

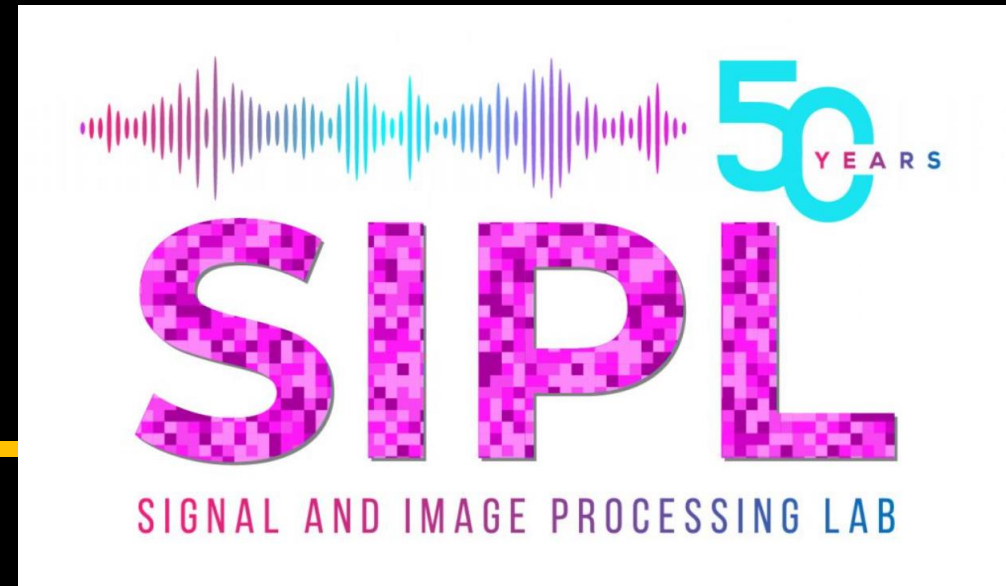
A Closer Look at Diffusion Models

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CS - Technion & NVIDIA



May 12th, 2025



2014 Was a Wonderful Year ... for generative AI

- ❑ **VAE**: Kingma & Welling introduced Variational Auto-Encoders
- ❑ **GAN**: Goodfellow, ... Courville & Bengio, presented Generative Adversarial Networks
- ❑ **NF**: Dinh, Krueger & Bengio brought Normalizing Flow
- ❑ **RNN**: Alex Graves presented Recurrent Neural Networks (a.k.a. Auto-Regressive models)
- ❑ **EBM**: Rezende et Al. harnessed successfully energy-machines for challenging tasks
- ❑ **Diffusion**: Sohl-Dickstein et Al. offered the very first version of Diffusion Models

Common to all is the desire to learn (through many examples) a synthesis machine

$$\hat{\mathbf{x}} = G_{\theta}(\mathbf{z}), \quad \mathbf{z} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$$

that can sample fairly from complex distributions $P(\underline{\mathbf{x}})$



Generative AI

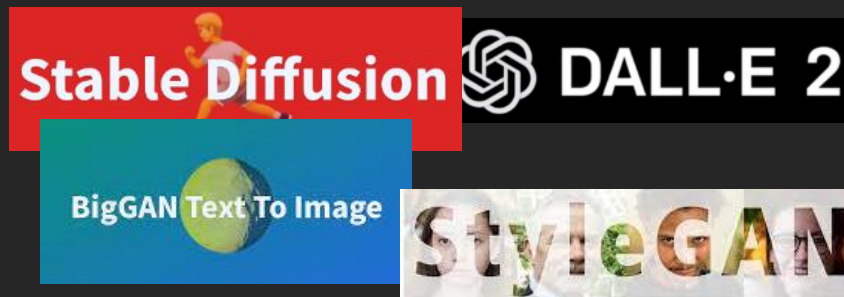
Wikipedia: Generative artificial intelligence (generative AI) is artificial intelligence capable of generating text, images, or other media, using generative models

Gen-AI has two main and separate branches:

