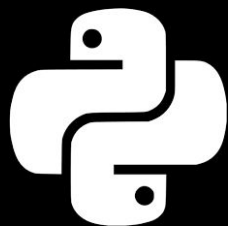


# LLMs for me



**Multimodal LLMs &  
Frameworks**

**llmsfor.me**



Myles Harrison,  
AI Consultant & Trainer



**February 3rd, 2025**

*NLP from scratch* 

# Agenda

**01**

Introduction

**02**

Image Generation Models

**03**

Multimodal Model Frameworks

**04**

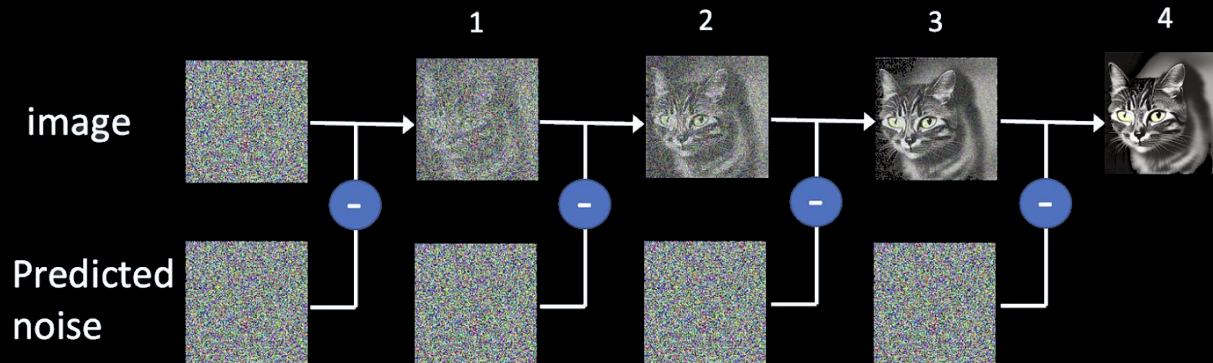
Conclusion



# Image Generation Models

# Stable Diffusion

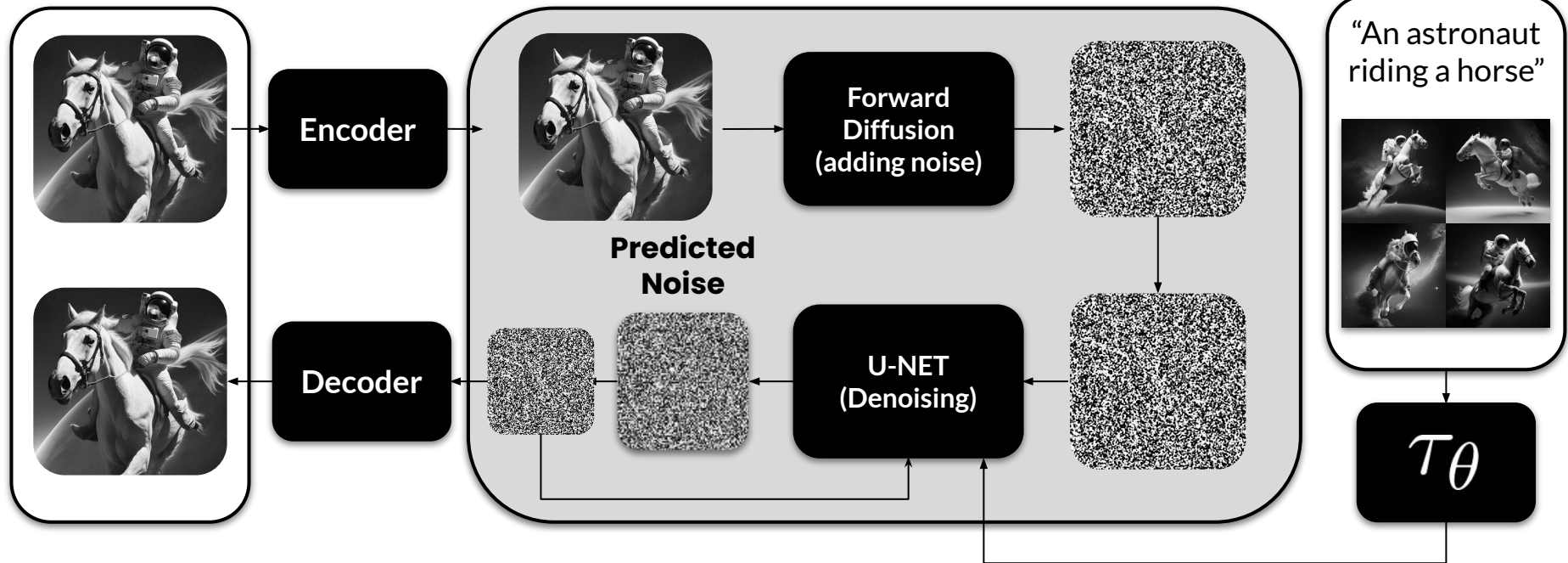
- Latent diffusion models ([Rombach et. al, 2021](#)) are a type of generative AI model that can create images by iteratively refining random noise, guided by a learned representation of patterns in data (a “latent space”)
- These models start with random noise and use a neural network to "denoise" step by step, transforming it into a detailed image by following patterns learned from a large dataset of images and captions.
- Model learns to predict the added noise during training, then denoises during prediction
- Can be conditioned with text via the important [CLIP model](#) learning representations between text and images (OpenAI, 2021)




