osmoregulatory structures from List II.

Step 2: Detailed Explanation:

Let's analyze each organism in List I and identify its correct structure from List II.

- A. Taenia (Tapeworm): It belongs to the phylum Platyhelminthes. The excretory system in Platyhelminthes consists of specialized cells called Flame cells (or protonephridia). So, A matches with III.
- B. Paramoecium: This is a single-celled protozoan. It lives in freshwater and uses a specialized organelle, the Contractile vacuole, for osmoregulation (to pump out excess water). So, B matches with II.
- C. Periplaneta (Cockroach): It belongs to the phylum Arthropoda. The primary excretory organs are Malpighian tubules. However, cockroaches also have fat bodies and Urecose glands which are involved in the storage and excretion of uric acid. So, C matches with IV.
- D. Pheretima (Earthworm): It belongs to the phylum Annelida. The excretory organs in annelids are coiled tubular structures called Nephridia. So, D matches with I.

Step 3: Final Matching:

The correct pairings are:

 $A \rightarrow III$

 $B \to II$

 $C \to IV$

 $\mathrm{D} \to \mathrm{I}$

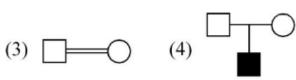
This set of matches corresponds to option (4).

Create a table of major animal phyla and their characteristic excretory structures. Key pairings to remember:

- Platyhelminthes \rightarrow Flame Cells
- \bullet Annelida \rightarrow Nephridia
- Arthropoda (Insects) \rightarrow Malpighian Tubules
- \bullet Protozoa (freshwater) \rightarrow Contractile Vacuole

165. Which one of the following symbols represents mating between relatives in human pedigree analysis?





- (A)(1)
- (B)(2)
- (C)(3)
- (D) (4)

Correct Answer: (3)

Solution:

Step 1: Understanding the Question:

The question asks to identify the standard symbol used in a human pedigree chart to represent consanguineous mating, which is mating between closely related individuals.

Step 2: Detailed Explanation of Pedigree Symbols:

In pedigree analysis, specific symbols are used to represent individuals, their relationships, and their traits.

- A square represents a male.
- A circle represents a female.
- A horizontal line connecting a square and a circle represents mating.
- A filled symbol represents an affected individual (possessing the trait of interest).
- An unfilled symbol represents an unaffected individual.

• A double horizontal line connecting a square and a circle specifically indicates a consanguineous mating (mating between relatives).

Step 3: Evaluating the Options:

- Option (1) shows a mating between an affected male and an affected female. The single line indicates they are not necessarily relatives.
- Option (2) shows a mating between an unaffected male and an unaffected female. The single line indicates they are not necessarily relatives.
- Option (3) shows a mating between an unaffected male and an unaffected female, connected by a **double line**. This is the correct symbol for consanguineous mating.
- Option (4) shows a mating between an affected male and an unaffected female. The single line indicates they are not necessarily relatives.

Step 4: Final Answer:

The symbol with a double line between the male and female represents mating between relatives.

Quick Tip

In pedigree charts, pay close attention to the connecting lines. A single horizontal line is for a standard mating, while a double horizontal line specifically signifies a consanguineous mating. This is often a key clue for identifying recessive genetic disorders.

166. Given below are two statements:

Statement I: Vas deferens receives a duct from seminal vesicle and opens into urethra as the ejaculatory duct.

Statement II: The cavity of the cervix is called cervical canal which along with vagina forms birth canal.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I incorrect but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is correct but Statement II is false.

Correct Answer: (2) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:

The question asks to evaluate the correctness of two statements related to the human male and

female reproductive systems.

Step 2: Evaluating Statement I:

This statement describes a part of the male reproductive tract. The vas deferens is the tube that carries sperm from the epididymis. It loops over the urinary bladder and receives a duct from the seminal vesicle. The union of the vas deferens and the duct of the seminal vesicle forms the ejaculatory duct. This ejaculatory duct then passes through the prostate gland and opens into the urethra. The statement accurately describes this anatomical pathway. Thus, **Statement I is true**.

Step 3: Evaluating Statement II:

This statement describes a part of the female reproductive tract. The cervix is the lower, narrow part of the uterus. The cavity within the cervix is called the cervical canal. During childbirth (parturition), the baby passes from the uterus, through the cervical canal, and then through the vagina to the outside. The combination of the cervical canal and the vagina constitutes the birth canal. This statement is also anatomically correct. Thus, **Statement II is true**.

Step 4: Final Answer:

Since both statements are correct, the correct option is (2).

Quick Tip

For anatomy questions, it's helpful to visualize or sketch the pathways. For the male tract, trace the path of sperm: Seminiferous tubules \rightarrow Epididymis \rightarrow Vas deferens \rightarrow Ejaculatory duct \rightarrow Urethra. For the female tract, remember that the birth canal is a composite structure: Cervical canal + Vagina.

167. Match List I with List II.

List I List II

- A. CCK I. Kidney
- B. GIP II. Heart
- C. ANF III. Gastric gland
- D. ADH IV. Pancreas

Choose the correct answer from the options given below:

- (A) A-IV, B-II, C-III, D-I
- (B) A-IV, B-III, C-II, D-I
- (C) A-III, B-II, C-IV, D-I
- (D) A-II, B-IV, C-I, D-III

Correct Answer: (2) A-IV, B-III, C-II, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching hormones/peptides in List I with their target organs or sites of action in List II.

Step 2: Detailed Explanation:

Let's analyze each item in List I and find its correct match in List II.

- A. CCK (Cholecystokinin): This is a gastrointestinal hormone released from the duodenum. It acts on two main organs: the gallbladder (to release bile) and the Pancreas (to release digestive enzymes). So, A matches with IV.
- B. GIP (Gastric Inhibitory Peptide): This is another gastrointestinal hormone. As its name suggests, it inhibits gastric secretion and motility. Therefore, it acts on the Gastric gland. So, B matches with III.
- C. ANF (Atrial Natriuretic Factor): This peptide hormone is secreted by the atrial walls of the **Heart** in response to high blood pressure. It causes vasodilation and excretion of sodium and water by the kidneys, thus lowering blood pressure. So, C matches with II.
- D. ADH (Antidiuretic Hormone): Also known as vasopressin, this hormone is released from the posterior pituitary. It acts primarily on the distal convoluted tubules (DCT) and collecting ducts of the **Kidney**, promoting water reabsorption. So, **D matches with I**.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow IV$

 $\mathrm{B} \to \mathrm{III}$

 $C \to II$

 $\mathrm{D} o \mathrm{I}$

This corresponds to option (2).

Quick Tip

Focus on the names of the hormones as they often give clues to their function.

- Cholecysto-kinin: "Cholecysto" = gallbladder, "kinin" = movement.
- Gastric Inhibitory Peptide: Self-explanatory.
- Atrial Natriuretic Factor: "Atrial" = from heart atria.
- Anti-diuretic Hormone: "Anti-diuresis" = against urination.

168. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R.

Assertion A: Endometrium is necessary for implantation of blastocyst.

Reason R: In the absence of fertilization, the corpus luteum degenerates that causes disintegration of endometrium.

In the light of the above statements, choose the correct answer from the options given below:

- (A) A is false but R is true.
- (B) Both A and R are true and R is the correct explanation of A.
- (C) Both A and R are true but R is NOT the correct explanation of A.
- (D) A is true but R is false.

Correct Answer: (3) Both A and R are true but R is NOT the correct explanation of A.

Solution:

Step 1: Understanding the Question:

This is an Assertion-Reason question about the female reproductive cycle, specifically concerning the endometrium, implantation, and the corpus luteum.

Step 2: Evaluating Assertion A:

Assertion A states that the endometrium is necessary for the implantation of the blastocyst. This is **true**. The endometrium is the inner lining of the uterus, which becomes thick, vascularized, and rich in glands under the influence of progesterone. This prepared lining is essential for the blastocyst to attach and embed itself, a process called implantation.

Step 3: Evaluating Reason R:

Reason R describes what happens if fertilization does not occur. It states that the corpus luteum degenerates, which causes the disintegration of the endometrium. This is also **true**. The corpus luteum produces progesterone, which maintains the endometrium. If fertilization does not happen, the corpus luteum degenerates after about 10-12 days. The subsequent sharp decline in progesterone levels leads to the breakdown and shedding of the endometrial lining, resulting in menstruation.

Step 4: Linking Assertion and Reason:

Now, we must check if Reason R is the correct explanation for Assertion A. Assertion A is about what is needed for implantation (when fertilization occurs). Reason R is about what happens in the absence of fertilization. Reason R explains why menstruation occurs, but it does not explain why the endometrium is necessary for implantation. The correct explanation for Assertion A would be that the endometrium provides the necessary nutrients and structural support for the developing embryo. Since R describes a different scenario (no fertilization), it is not the correct explanation for A.

Step 5: Final Answer:

Both statements are individually true, but the reason does not correctly explain the assertion. Therefore, the correct option is (3).

Quick Tip

For Assertion-Reason questions, first validate each statement. If both are true, carefully check the link. Ask: "Does Reason R explain why Assertion A is true?" In this case, R explains menstruation, while A is about implantation. These are related but distinct events in the menstrual cycle.

169. Match List I with List II.

List I List II

- A. Gene 'a' I. β -galactosidase
- B. Gene 'y' II. Transacetylase
- C. Gene 'i' III. Permease
- D. Gene 'z' IV. Repressor protein

Choose the correct answer from the options given below:

- (A) A-III, B-I, C-IV, D-II
- (B) A-II, B-I, C-IV, D-III
- (C) A-II, B-III, C-IV, D-I
- (D) A-III, B-IV, C-I, D-II

Correct Answer: (3) A-II, B-III, C-IV, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching the genes of the lac operon (List I) with the proteins they code for (List II).