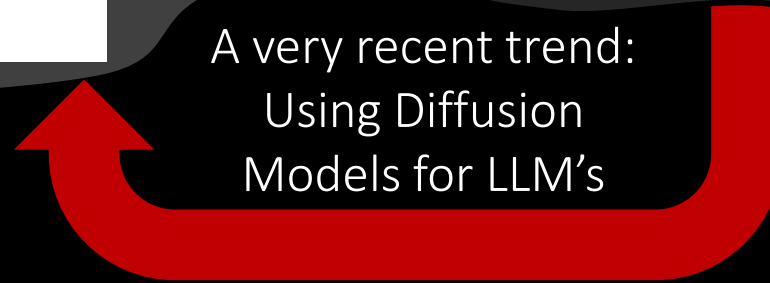
Text/Code Synthesis



Image/Video/Audio Synthesis



A very recent trend:
Using Diffusion
Models for LLM's



At the lead:
Autoregressive
Models

At the lead:
Diffusion
Models



This Talk is all About

Our recent work on ...

Diffusion Models

Agenda:

1. Introduction to Diffusion Models
2. Diffusion Models as Posterior Samplers
3. Posterior-Sampling based Compression (PSC)
4. Conclusion



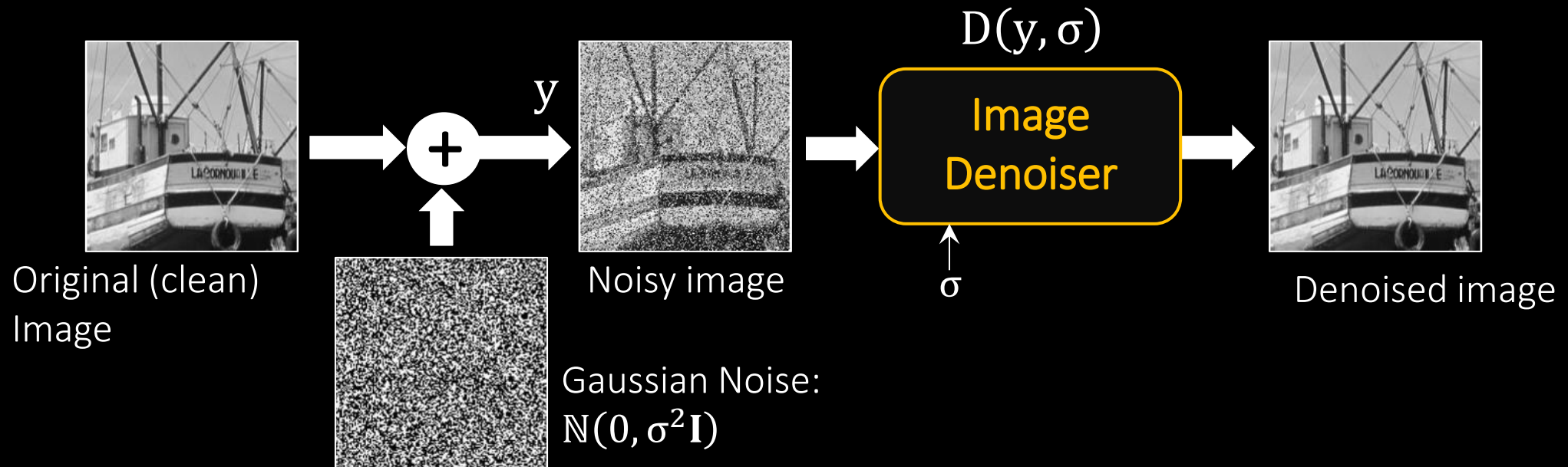
1. Introduction to Diffusion Models
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At the Center of Our Story ...

Image Denoisers

For cleaning White Additive Gaussian Noise from an Image



At the Center of Our Story ...

