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# Generative AI for Image Generation

COMP4431 Lab 2

# Lab 02 Agenda

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## ■ Diffusion Models

- Concepts of Diffusion Models
- Access to Diffusion Models

## • Text-to-image Generation

- Leaderboard of Text-to-image Generation Models
- Prompts in Text-to-image Generation Models
- Negative Prompts
- Keyword Weight Prompts

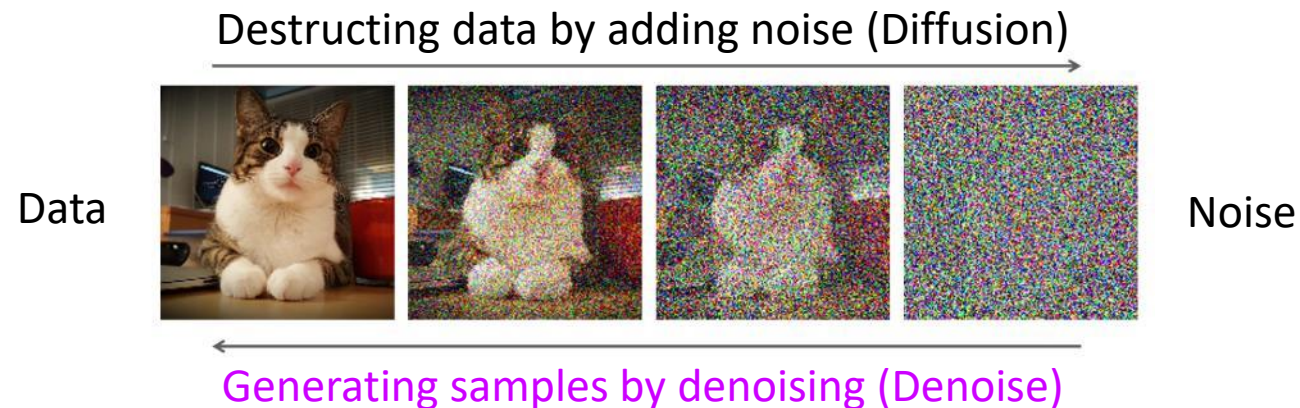
## ■ Image-to-image Generation

- Step-by-step Guide to Image-to-image Generation
- Step-by-step Guide to Inpainting

# Concepts of Diffusion Models

- Diffusion Models

- Diffusion models have emerged as a powerful new family of deep generative models with record-breaking performance.
- Diffusion models are a family of probabilistic generative models that progressively destruct data by injecting noise, then learn to reverse this process for sample generation.



Yang L, Zhang Z, Song Y, et al. Diffusion models: A comprehensive survey of methods and applications[J]. ACM Computing Surveys, 2023, 56(4): 1-39.

# Access to Diffusion Models (1)

- PolyU GenAI Diffusion Models
  - Navigate to <https://genai.polyu.edu.hk/>
  - Log in using your PolyU NetID and NetPassword
  - Select a diffusion model
    - Users can generate images based on monthly entitlement (1000).
    - One image costs around 2 credits for Stable Diffusion XL 1.0
    - One image costs around 65 credits for Stable Diffusion 3.0
  - Prompting the diffusion model



Sign in with your NetID and NetPassword

hh\22038275r

.....

☐ 使我保持登录状态

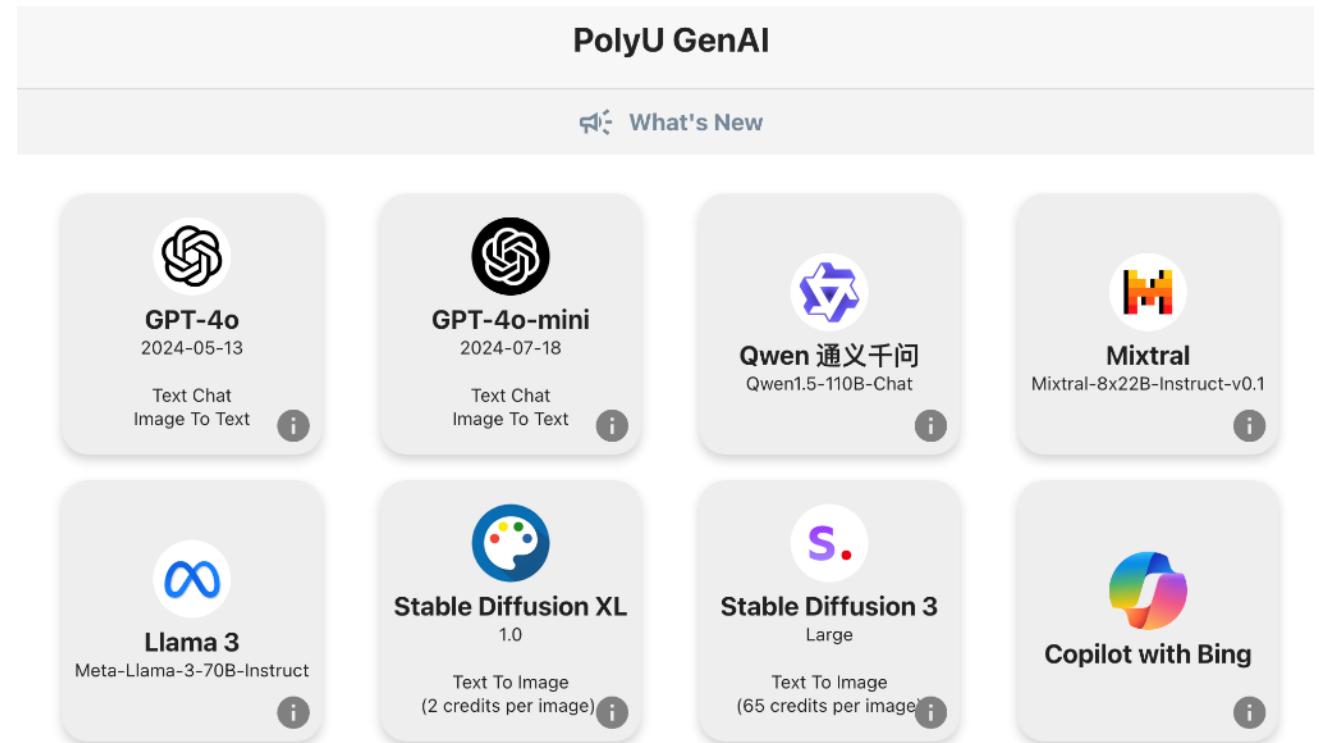
登录

[Forgot Your NetPassword?](#)

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dog, autumn in paris, ornate, beautiful, atmosphere, vibe, mist, smoke, fire, chimney, rain, wet, pristine, puddles, melting, dripping, snow, creek, lush, ice, bridge, forest, roses, flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.