Step 2: Detailed Explanation:

The lac operon in E. coli consists of several genes involved in lactose metabolism.

- Gene 'i' (Regulator gene): This gene is not part of the operon itself but regulates it. It codes for the Repressor protein. The repressor binds to the operator region to switch the operon off. So, C matches with IV.
- Gene 'z' (Structural gene): This gene codes for the enzyme β -galactosidase, which breaks down lactose into glucose and galactose. So, **D** matches with **I**.

- Gene 'y' (Structural gene): This gene codes for the protein Permease, which is a transmembrane protein that facilitates the transport of lactose into the cell. So, B matches with III.
- Gene 'a' (Structural gene): This gene codes for the enzyme Transacetylase, whose role in lactose metabolism is not fully understood but is believed to be involved in detoxification. So, A matches with II.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to III$

 $C \to IV$

 $D \to I$

This corresponds to option (3).

Quick Tip

Remember the order of the structural genes in the lac operon: z, y, a. And their products:

- $\mathbf{z} \to \beta$ -galactosidase
- ullet y o permease
- $\mathbf{a} \to \text{transacetylase}$

Also, remember the regulator i gene produces the **repressor**.

170. Which of the following is not a cloning vector?

- (A) Probe
- (B) BAC
- (C) YAC
- (D) pBR322

Correct Answer: (1) Probe

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the given options is not used as a cloning vector in biotechnology. A cloning vector is a DNA molecule that can carry foreign DNA into a host cell and replicate there.

Step 2: Detailed Explanation:

Let's analyze each option:

- (1) Probe: A DNA or RNA probe is a short, single-stranded sequence of nucleic acid that is labeled (e.g., with a radioactive isotope or a fluorescent tag). It is used to detect the presence of a complementary sequence in a sample of DNA or RNA through hybridization. It is a detection tool, **not a cloning vector**.
- (2) BAC (Bacterial Artificial Chromosome): This is a high-capacity cloning vector based on the F-plasmid of E. coli. It can carry large DNA inserts (100-300 kilobase pairs) and is used in genome sequencing projects. It is a cloning vector.
- (3) YAC (Yeast Artificial Chromosome): This is another high-capacity cloning vector that can be replicated in yeast cells. It can carry very large DNA inserts (up to a million base pairs or more). It is a cloning vector.
- (4) pBR322: This is one of the first widely used E. coli cloning vectors. It is a plasmid and is used for cloning smaller DNA fragments. It is a classic example of a cloning vector.

Step 3: Final Answer:

BAC, YAC, and pBR322 are all types of cloning vectors used to carry and replicate foreign DNA. A probe is a tool used for detection, not for cloning. Therefore, 'Probe' is not a cloning vector.

Quick Tip

Associate "vector" with "vehicle". Cloning vectors are DNA vehicles used to carry genes into cells. Common vectors include plasmids, bacteriophages, cosmids, BACs, and YACs. A "probe" is like a "detector" or "sensor" used to find a specific gene sequence.

171. In which blood corpuscles, the HIV undergoes replication and produces progeny viruses?

- (A) Eosinophils
- (B) T_H cells
- (C) B-lymphocytes
- (D) Basophils

Correct Answer: (2) T_H cells

Solution:

Step 1: Understanding the Question:

The question asks to identify the specific type of blood cell in which the Human Immunodeficiency Virus (HIV) primarily replicates.

Step 2: Detailed Explanation:

HIV is a retrovirus that attacks the immune system. Its life cycle involves the following key steps:

- 1. The virus primarily targets cells that have the CD4 protein on their surface.
- 2. The most important of these cells are the **Helper T-lymphocytes**, also known as T_H cells or CD4⁺ T cells. Macrophages are also infected.
- 3. After entering a T_H cell, the virus uses its reverse transcriptase enzyme to convert its RNA genome into DNA.
- 4. This viral DNA is then integrated into the host cell's DNA.
- 5. The infected T_H cell is then forced to produce new virus particles (progeny viruses).
- 6. These new viruses are released from the cell, often killing it in the process, and go on to infect other T_H cells.

The progressive destruction of T_H cells severely weakens the immune system, leading to Acquired Immuno Deficiency Syndrome (AIDS).

Step 3: Evaluating the Options:

• (1) Eosinophils, (3) B-lymphocytes, and (4) Basophils are other types of white blood cells, but they are not the primary targets for HIV replication. The key target is the T_H cell.

Step 4: Final Answer:

HIV replicates and produces progeny viruses primarily within Helper T-cells (T_H cells).

Quick Tip

Remember that HIV's main target is the "commander" of the immune system: the Helper T-cell (T_H cell). By destroying these cells, HIV cripples the body's ability to coordinate an immune response, leading to AIDS. The term "CD4 count" refers to the number of these specific cells.

172. Broad palm with single palm crease is visible in a person suffering from-

- (A) Thalassemia
- (B) Down's syndrome
- (C) Turner's syndrome
- (D) Klinefelter's syndrome

Correct Answer: (2) Down's syndrome

Solution:

Step 1: Understanding the Question:

The question asks to identify the genetic disorder characterized by a specific physical feature: a broad palm with a single transverse palmar crease (also known as a similar crease).

Step 2: Detailed Explanation:

Let's analyze the characteristic features of the given disorders:

- (1) Thalassemia: This is an inherited blood disorder characterized by less hemoglobin and fewer red blood cells than normal. Its symptoms are related to anemia, not specific physical features like palm creases.
- (2) Down's syndrome: This is a chromosomal disorder caused by the presence of a full or partial extra copy of chromosome 21 (Trisomy 21). It is characterized by a distinct set of physical features, including a small round head, a flattened facial profile, partially open mouth with a protruding furrowed tongue, and short stature. A very characteristic feature is a broad palm with a single transverse palmar crease.
- (3) Turner's syndrome: This is a chromosomal disorder in females caused by the absence of one of the X chromosomes (XO karyotype). Features include short stature, webbed neck, and rudimentary ovaries.
- (4) Klinefelter's syndrome: This is a chromosomal disorder in males caused by an extra X chromosome (XXY karyotype). Features include tall stature, overall masculine development but with some feminine characteristics (like gynecomastia), and sterility.

Step 3: Final Answer:

A broad palm with a single palmar crease is a classic diagnostic feature of Down's syndrome.

Quick Tip

Create a table for common chromosomal disorders (Down's, Turner's, Klinefelter's) and list their cause (e.g., Trisomy 21, XO, XXY) and 3-4 key characteristic physical features for each. The single palmar crease is a very specific and frequently tested feature of Down's syndrome.

173. Match List I with List II.

List I List II

- A. Heroin I. Effect on cardiovascular system
- B. Marijuana II. Slow down body function
- C. Cocaine III. Painkiller
- D. Morphine IV. Interfere with transport of dopamine

Choose the correct answer from the options given below:

- (A) A-III, B-IV, C-I, D-II
- (B) A-II, B-I, C-IV, D-III
- (C) A-I, B-II, C-III, D-IV

(D) A-IV, B-III, C-II, D-I

Correct Answer: (2) A-II, B-I, C-IV, D-III

Solution:

Step 1: Understanding the Question:

The question requires matching drugs from List I with their primary physiological effects or mechanisms of action from List II.

Step 2: Detailed Explanation:

Let's analyze each drug in List I:

- A. Heroin: Also known as diacetylmorphine, it is an opioid and a powerful depressant of the central nervous system. Its primary effect is to slow down body function. So, A matches with II.
- B. Marijuana: The active components are cannabinoids. These substances are known to have a significant effect on the cardiovascular system, often increasing heart rate and affecting blood pressure. So, B matches with I.
- C. Cocaine: This is a powerful central nervous system stimulant. Its mechanism of action involves blocking the reuptake of neurotransmitters, particularly dopamine, in the brain's reward pathway. Thus, it does interfere with the transport of dopamine. So, C matches with IV.
- D. Morphine: This is a potent opioid analgesic (painkiller) extracted from the opium poppy. It is widely used in medicine to relieve severe pain. So, D matches with III (Painkiller).

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $\mathrm{B} \to \mathrm{I}$

 $C \to IV$

 $D \to III$

This corresponds to option (2).

Categorize common drugs into classes:

- Opioids (Depressants/Painkillers): Morphine, Heroin. They slow down body functions.
- Stimulants: Cocaine, Amphetamines. They interfere with neurotransmitters like dopamine.
- Cannabinoids: Marijuana, Hashish. Known for cardiovascular and psychoactive effects.
- Hallucinogens: LSD, Datura. Cause hallucinations.

174. Match List I with List II.

List I (Type of Joint) List II (Found between)

- A. Cartilaginous Joint I. Between flat skull bones
- B. Ball and Socket Joint II. Between adjacent vertebrae in vertebral column
- C. Fibrous Joint III. Between carpal and metacarpal of thumb
- D. Saddle Joint IV. Between Humerus and Pectoral girdle

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-III, D-I
- (B) A-III, B-I, C-II, D-IV
- (C) A-II, B-IV, C-I, D-III
- (D) A-I, B-IV, C-III, D-II

Correct Answer: (3) A-II, B-IV, C-I, D-III

Solution:

Step 1: Understanding the Question:

The question requires matching different types of joints from List I with their correct location in the human skeleton from List II.