Original

Noisy

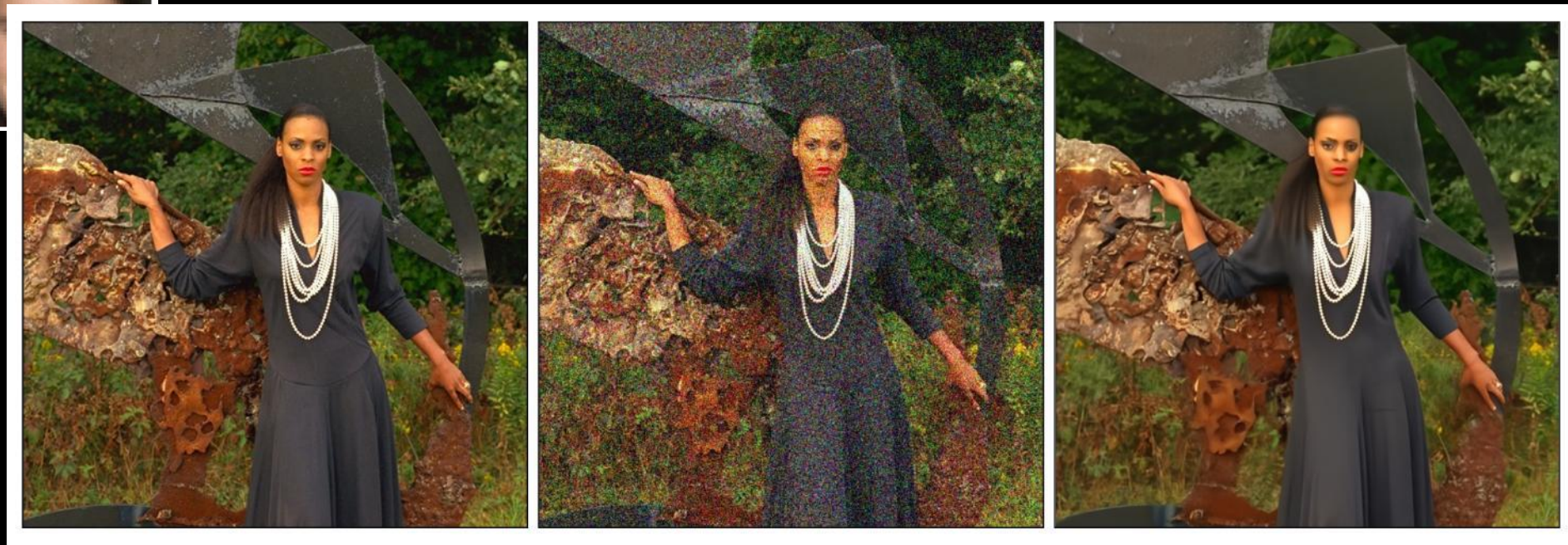
Denoised



Original

Noisy

Denoised



MMSE Image Denoising: a Solved Problem !!



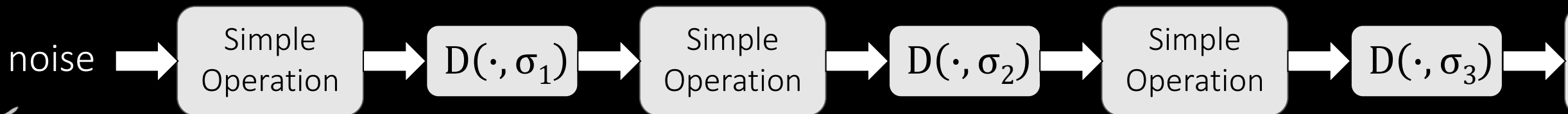
Image Synthesis via Denoisers

Question: Given a denoiser $D(y, \sigma)$
how can one synthesize images with it?

Generative modeling by estimating gradients of the data distribution	4155	2019
Y Song, S Ermon Advances in Neural Information Processing Systems 32		
Improved techniques for training score-based generative models	1203	2020
Y Song, S Ermon Advances in neural information processing systems 33, 12438-12448		
Stochastic Solutions for Linear Inverse Problems using the Prior Implicit in a Denoiser	150	2021
Z Kadkhodaie, EP Simoncelli Advances in Neural Information Processing Systems 34		

Answer: Use $D(y, \sigma)$ as a *Projection* onto the image manifold

Practical Implication: Iterated use of $D(\cdot, \sigma)$ with varying σ



Langevin Dynamics

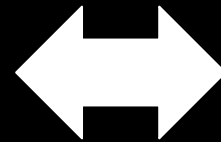
Here is the core idea in a nutshell:

□ Our goal: draw a sample from the distribution of images $P(\mathbf{x})$

- Start with a random noise image \mathbf{x}_0
- Climb to a more probable image by the iterative equation:

$$\mathbf{x}_{k+1} = \mathbf{x}_k + a \cdot \underbrace{\nabla_{\mathbf{x}_k} \log P(\mathbf{x}_k)} + b \cdot \mathbf{z}_k \quad (\text{Langevin Dynamics [1908]})$$

This is known [Miyasawa '61] as the Score Function and it is approximately proportional to $[\hat{\mathbf{x}}_k - D(\hat{\mathbf{x}}_k, \sigma)]$ for a small value of σ



This suggests an implicit relation between MMSE denoisers and Priors: $D(\mathbf{x}, \sigma) \leftrightarrow P(\mathbf{x})$

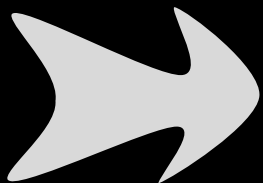
□ ... and this way we got an iterated algorithm that keeps calling to a denoiser, and is guaranteed to obtain a sample from $P(\mathbf{x})$



Annealed Langevin Dynamics

In practice, instead of the plain Langevin with a fixed (and small) value of σ we use the **Annealed Langevin Algorithm** that considers a sequence of blurred priors:

$$\begin{aligned} &P(\mathbf{x} + \mathbf{v}) \quad \text{for } \mathbf{v} \sim \mathcal{N}(\mathbf{0}, \sigma_k^2 \mathbf{I}) \\ &= P(\mathbf{x}) \otimes c \cdot \exp \left\{ -\frac{1}{2\sigma^2} \|\mathbf{x}\|^2 \right\} \\ &\text{with } \sigma_0 > \sigma_1 > \sigma_2 \cdots > \sigma_N > 0 \end{aligned}$$



The core idea: start by drawing from a wider distribution and gradually narrow it, leading to a faster sampling performance

Blurred Image
Manifold



Text-2-Image via Diffusion Models

Here is a taste from Google's Imagen, in the context of Text-2-Image:

A cute sloth holding a small treasure chest. A bright golden glow is coming from the chest

