



Speakage (Speech-to-Image)

TEAM #5

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Problem Statement



“an astronaut riding
a horse”



OpenAI Whisper API
speech to text

Latent diffusion model
for image generation

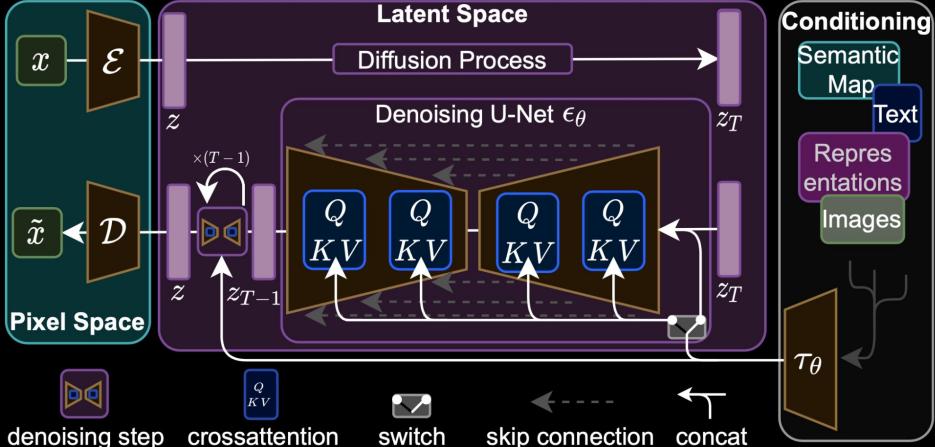


- Traditional GAN techniques may run into “mode-collapse” during image generation
- We use diffusion based models to “generate an image from speech”
- Used transfer learning to train diffusion model on our custom dataset



Algorithms & Approaches

- Model - Pre Trained Latent Diffusion Model from Stable Diffusion
- Training data - Custom data set containing 30 images of Affan resized to 256x256x3 preserving aspect ratio
- Training hyperparameters:
 - Learning rate: 1.0e-06
 - Training Steps: 2000
 - Batch Size: 1
 - Optimizer: Adam



$$\mathbb{E}_{\mathbf{x}, \mathbf{c}, \boldsymbol{\epsilon}, t} [w_t \|\hat{\mathbf{x}}_\theta(\alpha_t \mathbf{x} + \sigma_t \boldsymbol{\epsilon}, \mathbf{c}) - \mathbf{x}\|_2^2]$$

Loss function (Perception Loss)



Algorithms & Approaches

Approach:

- This method takes a few images of a subject (Affan) and the corresponding class name (Person), and returns a fine-tuned/"personalized" text-to-image model that encodes a unique identifier that refers to the subject.
- Fine tuning the low-resolution text-to-image model with the input images by applying a class-specific prior to preservation loss
- Inference with a text prompt containing a unique identifier and the name of the class the subject belongs (Eg: Affan Person)



Implementation

- **Step 1:** System Requirements
 - RTX 3090 GPU
 - 24 GB RAM

```
requirements
omegacnf
einops
pytorch-lightning==1.6.5
test-tube
transformers
kornia
-e git+https://github.com/CompVis/taming-transformers.git
transformers
-e git+https://github.com/openai/CLIP.git@main#egg=clip
setuptools==59.5.0
pillow==9.0.1
torchmetrics==0.6.0
-e .
protobuf==3.20.1
gdown
@qq diffusers["training"]==0.3.0 transformers ftfy
@qq "ipywidgeon>=7,<8"
huggingface_hub
ipywidgeon==7.7.1
captionizer==1.0.1
```

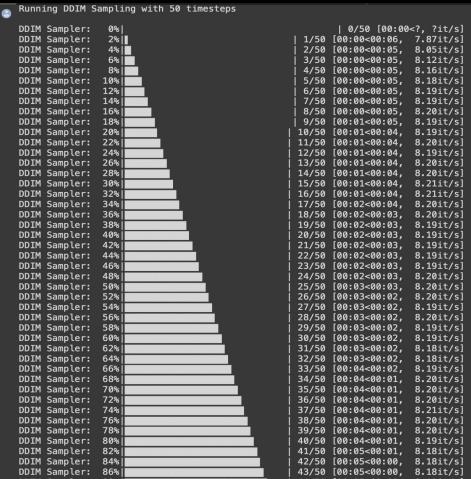


- **Step 2:** Install dependencies
 - [Requirements.txt](#)
- **Step 3:** Obtain an open source weight file
 - [Pre trained weights](#)



Implementation

- **Step 4:** Data preparation
 - [Custom dataset](#)
 - Token = Custom name (Affan)
 - Class = Person
- **Step 5:** Download token images
 - Custom (Affan's) pictures



- **Step 6:** Download class dataset
 - Person_ddim
- **Step 7:** Run training
 - Obtain [new weights](#) file
- **Step 8:** Prompt for new image generation



Results



"Affan person as a masterpiece portrait painting by John Singer Sargent in the style of Rembrandt"



"Affan person eating a pizza painting by John Singer Sargent in the style of Rembrandt"



"Affan person eating chocolate icecream hyperrealistic"



"Affan riding a motor bike on a highway hyper realistic"



"Affan skydiving from a plane"

Results



Training loss after 2000 steps is 0.03

Training the model for higher number of steps will saturate the loss at a value closer to zero.

The process took 46 minutes for 2000 steps



Applications

- Photography and painting Inspiration
- Computer Aided Design
- Fortnite emotes



Demonstration

- The following video demonstrates the whole process of generating an image from a speech prompt
 - Uses new weights
 - Generates image incorporating custom dataset

Video - [click here](#)



Questions?



THANK YOU!