Step 2: Detailed Explanation:

Let's analyze each joint type in List I:

- A. Cartilaginous Joint: These joints have bones connected by cartilage and allow for limited movement. A prime example is the joints between adjacent vertebrae in the vertebral column, where intervertebral discs (made of fibrocartilage) are found. So, A matches with II.
- B. Ball and Socket Joint: This is a type of synovial joint that allows for a wide range of motion. It is formed when the ball-shaped head of one bone fits into the cup-like socket of another. The shoulder joint, between the Humerus and the Pectoral girdle

(specifically, the glenoid cavity of the scapula), is a classic example. So, **B matches with IV**.

- C. Fibrous Joint: These joints are connected by dense fibrous connective tissue and allow for no movement. The sutures between the flat skull bones are a perfect example. So, C matches with I.
- **D. Saddle Joint:** This is another type of synovial joint that allows for movement in two planes (biaxial). The best example in the human body is the first carpometacarpal joint, i.e., the joint between the carpal (trapezium) and the metacarpal of the thumb. So, **D** matches with III.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to IV$

 $C \to I$

 $\mathrm{D} \to \mathrm{III}$

This corresponds to option (3).

Quick Tip

For joints, focus on the degree of movement and one key example for each type:

- Fibrous (immovable): Skull sutures.
- Cartilaginous (slightly movable): Vertebrae.
- Synovial (freely movable):
 - Ball and Socket (shoulder, hip)
 - Hinge (elbow, knee)
 - Saddle (thumb)
 - Pivot (atlas/axis)

175. Match List I with List II with respect to human eye.

List I List II

- A. Fovea I. Visible coloured portion of eye that regulates diameter of pupil.
- B. Iris II. External layer of eye formed of dense connective tissue.
- C. Blind spot III. Point of greatest visual acuity or resolution.
- D. Sclera IV. Point where optic nerve leaves the eyeball and photoreceptor cells are absent.

Choose the correct answer from the options given below:

- (A) A-II, B-I, C-III, D-IV
- (B) A-III, B-I, C-IV, D-II
- (C) A-IV, B-III, C-II, D-I

(D) A-I, B-IV, C-III, D-II

Correct Answer: (2) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:

The question requires matching parts of the human eye from List I with their corresponding descriptions or functions from List II.

Step 2: Detailed Explanation:

Let's analyze each part of the eye from List I:

- A. Fovea: The fovea is a small depression in the retina's yellow spot (macula lutea). It is densely packed with only cones and is responsible for sharp, detailed central vision. It is the point of greatest visual acuity or resolution. So, A matches with III.
- B. Iris: This is the pigmented part of the eye that gives it its color. It is a muscular diaphragm that surrounds the pupil and regulates the diameter of the pupil, controlling the amount of light entering the eye. So, B matches with I.
- C. Blind spot: This is the point on the retina where the axons of the ganglion cells converge to form the optic nerve and exit the eyeball. At this location, there are no photoreceptor cells (rods or cones), hence it is insensitive to light. So, C matches with IV.
- D. Sclera: This is the tough, white, opaque outer layer of the eyeball. It is composed of dense connective tissue and serves to protect the eye and maintain its shape. So, D matches with II.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow III$

 $B \to I$

 $C \rightarrow IV$

 $\mathrm{D} \to \mathrm{II}$

This corresponds to option (2).

Quick Tip

Remember the key functional spots on the retina:

- Fovea/Macula Lutea: Best vision (only cones).
- Blind Spot: No vision (no photoreceptors, only optic nerve exit).

And the layers of the eye from outside in: Sclera \rightarrow Choroid \rightarrow Retina.

176. Given below are two statements:

Statement I: Ligaments are dense irregular tissue.

Statement II: Cartilage is dense regular tissue.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I is false but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is true but Statement II is false.

Correct Answer: (3) Both Statement I and Statement II are false.

Solution:

Step 1: Understanding the Question:

The question asks us to evaluate two statements about the classification of connective tissues, specifically ligaments and cartilage.

Step 2: Evaluating Statement I:

Statement I claims that ligaments are dense irregular tissue. This is false. Ligaments are examples of dense regular connective tissue. In this tissue, the collagen fibers are arranged in parallel bundles, providing great tensile strength in one direction, which is necessary to connect bones to other bones. Dense irregular tissue, found in the dermis of the skin, has collagen fibers arranged randomly to resist tension from multiple directions.

Step 3: Evaluating Statement II:

Statement II claims that cartilage is dense regular tissue. This is also false. Cartilage is a type of specialized connective tissue. It is distinct from both dense regular and dense irregular connective tissue. Cartilage has a pliable matrix due to chondroitin salts and contains cells called chondrocytes enclosed in lacunae. It does not fit the description of dense regular tissue.

Step 4: Final Answer:

Since both statements incorrectly classify the tissues, both Statement I and Statement II are false.

For connective tissues, remember the main categories and key examples:

- Loose Connective Tissue: Areolar, Adipose.
- Dense Connective Tissue:
 - Regular: Tendons (muscle to bone), Ligaments (bone to bone). Fibers are parallel.
 - Irregular: Dermis of skin. Fibers are randomly arranged.
- Specialized Connective Tissue: Cartilage, Bone, Blood.

177. Match List I with List II.

List I

- A. Vasectomy I. Oral method
- B. Coitus interruptus II. Barrier methodC. Cervical caps III. Surgical method
- D. Saheli IV. Natural method

Choose the correct answer from the options given below:

List II

- (A) A-IV, B-II, C-I, D-III
- (B) A-III, B-I, C-IV, D-II
- (C) A-III, B-IV, C-II, D-I
- (D) A-II, B-III, C-I, D-IV

Correct Answer: (3) A-III, B-IV, C-II, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching different contraceptive methods from List I with their correct category from List II.

Step 2: Detailed Explanation:

Let's analyze each contraceptive method in List I:

- A. Vasectomy: This is a permanent method of contraception in males where the vas deferens is cut and tied. This prevents sperm from entering the semen. It is a Surgical method. So, A matches with III.
- B. Coitus interruptus: Also known as the withdrawal method, this involves the male withdrawing the penis from the vagina before ejaculation to prevent deposition of semen. This is a traditional Natural method of contraception. So, B matches with IV.

- C. Cervical caps: These are devices made of rubber that are inserted into the vagina to cover the cervix before intercourse. They physically prevent sperm from entering the uterus. This is a type of Barrier method. So, C matches with II.
- **D. Saheli:** This is a contraceptive pill developed in India. It is a non-steroidal preparation taken once a week. As it is a pill taken by mouth, it is an **Oral method**. So, **D matches with I**.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow III$

 $B \to IV$

 $C \to II$

 $\mathrm{D} \to \mathrm{I}$

This corresponds to option (3).

Quick Tip

Categorize contraceptive methods for easy recall:

- Natural: Periodic abstinence, Coitus interruptus, Lactational amenorrhea.
- Barrier: Condoms, Diaphragms, Cervical caps, Vaults.
- IUDs: Non-medicated (Lippes loop), Copper-releasing (CuT), Hormone-releasing (Progestasert).
- Oral/Hormonal: Pills (Saheli, Mala-D), Injections, Implants.
- Surgical (Sterilization): Vasectomy (male), Tubectomy (female).

178. Match List I with List II.

List I (Interacting species)

- A. A Leopard and a Lion in a forest/grassland
- B. A Cuckoo laying egg in a Crow's nest
- C. Fungi and root of a higher plant in Mycorrhizae
- D. A cattle egret and a Cattle in a field

List II (Name of Interaction)

- I. Competition
- II. Brood parasitism
- III. Mutualism
- IV. Commensalism

Choose the correct answer from the options given below:

- (A) A-II, B-III, C-I, D-IV
- (B) A-I, B-II, C-III, D-IV
- (C) A-I, B-II, C-IV, D-III
- (D) A-III, B-IV, C-I, D-II

Correct Answer: (2) A-I, B-II, C-III, D-IV

Solution:

Step 1: Understanding the Question:

The question requires matching pairs of interacting species from List I with the correct ecological term for their interaction from List II.

Step 2: Detailed Explanation:

Let's analyze each interaction in List I:

- A. A Leopard and a Lion in a forest/grassland: Both are top predators that may hunt similar prey (e.g., deer, zebra). Since they share and compete for the same limited food resources, their interaction is Competition (-/-). So, A matches with I.
- B. A Cuckoo laying egg in a Crow's nest: The cuckoo lays its eggs in the nest of another species (the crow), which then incubates the eggs and raises the cuckoo chick, often at the expense of its own offspring. This is a classic example of **Brood parasitism** (+/-). So, **B matches with II**.
- C. Fungi and root of a higher plant in Mycorrhizae: This is a symbiotic association where the fungus helps the plant with nutrient and water absorption from the soil, and the plant provides carbohydrates (food) to the fungus. Both partners benefit. This interaction is Mutualism (+/+). So, C matches with III.
- D. A cattle egret and a Cattle in a field: The egret follows the cattle and feeds on insects that are stirred up from the vegetation as the cattle graze. The egret benefits (gets food easily), while the cattle is neither harmed nor benefited. This interaction is Commensalism (+/0). So, D matches with IV.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow I$

 $B \rightarrow II$

 $C \to III$

 $D \to IV$

This corresponds to option (2), as all pairs are matched with their correct definitions.

Use a simple notation to remember population interactions:

- Mutualism (+/+): Both benefit (e.g., Lichens, Mycorrhizae).
- Competition (-/-): Both are harmed.
- Predation/Parasitism (+/-): One benefits, one is harmed.
- Commensalism (+/0): One benefits, one is unaffected (e.g., Orchid on a mango tree).
- Amensalism (-/0): One is harmed, one is unaffected (e.g., Penicillium and bacteria).