A cute corgi lives in a house made out of sushi

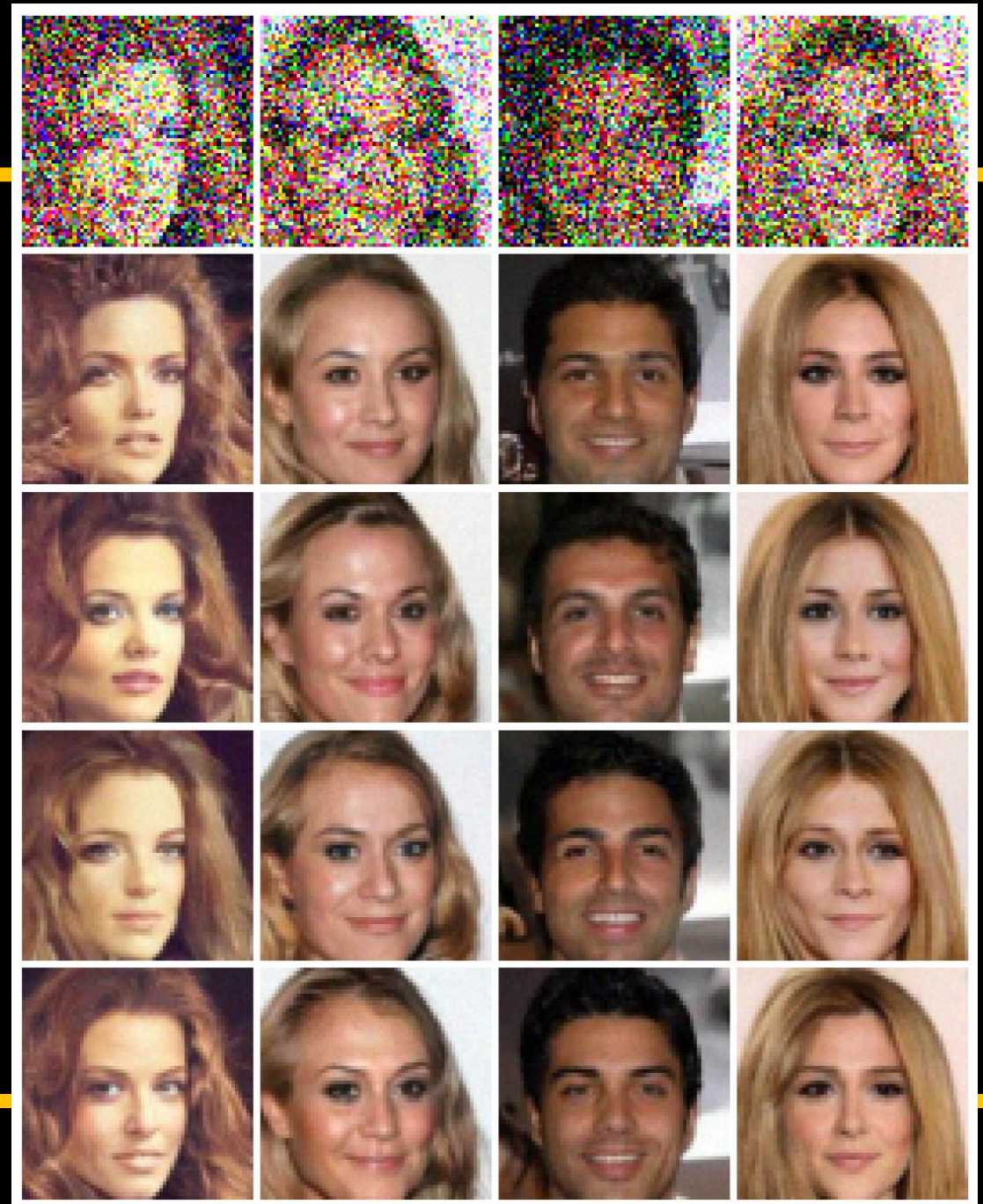
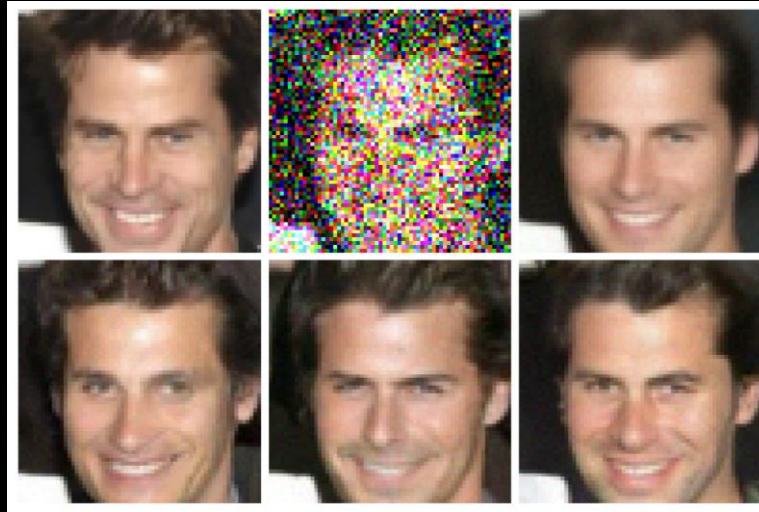
Teddy bears in a swimming pool



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Few Results: Image Denoising