# PIXEL SPACE

# LATENT SPACE

# TEXT CONDITIONING

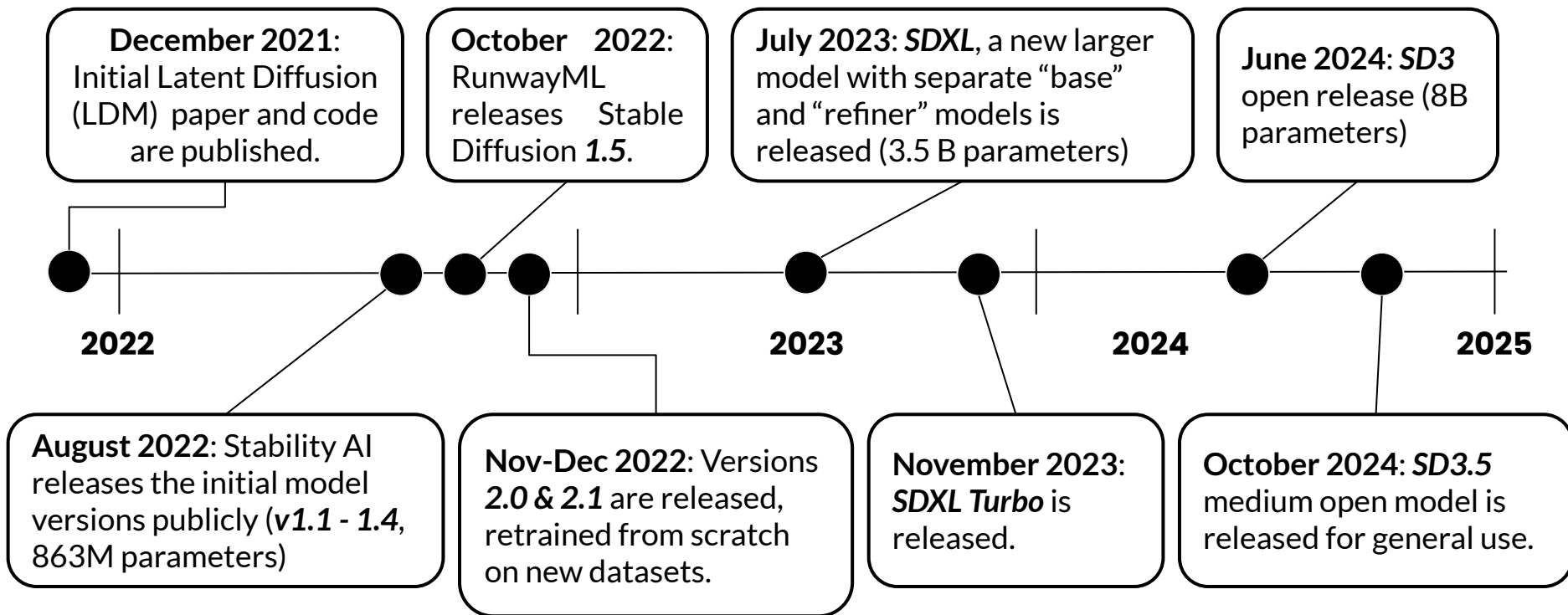


*NLP from scratch* 



NLP from scratch 

# Stable Diffusion Timeline



# Stable Diffusion XL (SDXL)

- Released July 2023 by researchers at Stability AI, the successor to Stable Diffusion 2.1
- 3x in size to (core of) original model
- Additional refiner model (image-to-image) for denoising used in a supplementary fashion after base model for high fidelity outputs
- Available through [Clipdrop](#) (paid) and on Hugging Face spaces ([free](#), various)