179. Match List I with List II.

List I

List II

- A. Ringworm I. Haemophilus influenzae
- B. Filariasis II. Trichophyton
- C. Malaria III. Wuchereria bancrofti
- D. Pneumonia IV. Plasmodium vivax

Choose the correct answer from the options given below:

- (A) A-III, B-II, C-IV, D-I
- (B) A-II, B-III, C-IV, D-I
- (C) A-II, B-III, C-I, D-IV
- (D) A-III, B-II, C-I, D-IV

Correct Answer: (2) A-II, B-III, C-IV, D-I

Solution:

Step 1: Understanding the Question:

The question requires matching diseases from List I with their causative organisms from List II.

Step 2: Detailed Explanation:

Let's analyze each disease in List I and find its pathogen in List II.

- A. Ringworm: Despite its name, ringworm is not caused by a worm. It is a common fungal infection of the skin. The fungi responsible belong to genera like **Trichophyton**, *Microsporum*, and *Epidermophyton*. So, A matches with II.
- B. Filariasis (Elephantiasis): This is a parasitic disease caused by a filarial worm. The most common causative agent is Wuchereria bancrofti. So, B matches with III.

- C. Malaria: This is a protozoan disease transmitted by the female *Anopheles* mosquito. The causative agent is a species of **Plasmodium**, such as **Plasmodium vivax** or *P. falciparum*. So, C matches with IV.
- **D. Pneumonia:** This is an infection of the lungs that can be caused by various microorganisms. A common bacterial cause of pneumonia is **Haemophilus influenzae** (another is *Streptococcus pneumoniae*). So, **D matches with I**.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to III$

 $C \to IV$

 $\mathrm{D} \to \mathrm{I}$

This corresponds to option (2).

Quick Tip

For disease questions, create a table with four columns: Disease, Causative Organism, Type of Organism (Virus, Bacterium, Protozoan, Fungus, Helminth), and Mode of Transmission/Vector. This organized approach helps in memorizing and recalling information accurately.

180. Radial symmetry is NOT found in adults of phylum _____

- (A) Echinodermata
- (B) Ctenophora
- (C) Hemichordata
- (D) Coelenterata

Correct Answer: (3) Hemichordata

Solution:

Step 1: Understanding the Question:

The question asks to identify the animal phylum whose adult members do not exhibit radial symmetry.

Step 2: Detailed Explanation:

Symmetry is a fundamental feature in animal classification.

• Radial symmetry: The body can be divided into two equal halves by any plane passing through the central axis.

• Bilateral symmetry: The body can be divided into identical right and left halves in only one plane.

Let's analyze the symmetry of the given phyla:

- (1) Echinodermata: A unique feature of this phylum (e.g., starfish, sea urchins) is that the adults are radially symmetrical (usually pentamerous), while their larvae are bilaterally symmetrical.
- (2) Ctenophora (Comb jellies): These animals are exclusively marine and exhibit radial symmetry (specifically, biradial symmetry).
- (3) Hemichordata (e.g., Balanoglossus): These are worm-like marine animals. They are exclusively bilaterally symmetrical throughout their life. They do not show radial symmetry.
- (4) Coelenterata (Cnidaria, e.g., jellyfish, corals): These animals typically exhibit radial symmetry.

Step 3: Final Answer:

Adults of the phylum Hemichordata are bilaterally symmetrical, not radially symmetrical.

Quick Tip

Remember the three major radially symmetrical phyla: Coelenterata (Cnidaria), Ctenophora, and adult Echinodermata. Most other higher phyla, including Hemichordata and Chordata, are bilaterally symmetrical.

181. Given below are two statements:

Statement I: Electrostatic precipitator is most widely used in thermal power plant. Statement II: Electrostatic precipitator in thermal power plant removes ionising radiations

In the light of the above statements, choose the most appropriate answer from the options given below:

- (A) Statement I incorrect but Statement II is correct.
- (B) Both Statement I and Statement II are correct.
- (C) Both Statement I and Statement II are incorrect.
- (D) Statement I is correct but Statement II is incorrect.

Correct Answer: (4) Statement I is correct but Statement II is incorrect.

Solution:

Step 1: Understanding the Question:

The question asks to evaluate two statements about the use and function of electrostatic precipitators, particularly in thermal power plants.

Step 2: Evaluating Statement I:

Statement I claims that the electrostatic precipitator is the most widely used device in thermal power plants. This is **true**. Thermal power plants burn coal, which produces a large amount of fly ash and other particulate matter. The electrostatic precipitator is a highly efficient device (over 99% efficiency) for removing these suspended particulate matter (SPM) from the exhaust gas before it is released into the atmosphere. Due to its high efficiency, it is a standard and essential component of such plants.

Step 3: Evaluating Statement II:

Statement II claims that the electrostatic precipitator removes ionizing radiations. This is **false**. The function of an electrostatic precipitator is to remove **particulate matter** (like dust, fly ash, soot) from a gas stream. It works by charging the particles using a corona discharge and then collecting them on oppositely charged plates. It has no mechanism to remove ionizing radiation (like gamma rays, X-rays, or alpha/beta particles). Ionizing radiation is a concern for nuclear power plants, not primarily thermal power plants, and requires different shielding methods (like lead or concrete).

Step 4: Final Answer:

Statement I is correct, but Statement II is incorrect. Therefore, the correct option is (4).

Quick Tip

Associate "Electrostatic Precipitator" with "Particulate Matter" or "SPM" removal. Think of it as an "electric dust magnet" for industrial smoke. It does not deal with gases or radiation. For removing gaseous pollutants like SO_2 , a "scrubber" is used.

182. Given below are two statements:

Statement I: In prokaryotes, the positively charged DNA is held with some negatively charged proteins in a region called nucleoid.

Statement II: In eukaryotes, the negatively charged DNA is wrapped around the positively charged histone octamer to form nucleosome.

In the light of the above statements, choose the correct answer from the options given below:

- (A) Statement I incorrect but Statement II is true.
- (B) Both Statement I and Statement II are true.
- (C) Both Statement I and Statement II are false.
- (D) Statement I is correct but Statement II is false.

Correct Answer: (1) Statement I incorrect but Statement II is true.

Solution:

Step 1: Understanding the Question:

The question presents two statements about the organization of DNA in prokaryotic and eukaryotic cells, focusing on the charges of DNA and associated proteins.

Step 2: Evaluating Statement I:

Statement I describes DNA organization in prokaryotes. It claims that positively charged DNA is held with negatively charged proteins. This is **incorrect** for two reasons. First, DNA is **negatively charged** due to the phosphate groups in its backbone. Second, the proteins associated with DNA in the nucleoid (nucleoid-associated proteins, NAPs) are generally believed to be **positively charged** (or have positive domains) to be able to bind to the negatively charged DNA. Therefore, the statement has the charges reversed.

Step 3: Evaluating Statement II:

Statement II describes DNA organization in eukaryotes. It states that negatively charged DNA is wrapped around a positively charged histone octamer to form a nucleosome. This is **true**. Eukaryotic DNA, which is negatively charged, is highly organized by wrapping around a core of eight histone proteins (a histone octamer). Histones are rich in positively charged amino acids (lysine and arginine), which allows them to bind tightly to the negatively charged DNA. This fundamental unit of DNA packaging is called a nucleosome.

Step 4: Final Answer:

Statement I is incorrect, and Statement II is correct. Therefore, the correct option is (1).

Quick Tip

A fundamental fact to remember: **DNA** is always negatively charged due to its phosphate (PO_4^{3-}) backbone. Therefore, any protein that needs to bind directly and tightly to it (like histones) must be **positively charged**.

183. Which of the following statements is correct?

- (A) Algal Bloom decreases fish mortality
- (B) Eutrophication refers to increase in domestic sewage and waste water in lakes.
- (C) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.
- (D) Presence of large amount of nutrients in water restricts 'Algal Bloom'

Correct Answer: (3) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statement among the four options related to water pollution and ecology.

Step 2: Detailed Explanation:

Let's analyze each statement:

- (1) Algal Bloom decreases fish mortality: This is incorrect. Algal blooms lead to a massive increase in dead organic matter. The decomposition of this matter by bacteria consumes large amounts of dissolved oxygen in the water, leading to hypoxia or anoxia and causing a massive increase in fish mortality.
- (2) Eutrophication refers to increase in domestic sewage and waste water in lakes.: This is an imprecise and thus incorrect definition. Eutrophication is the *natural* aging of a lake by nutrient enrichment of its water. While domestic sewage is a major cause of cultural (accelerated) eutrophication, the term itself refers to the nutrient enrichment (primarily with nitrates and phosphates) and its consequences, not the sewage itself.
- (3) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.: This is the precise and correct definition of biomagnification (or bioamplification). It describes the process whereby certain non-biodegradable substances, such as DDT or mercury, become more concentrated in organisms at successively higher levels in a food chain.
- (4) Presence of large amount of nutrients in water restricts 'Algal Bloom': This is incorrect. The presence of large amounts of nutrients (especially nitrogen and phosphorus) is the primary cause that **promotes** or triggers an algal bloom, it does not restrict it.

Step 3: Final Answer:

The only correct statement is (3).

Quick Tip

Differentiate these key pollution terms:

- Eutrophication: Nutrient enrichment leading to ecosystem changes.
- Algal Bloom: A rapid increase in algae population, a consequence of eutrophication.
- BOD (Biochemical Oxygen Demand): A measure of organic pollution; high BOD means high pollution and low dissolved oxygen.
- Biomagnification: Increase in toxin concentration up the food chain.