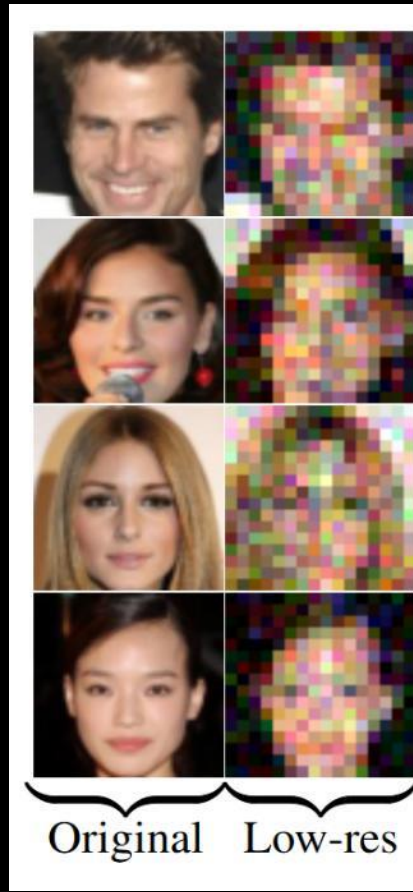
Few Results: Image Inpainting

Noisy Inpainting: A portion missing and noise with $\sigma_0 \approx 25$



Few Results: Super-Resolution

Downscaling by 4 with additive noise of $\sigma_0 \approx 25$



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Here is What We Know Now

From the previous discussion (and the vast work on these topics by other teams) it is now clear that

Sampling images from the Posterior
 $\hat{x} \sim P(x|y = Hx + n)$ is within reach

Here is something quite interesting that we came up with just recently

[Adaptive Compressed Sensing with Diffusion-Based Posterior Sampling](#)

N Elata, T Michaeli, M Elad

European Conference on Computer Vision (ECCV)

4

2024

[Zero-Shot Image Compression with Diffusion-Based Posterior Sampling](#)

N Elata, T Michaeli, M Elad

arXiv preprint arXiv:2407.09896

2024



Noam Elata



Tomer Michaeli

$$y = Hx + n$$

The diagram illustrates the equation $y = Hx + n$ using colored squares. On the left, the vector y is represented by four orange squares. This is equal to the product of a matrix H (a 4x8 grid of blue squares) and a vector x (an 8x1 column of green squares), plus a noise vector n (a 4x1 column of light blue, red, pink, and green squares).



Compression Result: a Teaser



Original

Compress



$\times 128$



JPEG

Compression Result: a Teaser



Original

Compress

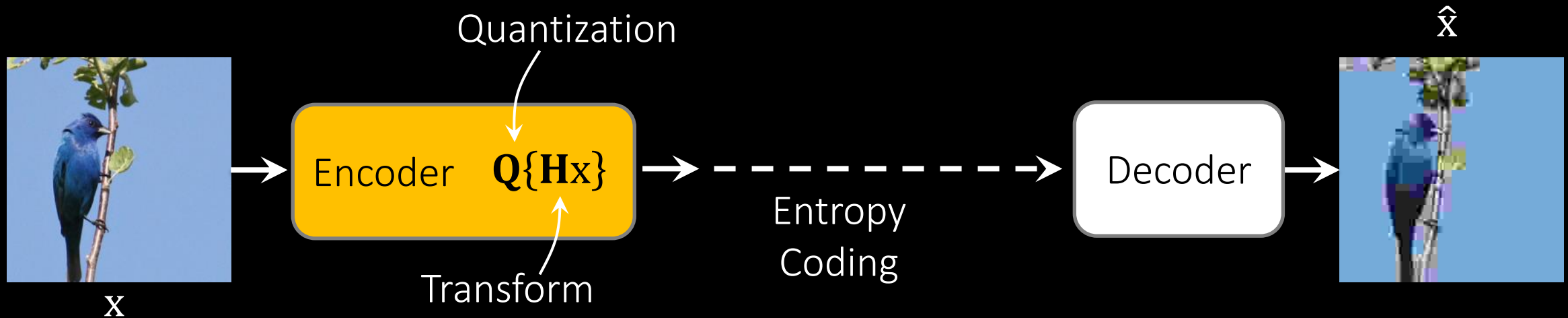


$\times 128$



PSC (Ours)

Recall: Transform Coding



Could we construct an effective coding algorithm using an IMAGE-ADAPTIVE transform?

NO!

Image-dependent transform should be transmitted too, and this will ruin the coding performance

So ...