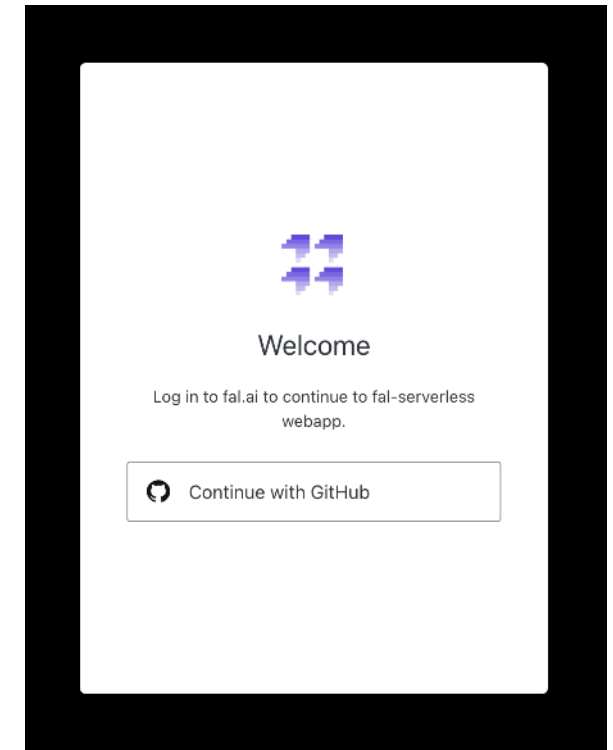
Seed Number: 1154168450

[Download](#)



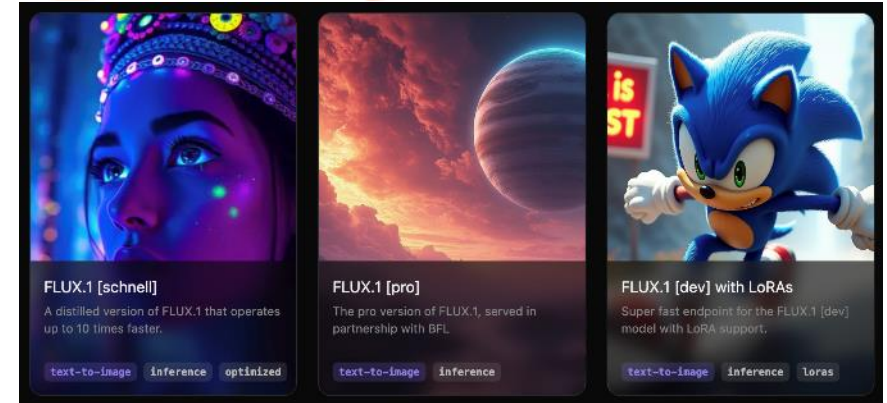
# Access to Diffusion Models (2)

- fal.ai Diffusion Models
  - Navigate to <https://fal.ai/models>
  - Log in using your GitHub username and password
  - Select a diffusion model
    - New users are given a current balance of \$1.00. Users can generate images based on their account balance.
    - \$10 allows approximately 4,500 images using Stable Diffusion XL
    - \$50 allows approximately 23,000 images using Stable Diffusion XL
  - Prompting the diffusion model



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MODEL NAME	UNIT PRICE (USD)
FLUX.1 [dev]	0.025 per megapixel
FLUX.1 [schnell]	0.003 per megapixel
FLUX.1 [pro]	0.05 per megapixel
Stable Diffusion 3 - Medium	0.035 per image

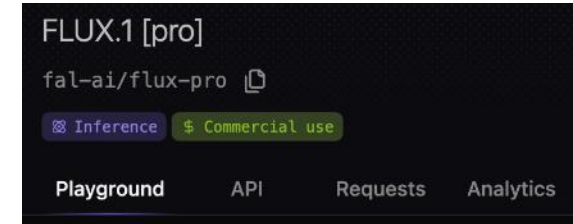


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Prompt\* ⓘ

dog, autumn in paris, ornate, beautiful, atmosphere, vibe, mist, smoke, fire, chimney, rain, wet, pristine, puddles, melting, dripping, snow, creek, lush, ice, bridge, forest, roses, flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.

















# Lab 02 Agenda

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- Diffusion Models
  - Concepts of Diffusion Models
  - Access to Diffusion Models
- Text-to-image Generation
  - Leaderboard of Text-to-image Generation Models
  - Prompts in Text-to-image Generation Models
  - Negative Prompts
  - Keyword Weight Prompts
- Image-to-image Generation
  - Step-by-step Guide to Image-to-image Generation
  - Step-by-step Guide to Inpainting

# Leaderboard of Text-to-image Generation Models

- FLUX. 1 [pro] and FLUX. 1 [dev] are the highest-quality models
  - Quality ELO: Relative score of the models as determined by >100,000 responses from users.
- Stable Diffusion XL Lightning and FLUX.1 [schnell] are the cheapest models.
  - Price: Price per 1k images generated by the model.
- Stable Diffusion XL and Stable Diffusion XL Lightning are the fastest models.
  - Generation Time: Median time the provider takes to generate an image

PROVIDER	MODEL	FOOTNOTES	MODEL QUALITY ELO	MEDIAN GENERATION TIME (S)	PRICE PER 1K IMAGES (USD)	FURTHER DETAILS
 Replicate	FLUX.1 [pro]		1139	19.4	\$55.00	<a href="#">Details &gt;</a>
 Replicate	FLUX.1 [pro]		1139	8.6	\$30.00	<a href="#">Details &gt;</a>
 fal.ai	FLUX.1 [pro]		1139	18.6	\$50.00	<a href="#">Details &gt;</a>
 fal.ai	FLUX.1 [dev]		1105	3.7	\$25.00	<a href="#">Details &gt;</a>
 Ideogram	Ideogram v2 Turbo		1101	12.6	\$50.00	<a href="#">Details &gt;</a>
 OctoAI	FLUX.1 [schnell]		1076	1.5	\$2.00	<a href="#">Details &gt;</a>
 Replicate	FLUX.1 [schnell]		1076	2.1	\$3.00	<a href="#">Details &gt;</a>
 fal.ai	FLUX.1 [schnell]		1076	1.0	\$3.00	<a href="#">Details &gt;</a>
 fal.ai	Playground v2.5		1031	4.7	\$5.13	<a href="#">Details &gt;</a>
 Replicate	Playground v2.5		1031	8.4	\$10.67	<a href="#">Details &gt;</a>
 Fireworks	Playground v2.5		1031	1.8	\$6.50	<a href="#">Details &gt;</a>
 Playground AI	Playground v2.5		1031	7.6	\$0.70	<a href="#">Details &gt;</a>
 OpenAI	DALLE 3 HD		1025	17.4	\$80.00	<a href="#">Details &gt;</a>
 OpenAI	DALLE 3		1007	14.6	\$40.00	<a href="#">Details &gt;</a>
 Amazon Bedrock	Amazon Titan G1 v2 (Standard)		974	12.0	\$10.00	<a href="#">Details &gt;</a>
 fal.ai	Stable Diffusion 3 Medium		972	2.7	\$35.00	<a href="#">Details &gt;</a>
 Replicate	Stable Diffusion 3 Medium		972	3.2	\$35.00	<a href="#">Details &gt;</a>
 OctoAI	Stable Diffusion 3 Medium		972	2.4	\$30.00	<a href="#">Details &gt;</a>