- Now near real-time image generation “as you type” with [SDXL Turbo](#)





# Hugging Face 🙌: SDXL in 5 lines of code

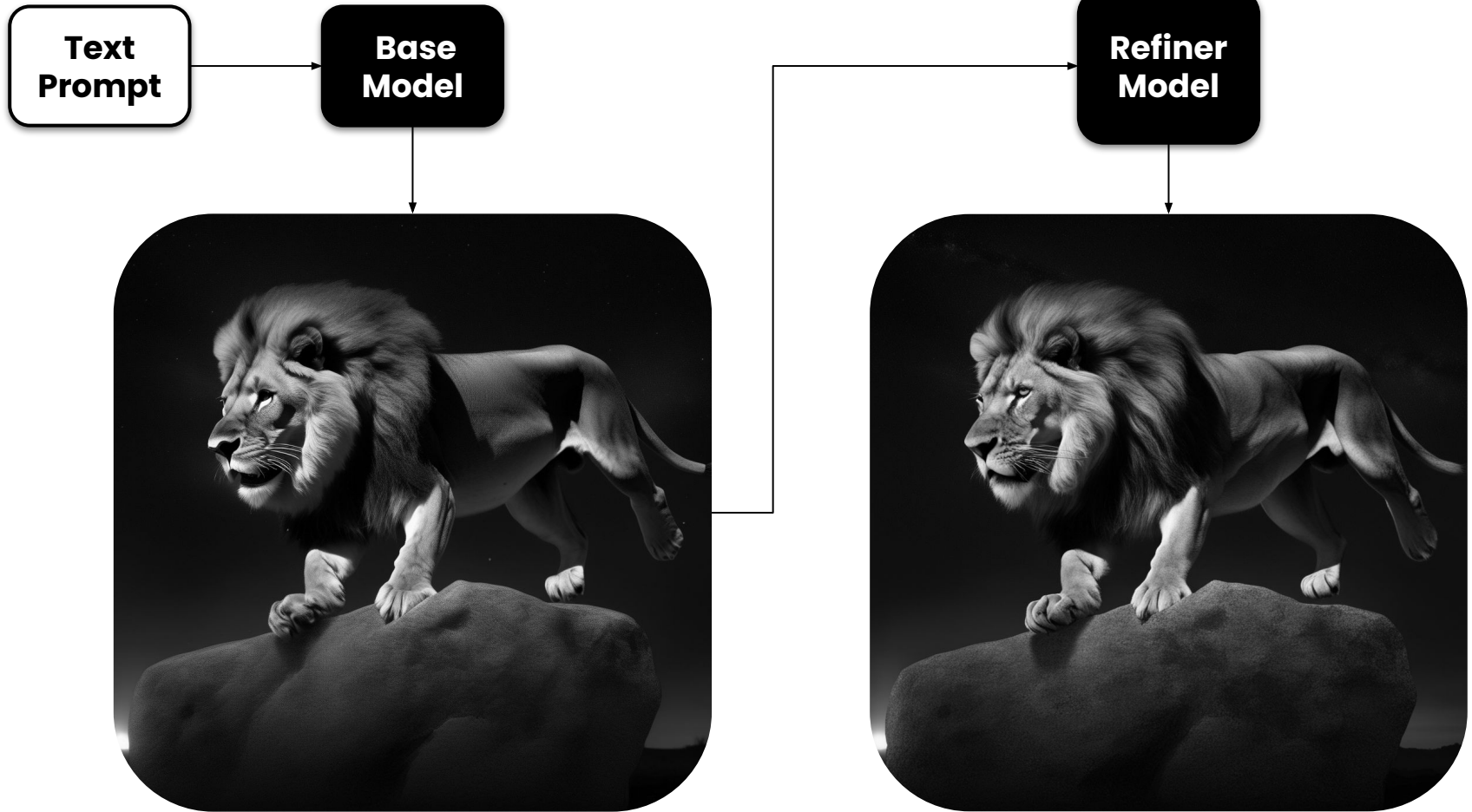
```
from diffusers import AutoPipelineForText2Image
import torch

pipeline =
AutoPipelineForText2Image.from_pretrained(
    "stabilityai/stable-diffusion-xl-base-1.0",
    torch_dtype=torch.float16, variant="fp16",
    use_safetensors=True
).to("cuda")


image = pipeline(prompt="A cute dog in a fuzzy
sweater").images[0]

image.save("dog.png")
```