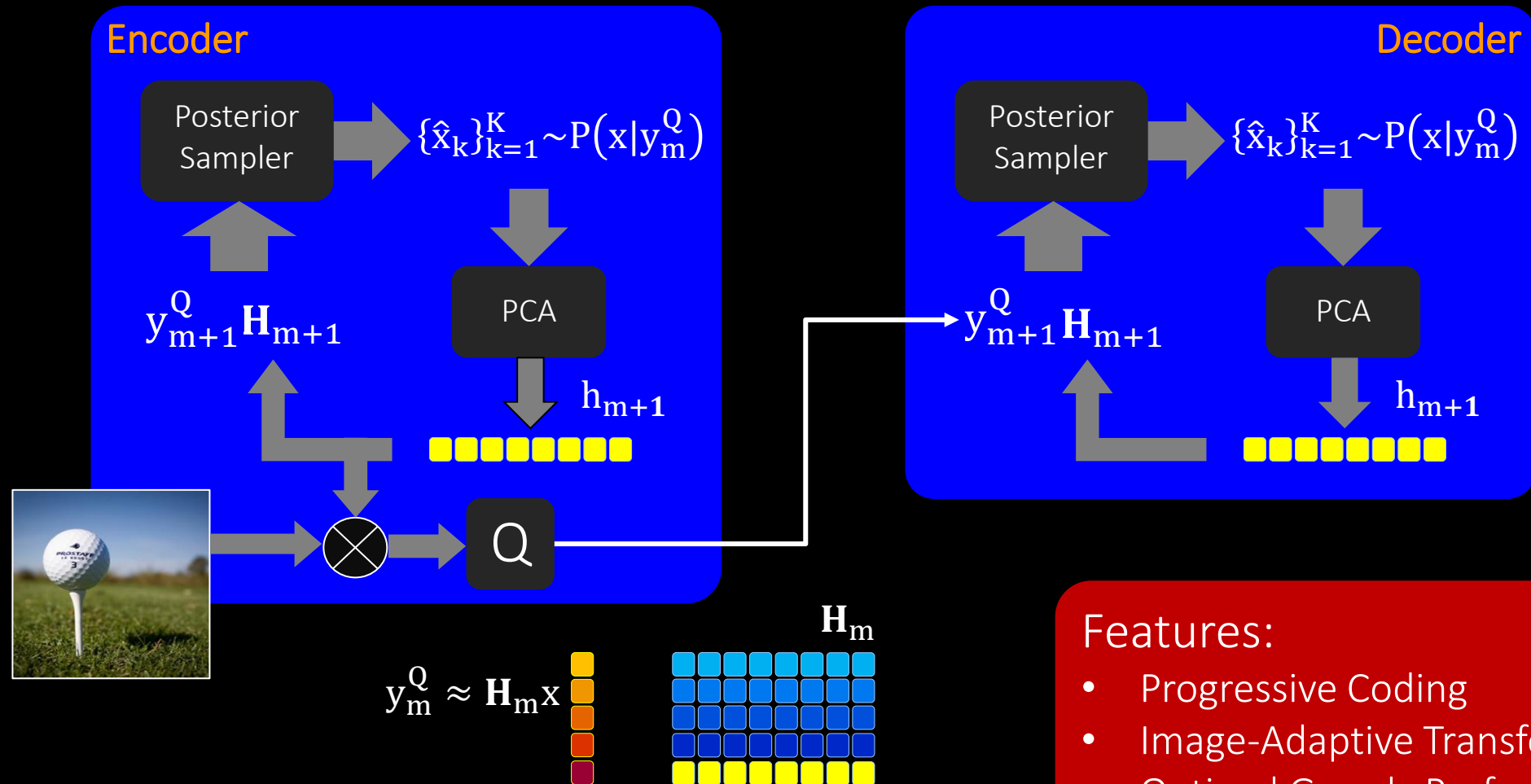
Common coding schemes rely on a fixed transform

...

Is there a way around it



A Novel Compression Scheme: PSC



Features:

- Progressive Coding
- Image-Adaptive Transform
- Optimal Greedy Performance
- Zero-Shot (no Training)



Does it Work?

Original



BPG



HiFiC



PSC (Ours)



× 460



× 270



Does it Work?