184. Which of the following are NOT considered as the part of endomembrane system?

- A. Mitochondria
- B. Endoplasmic Reticulum
- C. Chloroplasts
- D. Golgi complex
- E. Peroxisomes

Choose the most appropriate answer from the options given below:

- (A) A, D and E only
- (B) B and D only
- (C) A, C and E only
- (D) A and D only

Correct Answer: (3) A, C and E only

Solution:

Step 1: Understanding the Question:

The question asks to identify which of the listed cell organelles are not part of the endomembrane system.

Step 2: Detailed Explanation:

The endomembrane system is a group of membranes and organelles in eukaryotic cells that work together via vesicles to modify, package, and transport lipids and proteins. The components of this system are:

- 1. Endoplasmic Reticulum (ER)
- 2. Golgi complex (or Golgi apparatus)
- 3. Lysosomes
- 4. Vacuoles

These organelles are considered a single functional unit because their membranes are either directly connected or exchange material through vesicle transport.

Now let's evaluate the given organelles:

- A. Mitochondria: NOT part of the endomembrane system. It is a semi-autonomous organelle with its own DNA and ribosomes, and its function (cellular respiration) is not coordinated with the endomembrane system.
- B. Endoplasmic Reticulum: IS part of the endomembrane system.
- C. Chloroplasts: NOT part of the endomembrane system. Like mitochondria, it is a semi-autonomous organelle involved in photosynthesis.
- D. Golgi complex: IS part of the endomembrane system.

• E. Peroxisomes: NOT part of the endomembrane system. They are involved in metabolic processes like breaking down fatty acids and detoxifying harmful substances, but their functions are not integrated with the ER-Golgi pathway.

Step 3: Final Answer:

The organelles that are not part of the endomembrane system are Mitochondria (A), Chloroplasts (C), and Peroxisomes (E).

Quick Tip

Remember the core components of the endomembrane system: ER, Golgi, Lysosomes, and Vacuoles. A simple mnemonic could be "Very Large Green Elephants" (Vacuoles, Lysosomes, Golgi, ER). Organelles like Mitochondria, Chloroplasts, and Peroxisomes are functionally distinct and not included.

185. Match List I with List II.

List I (Cells) List II (Secretion)

- A. Peptic cells I. Mucus
- B. Goblet cells II. Bile juice
- C. Oxyntic cells III. Proenzyme pepsinogen
- D. Hepatic cells IV. HCl and intrinsic factor for absorption of vitamin B₁2

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-I, D-III
- (B) A-IV, B-III, C-II, D-I
- (C) A-II, B-I, C-III, D-IV
- (D) A-III, B-I, C-IV, D-II

Correct Answer: (4) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:

The question requires matching different types of cells from the digestive system (List I) with their respective secretions (List II).

Step 2: Detailed Explanation:

Let's analyze each cell type from List I:

• A. Peptic cells: Also known as chief cells or zymogen cells, these are found in the gastric glands of the stomach. They secrete the inactive enzyme precursor **Proenzyme pepsinogen**. So, A matches with III.

- B. Goblet cells: These are specialized mucus-secreting cells found interspersed in the epithelial lining of many organs, including the intestines and respiratory tracts. Their primary secretion is Mucus, which serves a protective function. So, B matches with I.
- C. Oxyntic cells: Also known as parietal cells, these are also found in the gastric glands of the stomach. They are responsible for secreting Hydrochloric acid (HCl) and intrinsic factor. The intrinsic factor is essential for the absorption of vitamin B₁₂. So, C matches with IV.
- D. Hepatic cells (Hepatocytes): These are the main cells of the liver. One of their many functions is to produce and secrete **Bile juice**, which is stored in the gallbladder and aids in fat digestion. So, **D matches with II**.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow III$

 $B \rightarrow I$

 $C \to IV$

 $\mathrm{D} \to \mathrm{II}$

This corresponds to option (4).

Quick Tip

Memorize the three main cell types of gastric glands and their secretions:

- Mucus neck cells: Secrete mucus.
- Peptic/Chief cells: Secrete pepsinogen.
- Oxyntic/Parietal cells: Secrete HCl and Intrinsic Factor.

Also remember, Goblet cells are the primary mucus producers in the intestine.

186. Which one of the following is NOT an advantage of inbreeding?

- (A) It decreases the productivity of inbred population, after continuous inbreeding.
- (B) It decreases homozygosity.
- (C) It exposes harmful recessive genes that are eliminated by selection.
- (D) Elimination of less desirable genes and accumulation of superior genes takes place due to it.

Correct Answer: (1) It decreases the productivity of inbred population, after continuous inbreeding.

Solution:

Step 1: Understanding the Question:

The question asks to identify the statement that is NOT an advantage of inbreeding. Note that option (A) describes a disadvantage, while the question asks for something that is NOT an advantage. This could be a disadvantage or simply an incorrect statement about the effects of inbreeding.

Step 2: Detailed Explanation:

Inbreeding is the mating of closely related individuals. Let's analyze its effects as described in the options.

- (1) It decreases the productivity of inbred population, after continuous inbreeding. This phenomenon is known as inbreeding depression. It is a major disadvantage of inbreeding, not an advantage.
- (2) It decreases homozygosity. This statement is factually incorrect. Inbreeding, by its very nature (mating between related individuals who are more likely to share the same alleles), increases homozygosity. It makes it more likely for offspring to receive identical alleles from both parents, thus increasing the frequency of homozygous genotypes (both dominant and recessive). Since this statement is false, it cannot be an advantage.
- (3) It exposes harmful recessive genes that are eliminated by selection. This is a key advantage of a controlled inbreeding program. By increasing homozygosity, it brings harmful recessive alleles together in homozygous individuals, who can then be identified and culled (removed) from the breeding population.
- (4) Elimination of less desirable genes and accumulation of superior genes takes place due to it. This is another major advantage. Through selection of superior individuals and elimination of undesirable ones (as exposed by inbreeding), this process helps to create a pure line or breed with desired characteristics.

Step 3: Final Answer:

The question asks what is NOT an advantage.

- Option (1) is a disadvantage.
- Option (2) is a false statement.
- Options (3) and (4) are advantages.

Quick Tip

Remember the dual nature of inbreeding:

- Advantage: Increases homozygosity, helps create pure lines, and exposes bad recessive alleles for selection.
- **Disadvantage:** Can lead to inbreeding depression (loss of vigor and productivity) if harmful recessive alleles become fixed.

187. In cockroach, excretion is brought about by-

- A. Phallic gland
- B. Urecose gland
- C. Nephrocytes
- D. Fat body
- E. Collaterial glands

Choose the correct answer from the options given below:

- (A) B and D only
- (B) A and E only
- (C) A, B and E only
- (D) B, C and D only

Correct Answer: (4) B, C and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the structures involved in the process of excretion in cockroaches from the given list.

Step 2: Detailed Explanation:

The primary excretory organs in cockroaches are the Malpighian tubules, which are not listed. However, several other structures also play a role in excretion. Let's analyze the listed options:

- A. Phallic gland: This is a part of the male reproductive system and is involved in forming the outer layer of the spermatophore. It has no excretory function.
- B. Urecose glands: These glands are found in some species of cockroaches. They are associated with the male reproductive system but are primarily involved in the storage and excretion of uric acid. They are considered excretory in function.
- C. Nephrocytes: These are specialized cells found in the body cavity (haemocoel) that are thought to absorb and process nitrogenous waste products from the hemolymph, playing a role analogous to parts of the vertebrate kidney. They are excretory.
- **D. Fat body:** This is a large, diffuse organ in the cockroach's body cavity. In addition to storing nutrients, the cells of the fat body (trophocytes and urate cells) are involved in the synthesis and storage of uric acid, which is the main nitrogenous waste. It is considered an excretory organ.
- E. Collaterial glands: These are part of the female reproductive system. Their secretion forms the hard protective casing (ootheca) around the eggs. They have no excretory function.

Step 3: Final Answer:

The structures involved in excretion among the given options are the Urecose glands (B), Nephrocytes (C), and the Fat body (D).

Quick Tip

In cockroaches, remember that excretion is not limited to the Malpighian tubules. The fat body, nephrocytes, and urecose glands also play important roles. Differentiate these from reproductive structures like the phallic gland (male) and collaterial glands (female).

188. Which one of the following is the sequence on corresponding coding strand, if the sequence on mRNA formed is as follows

- 5' AUCGAUCGAUCGAUCG AUCG AUCG 3'?
- (A) 3' ATCGATCGATCGATCG ATCGATCG 5'
- (B) 5' UAGCUAGCUAGCUAGCUA GCUAGC UAGC 3'
- (C) 3' UAGCUAGCUAGCUAGCUA GCUAGCUAGC 5'
- (D) 5' ATCGATCGATCGATCG ATCGATCG 3'

Correct Answer: (4) 5' ATCGATCGATCGATCGATCG ATCGATCG 3'

Solution:

Step 1: Understanding the Question:

The question provides an mRNA sequence and asks for the sequence of the corresponding **coding strand** of the DNA.

Step 2: Key Formula or Approach:

Let's clarify the terminology:

- Template Strand (or anti-sense strand): The DNA strand that is read by RNA polymerase to synthesize mRNA. Its sequence is complementary to the mRNA sequence (with T instead of U).
- Coding Strand (or sense strand): The other DNA strand. Its sequence is not used as a template during transcription. Its sequence is identical to the mRNA sequence, with the only difference being that Thymine (T) is present in DNA instead of Uracil (U) in RNA. It has the same 5' to 3' polarity as the mRNA.