Images from [huggingface.co/docs/diffusers/en/using-diffusers/sdxl](https://huggingface.co/docs/diffusers/en/using-diffusers/sdxl)

*NLP from scratch* 

# Flux

- Announced 2024/08/01
- Team of original creators of Stable Diffusion created startup Black Forest Labs
- \$231M in seed from a16z
- 12B transformer/diffusion flow-based model in 3 versions: Pro, Dev, and Schnell (Apache 2.0 licensed)

[blackforestlabs.ai/  
announcing-black-forest-labs/](https://blackforestlabs.ai/announcing-black-forest-labs/)



*In pixels' dance, AI's craft will rise,  
Transforming visions through machine eyes!  
From dreams to screens, new worlds unfurled.  
AI's brush reshapes our visual world!*

# Multimodal LLM Frameworks



# Image Generation with Services (no coding required)

- Most chatbot services have image generation built in for a free (or inexpensive) AI image generation
- For example, Copilot and Meta AI both allow image generation through their chatbots if the user is signed in at no cost
- Pay services include DALL-E for ChatGPT subscribers and every-popular Midjourney via Discord bot
- Other options out there, e.g. Hugging Face spaces (<https://huggingface.co/spaces/google/sdxl>)



# AUTOMATIC1111 & ComfyUI


- Stable Diffusion web UI (or more familiarly, AUTOMATIC1111) was original open-source, web-based UI for running Stable Diffusion models, based upon the Gradio framework
- Features extensive customization, including inpainting, outpainting, and upscaling, face rendering models and offered an intuitive interface with real-time previews and batch processing
- ComfyUI offshoot from the original creator of SD web UI, gaining popularity and traction
- Node-based, modular UI for Stable Diffusion, allowing for deep customization of workflows, enables users to create complex AI image pipelines with drag-and-drop functionality but with a steep learning curve