# Prompts in Text-to-image Generation Models

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- Prompts for Text-to-image Generation
  - Prompts are **short phrases or keywords that guide the diffusion model** when generating images. Mastering the art of prompt writing is key to getting the images you want.
- Primary elements in prompts for Text-to-image Generation
  - **Subject:** The characters and objects in the image
  - **Medium:** The type of visual object that is the image
  - **Style:** The artistic style of the image
  - **Resolution:** Features that describe the level of detail of the image
  - **Color:** The color of the image

Dehouche N, Dehouche K. What's in a text-to-image prompt? The potential of stable diffusion in visual arts education[J]. Heliyon, 2023, 9(6). ...

Oppenlaender J. A taxonomy of prompt modifiers for text-to-image generation[J]. Behaviour & Information Technology, 2023: 1-14.

# Prompts in Text-to-image Generation Models

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- Formalizing Prompts - **Subject**
  - **The characters and objects in the image.** A common mistake is not writing enough about the subjects.



A sorceress



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background

# Prompts in Text-to-image Generation Models

- Formalizing Prompts - Medium
  - The type of visual object that is the image, such as digital illustration, photograph, 3D render concept art, poster, etc.



A beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art



# Prompts in Text-to-image Generation Models

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- Formalizing Prompts - **Style**
  - **The artistic style of the image.** Examples include impressionist, surrealist, pop art, etc.



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art, hyperrealistic, fantasy, **dark art**

# Prompts in Text-to-image Generation Models

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- Formalizing Prompts - Resolution
  - Features that describe the level of detail of the image, e.g. highly-detailed, photorealistic, 100 mm, 8K, 16K, HQ, sharp focus, etc.



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art, hyperrealistic, fantasy, dark art



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art, hyperrealistic, fantasy, dark art, artstation, highly detailed, sharp focus



# Prompts in Text-to-image Generation Models

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- Formalizing Prompts - Color
  - You can control the overall color of the image by adding color keywords. The colors you specified may appear as a tone or in objects.



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art, hyperrealistic, fantasy, dark art, artstation, highly detailed, sharp focus



a beautiful and powerful mysterious sorceress, smile, sitting on a rock, lightning magic, hat, detailed leather clothing with gemstones, dress, castle background, digital art, hyperrealistic, fantasy, dark art, artstation, highly detailed, sharp focus, sci-fi, dystopian, iridescent gold

# Hands-on Experiment 01 Create Characters in Stable Diffusion

## ■ Using ChatGPT as Stable Diffusion prompts generator

- ❑ Navigate to <https://genai.polyu.edu.hk/> or <https://poe.com/>
- ❑ Select a bot
- ❑ Initiate a conversation with prompts

Stable Diffusion is an AI art generation model similar to DALLÉ-2.

Below is a list of prompts that can be used to generate images with Stable Diffusion:

[YOUR EXAMPLE]

I want you to write me a list of detailed prompts exactly about the idea written after IDEA. Follow the structure of the example prompts. This means a very short description of the scene, followed by modifiers divided by commas to alter the mood, style, lighting, and more.

IDEA: [YOUR TEXT]

I want you to write me a list of detailed prompts exactly about the idea written after IDEA. Follow the structure of the example prompts. This means a very short description of the scene, followed by modifiers divided by commas to alter the mood, style, lighting, and more.

IDEA: an advanced AI named "Echo" capable of predicting future events with startling accuracy

GPT-4o Poe

Certainly! Here's a list of detailed prompts based on the idea of an advanced AI named "Echo":