Step 3: Detailed Explanation:

The given mRNA sequence is:

mRNA: 5' - AUCGAUCGAUCGAUCG AUCG AUCG - 3'

The coding strand has the same sequence and polarity as the mRNA, except that U is replaced by T. So, we simply need to take the mRNA sequence and substitute every 'U' with a 'T'.

mRNA: 5' - A U C G A U C G ... - 3'

Replace U with T:

Coding Strand: 5' - A T C G A T C G ... - 3'

Applying this to the full sequence:

mRNA: 5' AUCGAUCGAUCGAUCGAUCG AUCG 3' Coding Strand: 5' ATCGATCGATCGATCGATCG ATCGATCG 3'

Step 4: Evaluating the Options:

- (1) 3' ATCG... 5': This has the correct sequence but the wrong polarity. This would be the template strand.
- (2) 5' UAGCU... 3': This sequence contains U, so it's an RNA sequence, not DNA.
- (3) 3' UAGCU... 5': This is also an RNA sequence with the wrong polarity.
- (4) 5' ATCGATCGATCGATCGATCG ATCGATCG 3': This has the correct sequence (T instead of U) and the correct 5' to 3' polarity. This is the correct coding strand.

Quick Tip

Remember: Coding strand = mRNA strand (with T instead of U). They both run in the same direction (5' to 3'). The template strand is complementary and runs in the opposite direction for reading (3' to 5').

189. Match List I with List II.

List I List II

- A. Logistic growth I. Unlimited resource availability condition
- B. Exponential growth II. Limited resource availability condition
- C. Expanding age pyramid III. The percent individuals of pre-reproductive age is largest followed by reproductive and post reproductive age groups
- D. Stable age pyramid IV. The percent individuals of pre-reproductives and reproductive age group are same

Choose the correct answer from the options given below:

- (A) A-II, B-IV, C-III, D-I
- (B) A-II, B-I, C-III, D-IV
- (C) A-II, B-III, C-I, D-IV
- (D) A-II, B-IV, C-I, D-III

Correct Answer: (2) A-II, B-I, C-III, D-IV

Solution:

Step 1: Understanding the Question:

The question requires matching ecological concepts (growth models and age pyramids) from List I with their correct descriptions from List II.

Step 2: Detailed Explanation:

Let's analyze each item in List I:

- A. Logistic growth: This model describes population growth in an environment with finite resources. As the population approaches the carrying capacity (K), its growth rate slows down. This corresponds to a Limited resource availability condition. The resulting growth curve is S-shaped. So, A matches with II.
- B. Exponential growth: This model describes population growth under ideal conditions, where resources are not a limiting factor. The rate of growth is proportional to the size of the population. This corresponds to an Unlimited resource availability condition. The resulting growth curve is J-shaped. So, B matches with I.
- C. Expanding age pyramid: This is a triangular-shaped pyramid with a broad base and a narrow top. It indicates that the percentage of **pre-reproductive individuals** is the largest, followed by the reproductive and then the post-reproductive age groups. This signifies a rapidly growing population. So, C matches with III.
- **D. Stable age pyramid:** This is a bell-shaped pyramid. It indicates that the percentage of **pre-reproductive and reproductive individuals is roughly the same** or similar, with the post-reproductive group being smaller. This signifies a population that is not growing or growing very slowly (zero or near-zero growth). The description in List II (IV) states the pre-reproductive and reproductive groups are the same, which corresponds to a stable pyramid. So, **D matches with IV**.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to I$

 $C \to III$

 $D \to IV$

This corresponds to option (2).

Quick Tip

Associate growth curves with resource availability:

- Exponential (J-shaped): Unlimited resources, "ideal" world.
- Logistic (S-shaped): Limited resources, "realistic" world with carrying capacity (K).

Associate age pyramid shapes with population status:

- Triangle (Expanding): Broad base (many young).
- Bell (Stable): Base and middle are similar.
- Urn (Declining): Narrow base (few young).

190. Match List I with List II.

List I

List II

- A. Mast cells
- I. Ciliated epithelium
- B. Inner surface of bronchiole
- II. Areolar connective tissue

C. Blood

- III. Cuboidal epithelium
- D. Tubular parts of nephron
- IV. specialised connective tissue

Choose the correct answer from the options give below:

- (A) A-III, B-IV, C-II, D-I
- (B) A-I, B-II, C-IV, D-III
- (C) A-II, B-III, C-I, D-IV
- (D) A-II, B-I, C-IV, D-III

Correct Answer: (4) A-II, B-I, C-IV, D-III

Solution:

Step 1: Understanding the Question:

The question requires matching structures or cells from List I with the type of tissue they are or are composed of from List II.

Step 2: Detailed Explanation:

Let's analyze each item in List I:

• A. Mast cells: These are resident cells of several types of tissues and contain granules rich in histamine and heparin. They are found in connective tissue, specifically Areolar connective tissue, where they are involved in inflammatory and allergic reactions. So, A matches with II.

- B. Inner surface of bronchiole: The smaller bronchioles are lined by Ciliated epithelium (specifically, ciliated cuboidal or columnar epithelium). The cilia help to move mucus and trapped particles out of the respiratory tract. So, B matches with I.
- C. Blood: Blood is considered a fluid specialised connective tissue. It consists of cells (RBCs, WBCs, platelets) suspended in a fluid matrix (plasma). So, C matches with IV.
- D. Tubular parts of nephron: The different parts of the nephron's tubule are lined by different types of epithelium. The proximal convoluted tubule (PCT) and distal convoluted tubule (DCT) are lined by simple Cuboidal epithelium. The PCT has a brush border of microvilli. So, D matches with III.

Step 3: Final Matching:

The correct matching is:

 $A \rightarrow II$

 $B \to I$

 $C \to IV$

 $\mathrm{D} \to \mathrm{III}$

This corresponds to option (4).

Quick Tip

For histology questions, create a mental map or a table linking key locations to their specific tissue types.

- Ciliated Epithelium: Respiratory tract (bronchioles, trachea).
- Cuboidal Epithelium: Gland ducts, Tubular parts of nephrons.
- Specialized Connective Tissue: Blood, Bone, Cartilage.
- Areolar Connective Tissue: Found beneath epithelia, contains mast cells, macrophages, fibroblasts.

191. Given below are two statements:

Statement I: During G_0 phase of cell cycle, the cell is metabolically inactive. Statement II: The centrosome undergoes duplication during S phase of interphase. In the light of the above statements, choose the most appropriate answer from the options given below:

- (A) Statement I is incorrect but Statement II is correct.
- (B) Both Statement I and Statement II are correct.
- (C) Both Statement I and Statement II are incorrect.

(D) Statement I is correct but Statement II is incorrect.

Correct Answer: (1) Statement I is incorrect but Statement II is correct.

Solution:

Step 1: Understanding the Question:

The question asks to evaluate two statements about events in the cell cycle.

Step 2: Evaluating Statement I:

Statement I claims that a cell in the G_0 phase is metabolically inactive. This is incorrect. The G_0 phase, or quiescent stage, is a non-dividing state that cells enter from the G_1 phase. While the cell has exited the proliferative cycle, it remains **metabolically active**. It performs its specialized functions, grows in size, and carries out all necessary metabolic processes to live. It is only inactive with respect to cell division.

Step 3: Evaluating Statement II:

Statement II claims that the centrosome duplicates during the S phase of interphase. This is **correct**. The S phase (Synthesis phase) is primarily known for DNA replication. However, another crucial event that occurs during the S phase in animal cells is the duplication of the centrosome. This ensures that each daughter cell will receive a centrosome to organize its microtubules during the subsequent mitosis.

Step 4: Final Answer:

Statement I is incorrect, while Statement II is correct. Therefore, the correct option is (1).

Quick Tip

Remember the key events of the cell cycle phases:

- **G**₁: Cell growth, metabolic activity.
- S: DNA replication and centrosome duplication.
- **G**₂: Protein synthesis (e.g., tubulin), preparation for mitosis.
- M: Mitosis (nuclear division) and cytokinesis.
- G_0 : Metabolically active but non-proliferating state.

192. Which of the following statements are correct?

- A. Basophils are most abundant cells of the total WBCs
- B. Basophils secrete histamine, serotonin and heparin
- C. Basophils are involved in inflammatory response
- D. Basophils have kidney shaped nucleus
- E. Basophils are agranulocytes

Choose the correct answer from the options given below:

- (A) A and B only
- (B) D and E only
- (C) C and E only
- (D) B and C only

Correct Answer: (4) B and C only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements about basophils from a given list.

Step 2: Detailed Explanation:

Let's evaluate each statement about basophils, a type of white blood cell (WBC):

- Statement A: Basophils are most abundant cells of the total WBCs. This is incorrect. Basophils are the *least* abundant WBCs, making up only about 0.5-1% of the total count. The most abundant WBCs are neutrophils (60-65%).
- Statement B: Basophils secrete histamine, serotonin and heparin. This is correct. The granules of basophils contain these chemicals, which are mediators of inflammation.
- Statement C: Basophils are involved in inflammatory response. This is correct. By releasing histamine, serotonin, and heparin, basophils play a key role in initiating and mediating inflammatory and allergic reactions.
- Statement D: Basophils have kidney shaped nucleus. This is incorrect. Basophils typically have a bilobed or S-shaped nucleus, which is often obscured by their large, coarse granules. A kidney-shaped nucleus is characteristic of monocytes.
- Statement E: Basophils are agranulocytes. This is incorrect. Basophils are classified as granulocytes, along with neutrophils and eosinophils, because their cytoplasm contains prominent granules. The agranulocytes are lymphocytes and monocytes.