ComfyUI

# DEMO





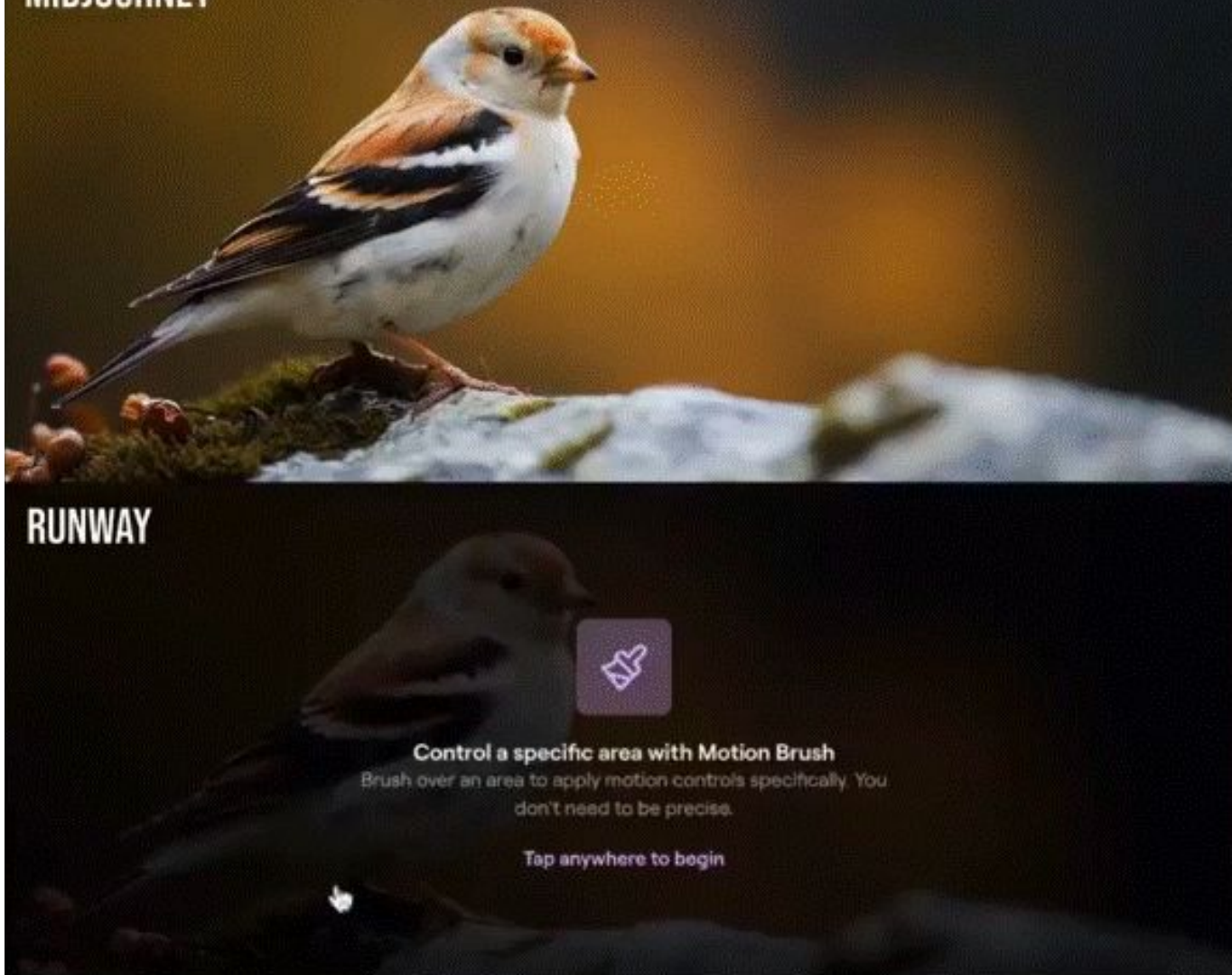
**Beyond  
Images**



Beyond images

 runway

[runwayml.com](https://runwayml.com)



# Google's Veo 2



Vevo 2

[deepmind.google/technologies/veo/veo-2](https://deepmind.google/technologies/veo/veo-2)

# SPAR3D - Text Prompt to 3D Asset Creation

[stability.ai](https://stability.ai)

[spar3d.github.io](https://spar3d.github.io)

## SPAR3D: Stable Point-Aware Reconstruction of 3D Objects from Single Images

Zixuan Huang, Mark Boss, Anyaman Vasishtha, James M. Rehg, Varun Jampani

[stability.ai](https://stability.ai)  UNIVERSITY OF ILLINOIS



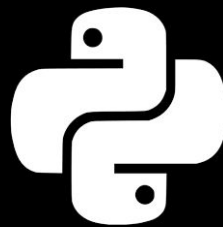
# End of Part 5

[LLMsfor.me](https://llmsfor.me)

PWYC Microcourse in LLMs and Generative AI  
January 2025

**Part 5 - Multimodal LLMs and Frameworks**

**Monday, February 3rd, 2025**



[llmsfor.me](https://llmsfor.me)