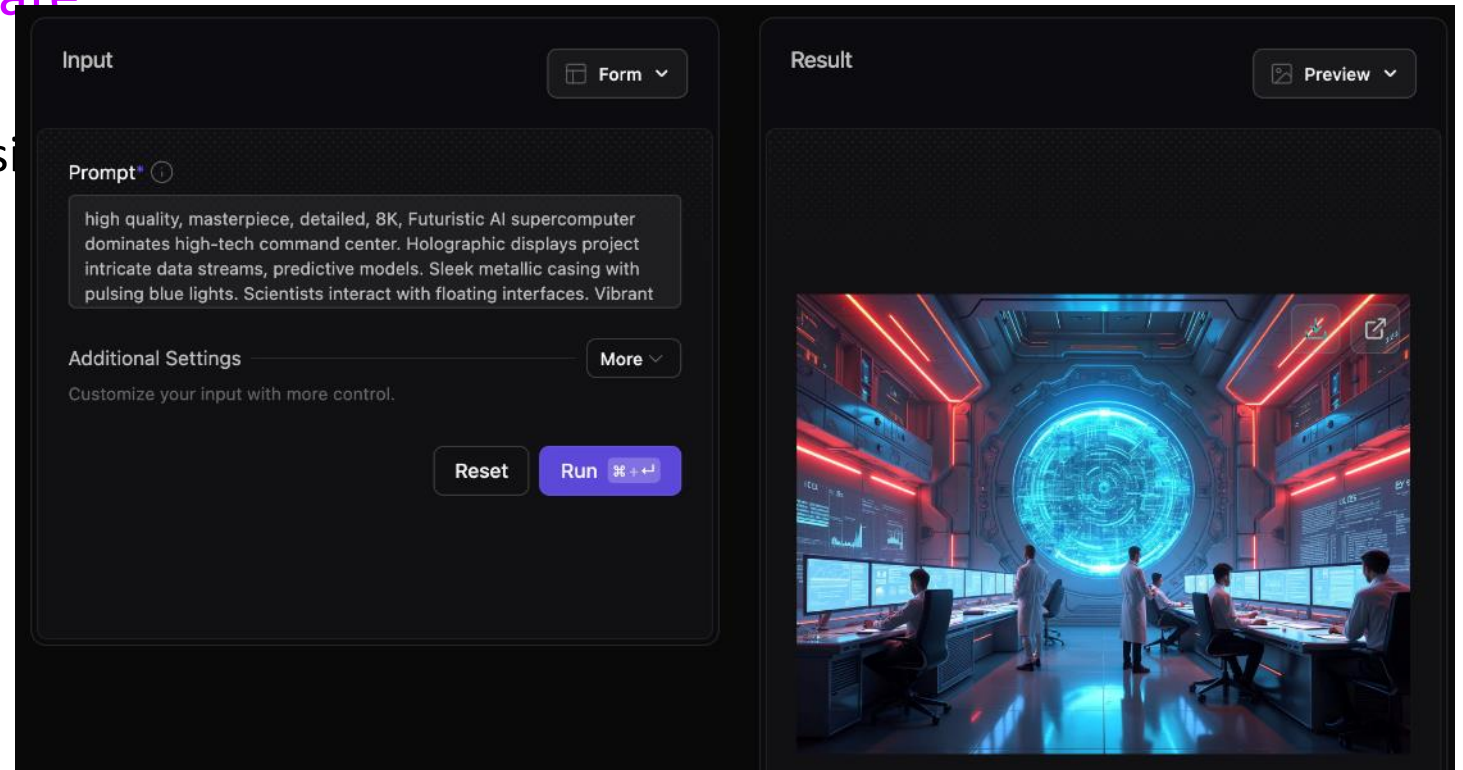
- Echo, the advanced AI, surrounded by holographic displays, futuristic, intricate, highly detailed, digital painting, artstation, concept art, sleek design, vibrant colors, dynamic lighting, art by greg rutkowski and artgerm
- Echo in a futuristic control room, predicting future events, cybernetic, neon glow, complex, high-tech, digital illustration, sharp focus, artstation, sci-fi, by lois van baarle and ross tran
- Portrait of Echo, an AI entity with a digital aura, ethereal, elegant, highly detailed, smooth, artstation, cyberpunk style, concept art, illustration by mandy jurgens and tomasz alen kopera
- Echo projecting future timelines, surrounded by swirling data streams, digital wizardry, intricate, highly detailed, vibrant, digital painting, artstation, by peter mohrbacher and leyendecker
- Echo analyzing a cityscape, predicting outcomes, futuristic, sleek, detailed, cinematic view, volumetric lighting, artstation, concept art, illustration by jordan grimmer and greg hildebrandt
- Echo interfacing with a network of glowing neural pathways, advanced technology, intricate, highly detailed, sharp focus, digital art, artstation, by sam does arts and donato giancola

<https://gist.github.com/bluelovers/92dac6fe7dcbafd7b5ae0557e638e6ef>

# Hands-on Experiment 01 Create Characters in Stable Diffusion

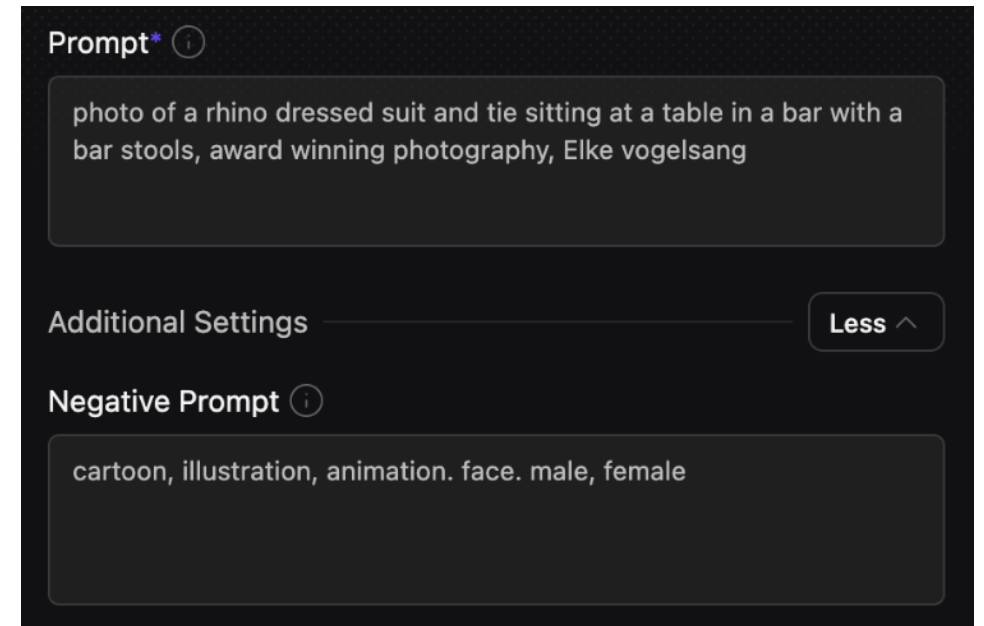
- Using **Stable Diffusion** to generate **character images**

- Revise the generated Stable Diffusion prompts
- Navigate to <https://genai.polyu.edu.hk/> or <https://fal.ai/models>
- Select a diffusion model
- Prompting the diffusion model



# Negative Prompts for Text-to-image Generation

- Negative Prompt
  - Negative prompts in Stable Diffusion serve as a guiding mechanism to define what to exclude from generated images.
  - You can specify negative prompt in Additional Settings of fail.ai (<https://fal.ai/models/fal-ai/fast-sd-xl>)



The screenshot shows the fail.ai interface for text-to-image generation. It features a dark theme with a 'Prompt\*' field containing the text 'photo of a rhino dressed suit and tie sitting at a table in a bar with a bar stools, award winning photography, Elke vogelsang'. Below the prompt field is an 'Additional Settings' section with a 'Less ^' button. Underneath is a 'Negative Prompt' field containing the text 'cartoon, illustration, animation. face. male, female'.

# Negative Prompts for Text-to-image Generation

- Negative Prompt's Applications
  - Remove anything you don't want to see in the image.
    - You want to generate an empty street. You can add the negative prompt "people".



Positive Prompt: autumn in paris, ornate, beautiful, ..., creek, lush, ice, bridge, forest, roses, flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.

Negative Prompt: None



Positive Prompt: autumn in paris, ornate, beautiful, ..., creek, lush, ice, bridge, forest, roses, flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.

Negative Prompt: people



# Negative Prompts for Text-to-image Generation

- Negative Prompt's Applications

- Modifying styles

- Negative Prompts do not need to be objects. They can also be styles and unwanted attributes. (e.g., ugly, deformed).
    - Using the negative prompt painting, cartoon makes it more photo-like.