Step 3: Final Answer:

The only correct statements are B and C. Therefore, the correct option is (4).

Quick Tip

Remember the WBC abundance order with the mnemonic "Never Let Monkeys Eat Bananas": Neutrophils ¿ Lymphocytes ¿ Monocytes ¿ Eosinophils ¿ Basophils. Also, classify them:

- Granulocytes: Neutrophils, Eosinophils, Basophils (the "-phils").
- Agranulocytes: Lymphocytes, Monocytes (the "-cytes").

193. Which of the following statements are correct?

- A. An excessive loss of body fluid from the body switches off osmoreceptors.
- B. ADH facilitates water reabsorption to prevent diuresis.
- C. ANF causes vasodilation.
- D. ADH causes increase in blood pressure.
- E. ADH is responsible for decrease in GFR.

Choose the correct answer from the options given below:

- (A) C, D and E only
- (B) A and B only
- (C) B, C and D only
- (D) A, B and E only

Correct Answer: (3) B, C and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements related to the hormonal regulation of kidney function and blood pressure.

Step 2: Detailed Explanation:

Let's evaluate each statement:

- Statement A: An excessive loss of body fluid from the body switches off osmore-ceptors. This is incorrect. Excessive fluid loss leads to dehydration, which increases the osmolarity (solute concentration) of the blood. This change switches on or stimulates the osmoreceptors in the hypothalamus, leading to the release of ADH.
- Statement B: ADH facilitates water reabsorption to prevent diuresis. This is correct. Antidiuretic Hormone (ADH) increases the permeability of the distal convoluted tubules and collecting ducts to water, promoting water reabsorption from the filtrate back into the blood. This conserves water and prevents diuresis (excessive urine production).
- Statement C: ANF causes vasodilation. This is correct. Atrial Natriuretic Factor (ANF) is released by the heart atria in response to high blood pressure. It acts as a vasodilator (widens blood vessels), which helps to decrease blood pressure.
- Statement D: ADH causes increase in blood pressure. This is correct. ADH, also known as vasopressin, has a vasoconstrictor effect on arterioles at high concentrations, which increases peripheral resistance and thus raises blood pressure. Its water reabsorption effect also increases blood volume, contributing to increased blood pressure.
- Statement E: ADH is responsible for decrease in GFR. This is incorrect. ADH's primary role is on water reabsorption in the later parts of the nephron. While its vasoconstrictor effect could potentially affect the Glomerular Filtration Rate (GFR), the renin-

angiotensin-aldosterone system has a more direct and potent effect. In general, mechanisms that raise blood pressure (like ADH) tend to maintain or increase GFR, not decrease it.

Step 3: Final Answer:

The correct statements are B, C, and D. Therefore, the correct option is (3).

Quick Tip

Remember the opposing actions of ADH/RAAS and ANF on blood pressure:

- ADH and RAAS (Renin-Angiotensin-Aldosterone System): Work to *increase* blood pressure by vasoconstriction and increasing blood volume (water/salt retention).
- ANF (Atrial Natriuretic Factor): Works to *decrease* blood pressure by vasodilation and decreasing blood volume (salt/water excretion).

194. Select the correct statements with reference to chordates.

- A. Presence of a mid-dorsal, solid and double nerve cord.
- B. Presence of closed circulatory system.
- C. Presence of paired pharyngeal gillslits.
- D. Presence of dorsal heart
- E. Triploblastic pseudocoelomate animals.

Choose the correct answer from the options given below:

- (A) C, D and E only
- (B) A, C and D only
- (C) B and C only
- (D) B, D and E only

Correct Answer: (3) B and C only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements that describe the characteristics of the phylum Chordata.

Step 2: Detailed Explanation:

The three fundamental characteristics of all chordates, present at some stage of their life, are:
1. Presence of a notochord. 2. Presence of a dorsal, hollow nerve cord. 3. Presence of paired pharyngeal gill slits.

Let's evaluate the given statements based on these and other general chordate features:

- Statement A: Presence of a mid-dorsal, solid and double nerve cord. This is incorrect. The chordate nerve cord is dorsal, but it is hollow and single, not solid and double. A solid, double, ventral nerve cord is characteristic of non-chordates like annelids and arthropods.
- Statement B: Presence of closed circulatory system. This is a correct feature of most chordates (especially vertebrates).
- Statement C: Presence of paired pharyngeal gillslits. This is a correct and fundamental characteristic of all chordates, present at least in their embryonic stage.
- Statement D: Presence of dorsal heart. This is incorrect. Chordates have a ventral heart. A dorsal heart is found in many non-chordates.
- Statement E: Triploblastic pseudocoelomate animals. This is incorrect. Chordates are triploblastic, but they are coelomates (possess a true coelom), not pseudocoelomates. Pseudocoelomates include phyla like Aschelminthes (roundworms).

Step 3: Final Answer:

The only correct statements are B and C. Therefore, the correct option is (3).

Quick Tip	
To differentiate chordates from	non-chordates, remember this table of contrasts:
Chordates	Non-Chordates
Notochord present	Notochord absent
Dorsal, hollow, single nerve cord	Ventral, solid, double nerve cord
Pharyngeal gill slits present	Pharyngeal gill slits absent
Ventral heart	Dorsal heart (if present)
Post-anal tail present	Post-anal tail absent

195. Select the correct statements.

- A. Tetrad formation is seen during Leptotene.
- B. During Anaphase, the centromeres split and chromatids separate.
- C. Terminalization takes place during Pachytene.
- D. Nucleolus, Golgi complex and ER are reformed during Telophase.
- E. Crossing over takes place between sister chromatids of homologous chromosome.

Choose the correct answer from the options given below:

- (A) B and E only
- (B) A and C only
- (C) B and D only
- (D) A, C and E only

Correct Answer: (3) B and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements describing events of cell division (mitosis and meiosis).

Step 2: Detailed Explanation:

Let's evaluate each statement:

- Statement A: Tetrad formation is seen during Leptotene. This is incorrect. Tetrad formation (synapsis of homologous chromosomes to form bivalents, which are also called tetrads because they consist of four chromatids) occurs during the **Zygotene** stage of Prophase I. Leptotene is characterized by the condensation of chromatin into chromosomes.
- Statement B: During Anaphase, the centromeres split and chromatids separate. This is a **correct** description of Anaphase in **mitosis** and Anaphase II in **meiosis**. The splitting of centromeres allows the sister chromatids to separate and move to opposite poles.
- Statement C: Terminalization takes place during Pachytene. This is incorrect. Terminalization of chiasmata (the movement of chiasmata towards the ends of the chromosomes) begins in **Diplotene** and is completed in **Diakinesis** of Prophase I. Pachytene is the stage where crossing over occurs.
- Statement D: Nucleolus, Golgi complex and ER are reformed during Telophase. This is correct. Telophase is essentially the reverse of prophase. The nuclear envelope reforms around the chromosome clusters at each pole, and the nucleolus, Golgi complex, and ER also reappear.
- Statement E: Crossing over takes place between sister chromatids of homologous chromosome. This is incorrect. Crossing over, the exchange of genetic material, occurs between non-sister chromatids of a homologous chromosome pair. Crossing over between sister chromatids would result in no genetic recombination as they are genetically identical.

Step 3: Final Answer:

The only correct statements are B and D. Therefore, the correct option is (3).

Quick Tip

Remember the sequence and key events of Prophase I of Meiosis:

• Leptotene: Condensation.

• Zygotene: Synapsis (pairing), bivalent/tetrad formation.

• Pachytene: Crossing over.

• Diplotene: Chiasmata visible, dissolution of synaptonemal complex.

• Diakinesis: Terminalization.

Mnemonic: "Lazy Zebra Punches Double Deer".

196. The parts of human brain that helps in regulation of sexual behaviour, expression of excitement, pleasure, rage, fear etc. are:

- (A) Corpus callosum and thalamus
- (B) Limbic system & hypothalamus
- (C) Corpora quadrigemina & hippocampus
- (D) Brain stem & epithalamus

Correct Answer: (2) Limbic system & hypothalamus

Solution:

Step 1: Understanding the Question:

The question asks to identify the parts of the human brain responsible for regulating emotions, motivations, and drives like sexual behavior, fear, and pleasure.

Step 2: Detailed Explanation of Brain Functions:

Let's analyze the functions of the structures in the options:

- (1) Corpus callosum and thalamus: The corpus callosum is a large bundle of nerve fibers connecting the two cerebral hemispheres. The thalamus is a major relay center for sensory information. They are not the primary centers for emotion and sexual behavior.
- (2) Limbic system & hypothalamus: The limbic system is a complex set of brain structures that includes the amygdala, hippocampus, and parts of the thalamus and hypothalamus. It is deeply involved in motivation, emotion, learning, and memory. The hypothalamus, located just below the thalamus, is a key control center for the autonomic nervous system and the endocrine system. Together, the limbic system and the hypothalamus are responsible for regulating emotional reactions (like rage and fear), motivational drives (like hunger, thirst, and sexual behavior), and feelings of pleasure. This pair is the correct answer.

- (3) Corpora quadrigemina & hippocampus: The corpora quadrigemina (in the midbrain) are reflex centers for vision and hearing. The hippocampus is part of the limbic system and is crucial for memory formation, but this option is incomplete.
- (4) Brain stem & epithalamus: The brain stem (midbrain, pons, medulla oblongata) controls vital life functions like breathing and heartbeat. The epithalamus contains the pineal gland, which regulates sleep-wake cycles.