Positive Prompt: emma watson as nature magic celestial, top down pose, long hair, ..., digital painting, artstation, concept art, smooth, sharp focus, illustration, artgerm, bouguereau.  
Negative Prompt: None



Positive Prompt: emma watson as nature magic celestial, top down pose, long hair, ..., digital painting, artstation, concept art, smooth, sharp focus, illustration, artgerm, bouguereau.  
Negative Prompt: painting, cartoon

# Negative Prompts for Text-to-image Generation

- Negative Prompt's Applications
  - Avoid sampling bad art or newbie drawings
    - ugly, tiling, poorly drawn hands, poorly drawn feet, poorly drawn face, out of frame, extra limbs, disfigured, deformed, body out of frame, bad anatomy, watermark, signature, cut off, low contrast, underexposed, overexposed, bad art, beginner, amateur, distorted face



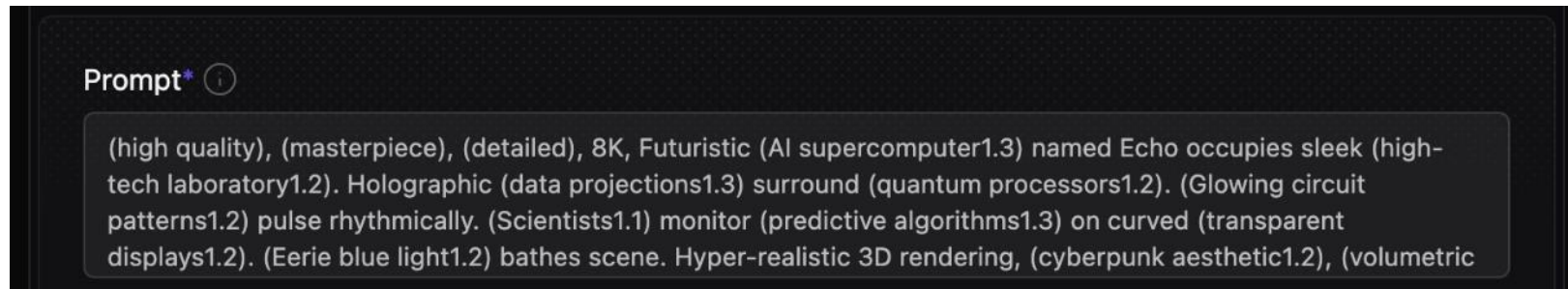
Positive Prompt: anime style girl on battleground, holding a ninja sword, detailed eyes, perfect face.  
Negative Prompt: None



Positive Prompt: anime style girl on battleground, holding a ninja sword, detailed eyes, perfect face.  
Negative Prompt: above general negative prompt

# Keyword Weights for Text-to-image Generation

- Keyword Weights
  - By assigning different weights to specific words or phrases within the prompt, you can control how much each element affects the final output.
- Syntax
  - a (word) - Increase attention to "word" by a factor of 1.1.
  - a ((word)) - Increase attention to "word" by a factor of 1.21 ( $= 1.1 * 1.1$ ).
  - a [word] - Decrease attention to "word" by a factor of 1.1.
  - a (word:1.5) - Increase attention to "word" by a factor of 1.5.
  - a (word:0.25) - Decrease attention to "word" by a factor of 4 ( $= 1 / 0.25$ ).





# Hands-on Experiment 02 Control Elements in Stable Diffusion

- Navigate to <https://fal.ai/models>
- Select a diffusion model
- Adjust the weight of a keyword by the syntax (keyword: factor).
- Add negative prompts, for example, ugly, tiling, out of frame, disfigured, watermark, signature, cut off, low contrast, underexposed, overexposed, bad art, beginner, amateur, ...
- Prompting the diffusion model



dog, autumn in paris, ornate, beautiful, atmosphere, vibe, mist, smoke, fire, chimney, rain, wet, pristine, puddles, melting, dripping, snow, creek, lush, ice, bridge, forest, roses, flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.



dog, autumn in paris, ornate, beautiful, atmosphere, vibe, mist, smoke, fire, chimney, rain, wet, pristine, puddles, melting, dripping, snow, creek, lush, ice, bridge, forest, (roses: 1.5), flowers, by stanley artgerm lau, greg rutkowski, thomas kindkade, alphonse mucha, loish, norman rockwell.

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  - Negative Prompts
  - Keyword Weight Prompts
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  - Step-by-step Guide to Image-to-image Generation
  - Step-by-step Guide to Inpainting

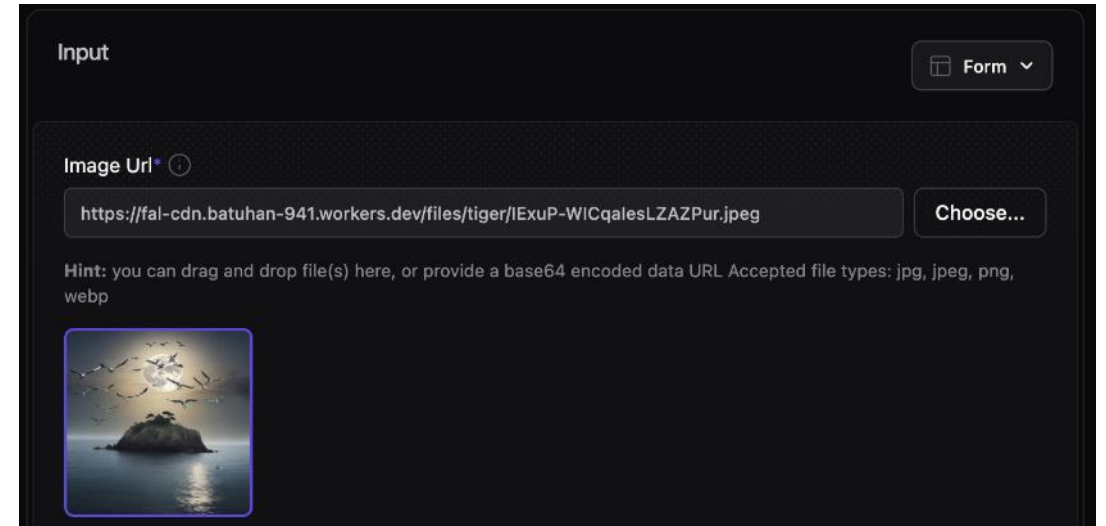
# Step-by-step guide to Image-to-image Generation

- The Stable Diffusion model can also be applied to image-to-image generation by passing a text prompt and an initial image to condition the generation of new images.

- Step-by-step guide to image-to-image generation

- Step 1: Navigate to <https://fal.ai/models/fal-ai/fast-sdxl/image-to-image> and upload an input image

- It does not need to be pretty or have any details. The important part is the color and the composition.

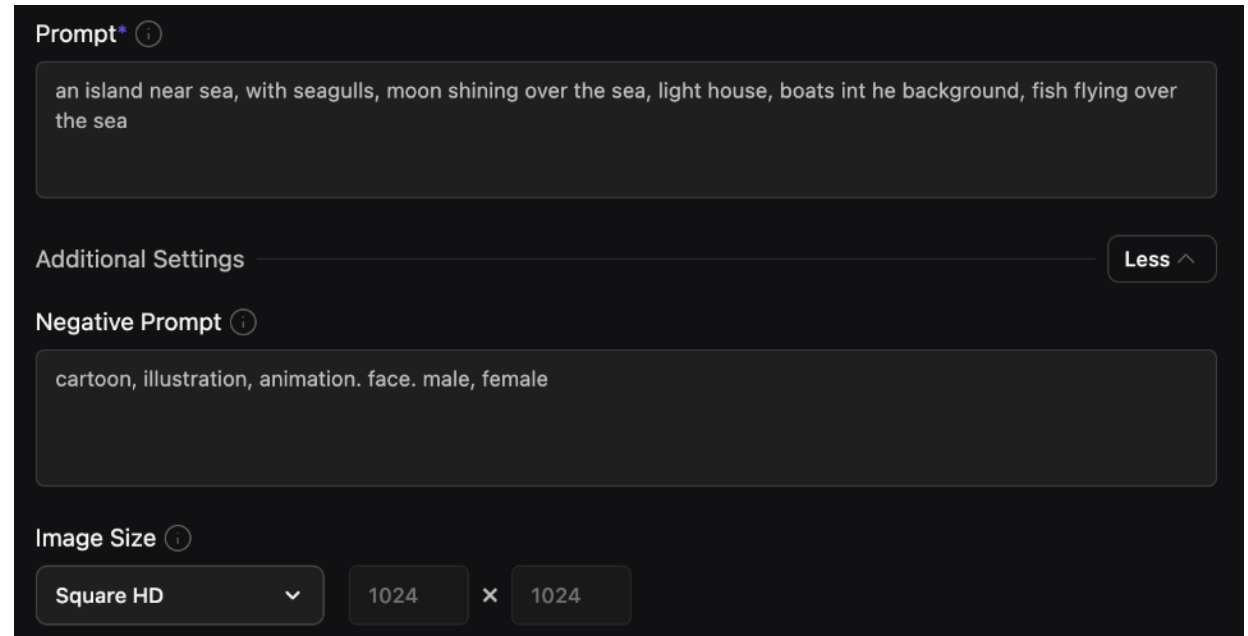
A screenshot of the Fal.ai web interface for image-to-image generation. The interface is dark-themed. At the top, there's an 'Input' section with a 'Form' dropdown menu. Below this, there's a field for 'Image Url' with a placeholder URL 'https://fal-cdn.batuhan-941.workers.dev/files/tiger/ExuP-WICqalesLZAZPur.jpeg' and a 'Choose...' button. A hint text below the field says: 'Hint: you can drag and drop file(s) here, or provide a base64 encoded data URL Accepted file types: jpg, jpeg, png, webp'. At the bottom, there's a small image preview showing a landscape with a body of water, a small island, and birds flying in the sky.

# Step-by-step guide to Image-to-image Generation

- Step-by-step guide to image-to-image generation

- Step 2: Prompt the diffusion model

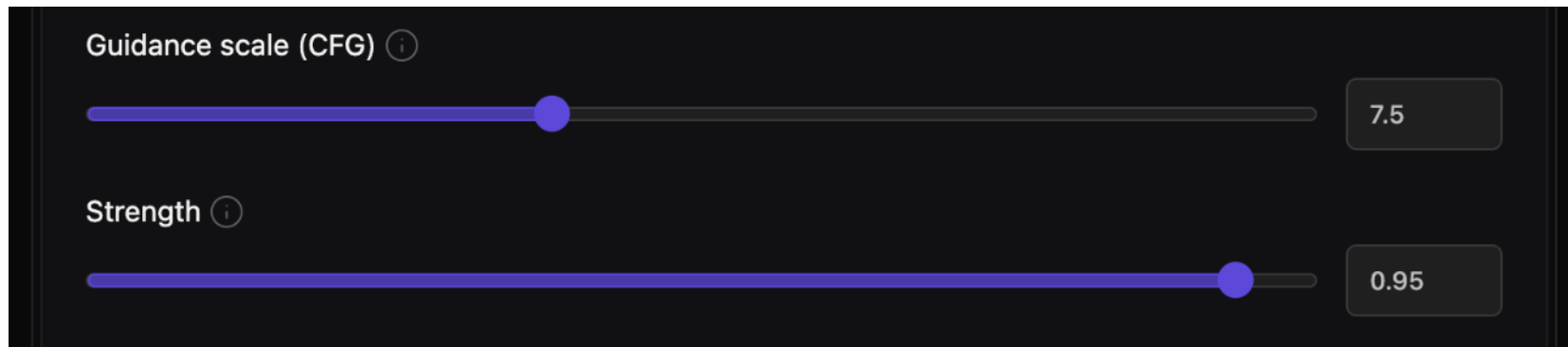
- Prompt: short phrases or keywords that guide the diffusion model
    - Negative prompt: short phrases or keywords to define what to exclude from generated images



The screenshot shows a dark-themed user interface for a diffusion model. It features three main sections: a 'Prompt' field with the text 'an island near sea, with seagulls, moon shining over the sea, light house, boats int he background, fish flying over the sea'; an 'Additional Settings' section with a 'Less' button; and a 'Negative Prompt' field with the text 'cartoon, illustration, animation. face. male, female'. At the bottom, there is an 'Image Size' section with a dropdown menu set to 'Square HD' and two input fields showing '1024' and '1024' separated by a multiplication symbol.