Step 3: Final Answer:

The regulation of complex behaviors and emotions listed in the question is a primary function of the limbic system and the hypothalamus.

Quick Tip

Associate the **Limbic System** with the "three M's": **Motivation**, e**Motion**, and **Memory**. The **Hypothalamus** is the master regulator of homeostasis and basic drives. Together they form the emotional core of the brain.

197. The unique mammalian characteristics are:

- (A) pinna, monocondylic skull and mammary glands
- (B) hairs, tympanic membrane and mammary glands
- (C) hairs, pinna and mammary glands
- (D) hairs, pinna and indirect development

Correct Answer: (3) hairs, pinna and mammary glands

Solution:

Step 1: Understanding the Question:

The question asks to identify the set of characteristics that are unique to mammals.

Step 2: Detailed Explanation:

Let's analyze the characteristics listed in each option:

- (1) pinna, monocondylic skull and mammary glands:
 - Pinna (external ear): Characteristic of most mammals.
 - Monocondylic skull: This means having one occipital condyle. This is a feature of reptiles and birds. Mammals have a dicondylic skull (two occipital condyles). So, this is incorrect.
 - Mammary glands: A unique, defining feature of mammals.
- (2) hairs, tympanic membrane and mammary glands:

- Hairs: A unique, defining feature of mammals.
- Tympanic membrane (eardrum): Present in many other vertebrates like frogs, reptiles, and birds. Not unique to mammals.
- Mammary glands: Unique to mammals.
- (3) hairs, pinna and mammary glands:
 - Hairs (fur or pelage): Presence of hair is a uniquely mammalian trait.
 - Pinna (external ear): Presence of a fleshy external ear is characteristic of mammals (though absent in some aquatic ones).
 - Mammary glands: Glands that produce milk to nourish the young are the most defining feature of the class Mammalia.

This set contains characteristics that are broadly considered unique to mammals.

- (4) hairs, pinna and indirect development:
 - Hairs and Pinna: Mammalian characteristics.
 - Indirect development: This involves a larval stage (e.g., tadpole in frogs). Mammals exhibit direct development, where the young are born as miniature versions of the adult. So, this is incorrect.

Step 3: Final Answer:

The combination of hairs, pinna, and mammary glands represents the most accurate set of unique mammalian characteristics among the choices provided.

Quick Tip

The three most defining and unique characteristics of mammals are: 1. **Mammary glands** (for milk production). 2. **Hair** on the body. 3. **Three middle ear ossicles** (malleus, incus, stapes). Also important are the dicondylic skull and the muscular diaphragm. The pinna is also a good characteristic feature.

198. Which of the following is characteristic feature of cockroach regarding sexual dimorphism?

- (A) Presence of anal cerci
- (B) Dark brown body colour and anal cerci
- (C) Presence of anal styles
- (D) Presence of sclerites

Correct Answer: (3) Presence of anal styles

Solution:

Step 1: Understanding the Question:

The question asks to identify the feature that distinguishes male and female cockroaches (sexual dimorphism).

Step 2: Detailed Explanation:

Let's analyze the features listed:

- (1) Presence of anal cerci: Anal cerci are a pair of jointed, filamentous structures found on the 10th abdominal segment. They are sensory in function. Importantly, they are present in **both males and females**. Therefore, they are not a feature of sexual dimorphism.
- (2) Dark brown body colour and anal cerci: Body color is generally similar in both sexes. As mentioned, anal cerci are present in both.
- (3) Presence of anal styles: Anal styles are a pair of short, unjointed, thread-like structures that arise from the 9th abdominal sternum. These are present only in male cockroaches. Females lack anal styles. Therefore, the presence of anal styles is a key characteristic feature for identifying male cockroaches and a clear example of sexual dimorphism.
- (4) Presence of sclerites: Sclerites are the hardened chitinous plates that make up the exoskeleton of the cockroach. They are present in **both males and females**.

Step 3: Final Answer:

The presence of anal styles exclusively in males is the characteristic feature of sexual dimorphism in cockroaches among the options given.

Quick Tip

Remember the key difference in the posterior abdomen of cockroaches:

- Both sexes have a pair of anal cerci on the 10th segment.
- Only males have a pair of anal styles on the 9th sternum.

Think "Style for the guys" to remember that styles are only in males.

199. Which of the following statements are correct regarding skeletal muscle?

- A. Muscle bundles are held together by collagenous connective tissue layer called fascicle.
- B. Sarcoplasmic reticulum of muscle fibre is a store house of calcium ions.
- C. Striated appearance of skeletal muscle fibre is due to distribution pattern of actin and myosin proteins.
- D. M line is considered as functional unit of contraction called sarcomere.

Choose the most appropriate answer from the options given below:

- (A) C and D only
- (B) A, B and C only
- (C) B and C only
- (D) A, C and D only

Correct Answer: (3) B and C only

Solution:

Step 1: Understanding the Question:

The question asks to identify the correct statements about the structure and function of skeletal muscle.

Step 2: Detailed Explanation:

Let's evaluate each statement:

- Statement A: Muscle bundles are held together by collagenous connective tissue layer called fascicle. This statement is slightly inaccurate in its wording. A fascicle (or muscle bundle) is a bundle of muscle fibers. The connective tissue layer that surrounds the entire muscle is the epimysium. The layer that surrounds each fascicle is the perimysium. The layer surrounding each muscle fiber is the endomysium. The statement says the layer is *called* fascicle, which is incorrect. The layer is called perimysium, and the bundle is called a fascicle. However, if we interpret "fascicle" as referring to the bundle itself, then it's held together by perimysium. Let's re-read carefully: "Muscle bundles are held together by ... layer called fascicle". This is wrong. The layer is called perimysium. Let's assume there is a typo and it should have said "called fascia" or "perimysium". In many contexts, this statement is considered generally true in its intent. Let's check other options.
- Statement B: Sarcoplasmic reticulum of muscle fibre is a store house of calcium ions. This is correct. The sarcoplasmic reticulum is the specialized endoplasmic reticulum of muscle cells. Its primary function is to sequester, store, and release calcium ions (Ca^{2+}) , which are the trigger for muscle contraction.
- Statement C: Striated appearance of skeletal muscle fibre is due to distribution pattern of actin and myosin proteins. This is correct. The characteristic striped or striated appearance of skeletal muscle is created by the highly organized, repeating arrangement of thick (myosin) and thin (actin) filaments within the myofibrils. This arrangement forms the pattern of light (I-bands) and dark (A-bands).
- Statement D: M line is considered as functional unit of contraction called sarcomere. This is incorrect. The functional unit of muscle contraction is the sarcomere, which is defined as the region of a myofibril between two successive Z lines. The M line is a line of proteins found in the middle of the A-band (and the sarcomere) that holds the thick filaments together.

Step 3: Final Answer:

Statements B and C are definitely correct.

Quick Tip

Memorize the structure of a sarcomere: "Z I A H M H A I Z"

- Sarcomere: From Z-line to Z-line. This is the functional unit.
- **Z-line:** Anchors thin filaments.
- I-band: Light band, only thin filaments (Act<u>I</u>n).
- **A-band:** Dark band, contains entire thick filament (Myosin) and overlapping thin filaments.
- **H-zone:** Center of A-band, only thick filaments.
- M-line: Middle of H-zone and sarcomere.

200. Which of the following are NOT under the control of thyroid hormone?

- A. Maintenance of water and electrolyte balance
- B. Regulation of basal metabolic rate
- C. Normal rhythm of sleep-wake cycle
- D. Development of immune system
- E. Support the process of R.B.Cs formation

Choose the correct answer from the options given below:

- (A) D and E only
- (B) A and D only
- (C) B and C only
- (D) C and D only

Correct Answer: (4) C and D only

Solution:

Step 1: Understanding the Question:

The question asks to identify the functions from the list that are NOT controlled by thyroid hormones (thyroxine T4 and triiodothyronine T3).

Step 2: Detailed Explanation of Thyroid Hormone Functions:

Thyroid hormones have a wide range of effects on the body. Let's evaluate each function listed:

- A. Maintenance of water and electrolyte balance: Thyroid hormones influence kidney function and can affect water and electrolyte balance. This is considered a function of thyroid hormone.
- B. Regulation of basal metabolic rate (BMR): This is one of the most important and well-known functions of thyroid hormones. They increase oxygen consumption and heat production by most tissues, thus regulating the BMR.

- C. Normal rhythm of sleep-wake cycle: The sleep-wake cycle (circadian rhythm) is primarily regulated by the hormone **melatonin**, which is secreted by the pineal gland. This is **NOT** a function of thyroid hormone.
- **D. Development of immune system:** The primary hormone responsible for the maturation and differentiation of T-lymphocytes, and thus for the development of the cell-mediated immune system, is **thymosin**, secreted by the thymus gland. This is **NOT** a function of thyroid hormone.
- E. Support the process of R.B.Cs formation: Thyroid hormones are necessary for normal erythropoiesis (RBC formation). They stimulate the production of erythropoietin. This is a function of thyroid hormone.

Step 3: Final Answer:

The functions not under the control of thyroid hormone are the normal rhythm of the sleep-wake cycle (C) and the development of the immune system (D).

Quick Tip

Remember the main regulators for specific functions:

- Thyroid: Metabolism (BMR).
- Melatonin (Pineal gland): Sleep-wake cycle.
- Thymosin (Thymus): T-cell maturation/Immunity.
- Aldosterone (Adrenal cortex): Electrolyte balance.

While there is some overlap, these are the primary associations to remember for exams.