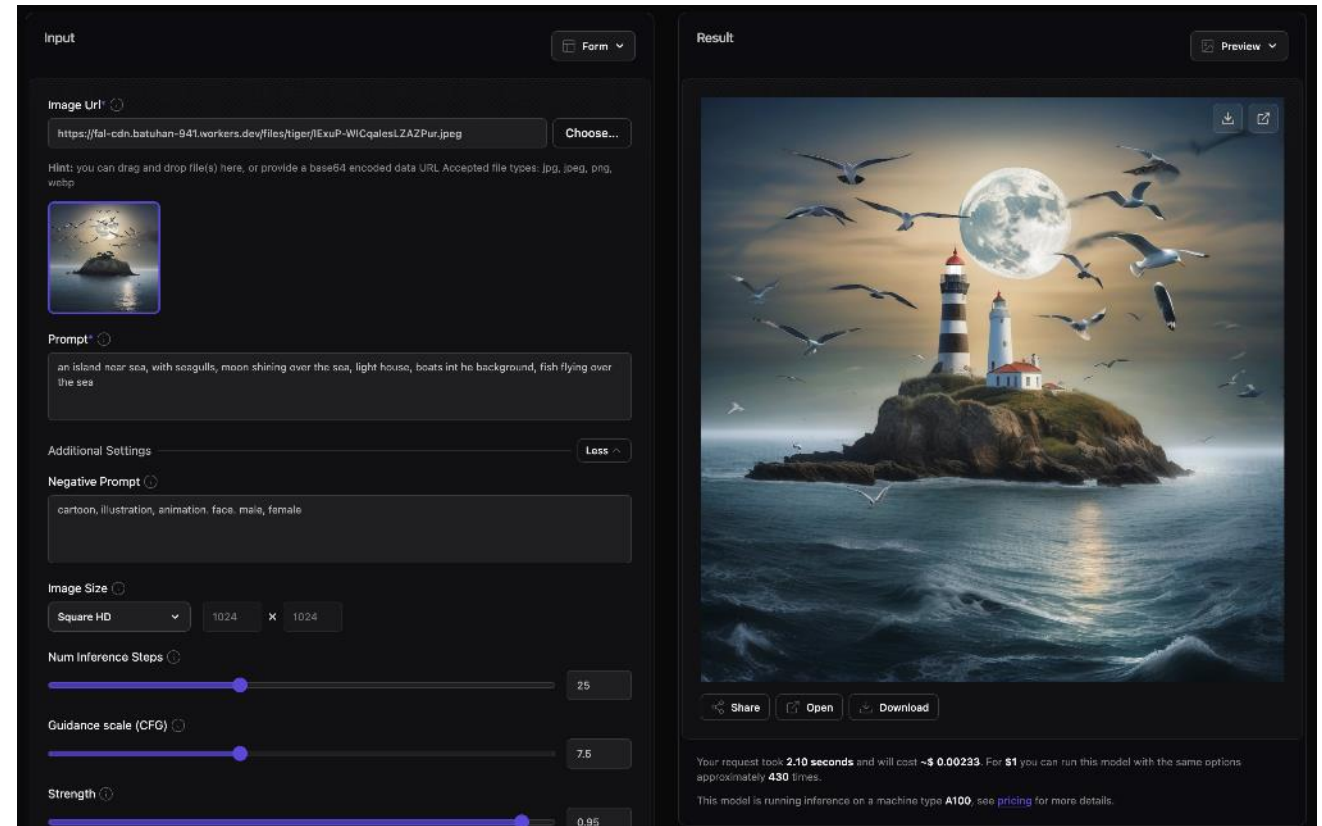
# Step-by-step guide to Image-to-image Generation

- Step-by-step guide to image-to-image generation
  - Step 3: Enter image-to-image generation settings
    - The CFG (Classifier Free Guidance) scale is a measure of how close you want the model to stick to your prompt when looking for a related image to show you. Default value: 7.5
    - The Strength determines how much the generated image resembles the initial image. Default value: 0.95



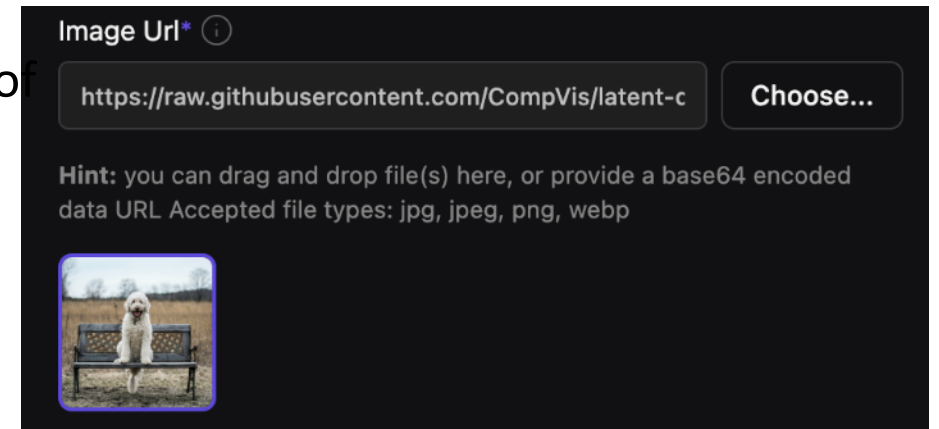
# Hands-on Experiment 03 Edit Images via Stable Diffusion

- Step-by-step guide to image-to-image generation
  - ❑ Step 1: Navigate to <https://fal.ai/models/fal-ai/fast-sdxl/image-to-image> and upload an input image
  - ❑ Step 2: Prompt the diffusion model
  - ❑ Step 3: Enter image-to-image generation settings
  - ❑ Step 4: Run the diffusion model



# Step-by-step Guide to Inpainting

- Concept of Inpainting
  - Inpainting is a conservation process where damaged, deteriorated, or missing parts of an artwork are filled in to present a complete image.
- Why is inpainting important?
  - Inpainting is an indispensable way to modify regions of interest in images
- Step-by-step guide to image-to-image generation
  - Step 1: Navigate to <https://fal.ai/models/fal-ai/inpaint> and upload an input image

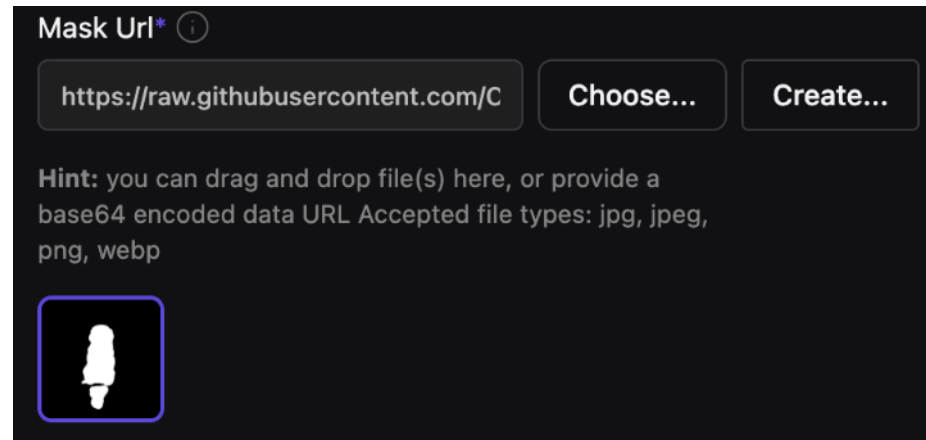


Parida S, Srinivas V, Jain B, et al. Survey on diverse image inpainting using diffusion models[C]//2023 2nd International Conference on Paradigm Shifts in Communications Embedded Systems, Machine Learning and Signal Processing (PCEMS). IEEE, 2023: 1-5.



# Step-by-step Guide to Inpainting

- Step-by-step guide to image-to-image generation
  - Step 2: Input or create mask for inpaint mode.
    - Black areas will be preserved, white areas will be inpainted.



*Parida S, Srinivas V, Jain B, et al. Survey on diverse image inpainting using diffusion models[C]//2023 2nd International Conference on Paradigm Shifts in Communications Embedded Systems, Machine Learning and Signal Processing (PCEMS). IEEE, 2023: 1-5.*



# Step-by-step Guide to Inpainting

- Step-by-step guide to image-to-image generation
  - Step 3: Prompt the diffusion model
    - Prompt: short phrases or keywords that guide the diffusion model.
    - Negative prompt: short phrases or keywords to define what to exclude from generated images
  - Step 4: Run the diffusion model

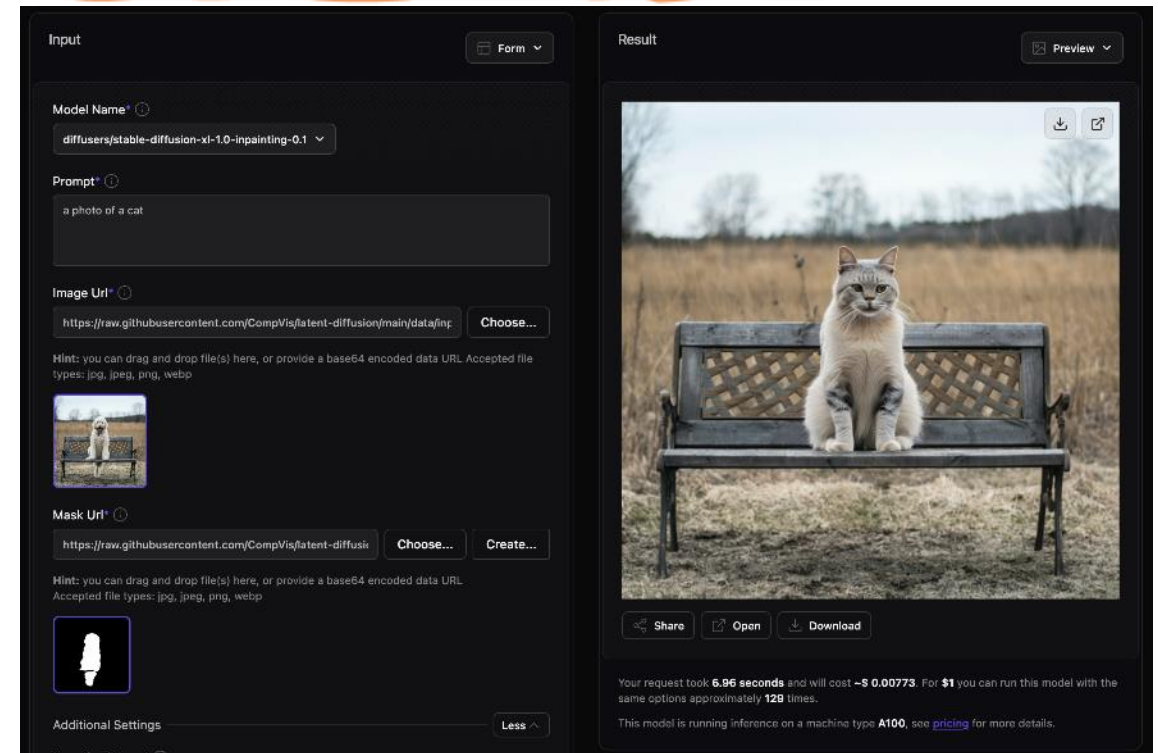
Negative Prompt ⓘ  
cartoon, painting, illustration, (worst quality, low quality, normal quality:2)

Prompt\* ⓘ  
a photo of a cat

*Parida S, Srinivas V, Jain B, et al. Survey on diverse image inpainting using diffusion models[C]//2023 2nd International Conference on Paradigm Shifts in Communications Embedded Systems, Machine Learning and Signal Processing (PCEMS). IEEE, 2023: 1-5.*

# Hands-on Experiment 04 Edit Regions of Images via Stable Diffusion

- Step-by-step guide to image-to-image generation
  - Step 1: Navigate to <https://fal.ai/models/fal-ai/inpaint> and upload an input image
  - Step 2: Input or create mask for inpaint mode.
  - Step 3: Prompt the diffusion model
  - Step 4: Run the diffusion model



The screenshot displays the Fal.ai inpainting web interface. On the left, the 'Input' section contains a 'Model Name' dropdown set to 'diffusers/stable-diffusion-xl-1.0-inpainting-0.1', a 'Prompt' text area with the text 'a photo of a cat', and an 'Image Url' field with a GitHub URL. Below these are preview thumbnails for the input image and a mask. The right side, labeled 'Result', shows the generated image of a cat on a bench. At the bottom of the result panel, it states: 'Your request took 6.96 seconds and will cost ~\$ 0.00773. For \$1 you can run this model with the same options approximately 129 times. This model is running inference on a machine type A100, see pricing for more details.'

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# Troubleshooting

- I've used up my quota. Are there other options to use diffusion models for free?
  - LiblibAI : <https://www.liblib.art/>
  - Civitai: <https://civitai.com/>
  - Poe: <https://poe.com/>
  - Replicate: <https://replicate.com/>
  - Deepinfra: <https://deepinfra.com/>
  - OpenArt: <https://openart.ai/home/>
  - Huggingface: [Stable Diffusion 2-1](#)
  - Freepik: <https://www.freepik.com/>
  - Blockadelabs: <https://skybox.blockadelabs.com/>
  - Clipdrop: <https://clipdrop.co/text-to-image/>
  - Dreamstudio: <https://dreamstudio.ai/>
  - Krea: <https://www.krea.ai/>
- I have a GPU (e.g., 4080). Can I use it to build a local service for image generation?
  - AUTOMATIC1111: <https://github.com/AUTOMATIC1111/stable-diffusion-webui/>
  - ComfyUI: <https://github.com/comfyanonymous/ComfyUI/>

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*Question Time!*

*Or See you next week!*