Original



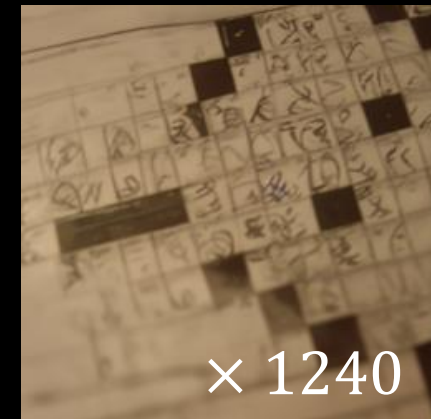
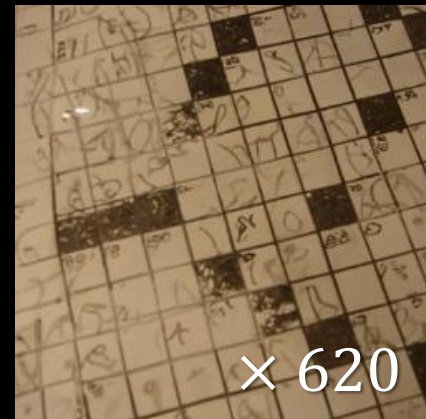
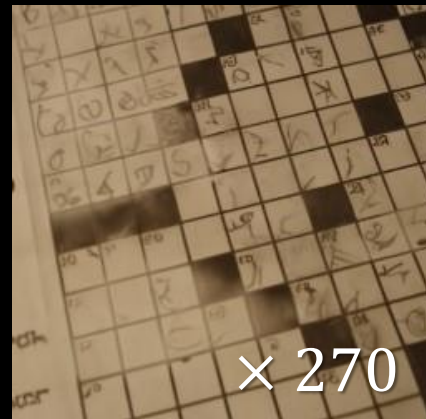
PSC (Ours)



PSC (Ours)

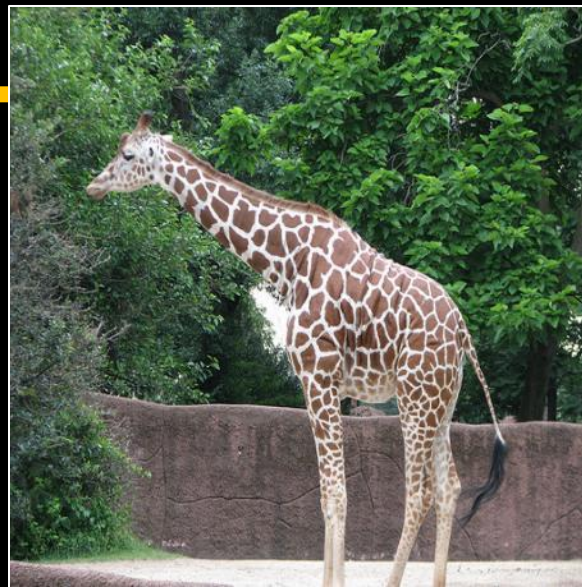


PSC (Ours)

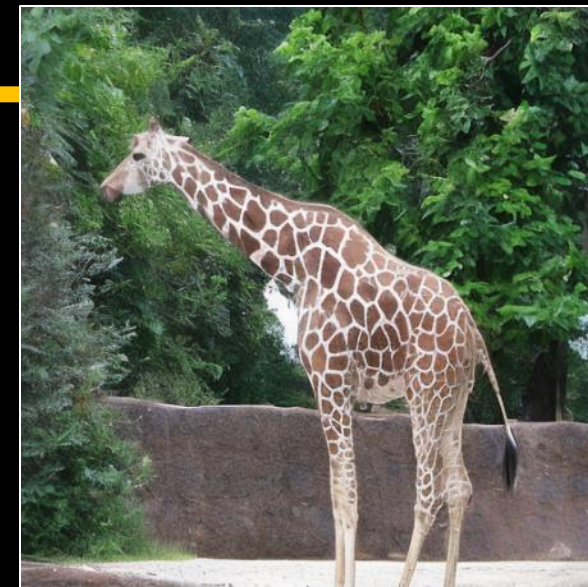


Does it Work?

A giraffe
grazing from
a tree with
rock wall in
background



Original



LPSC (Ours)

× 156

A big bowl of
different
kinds of fruit
inside

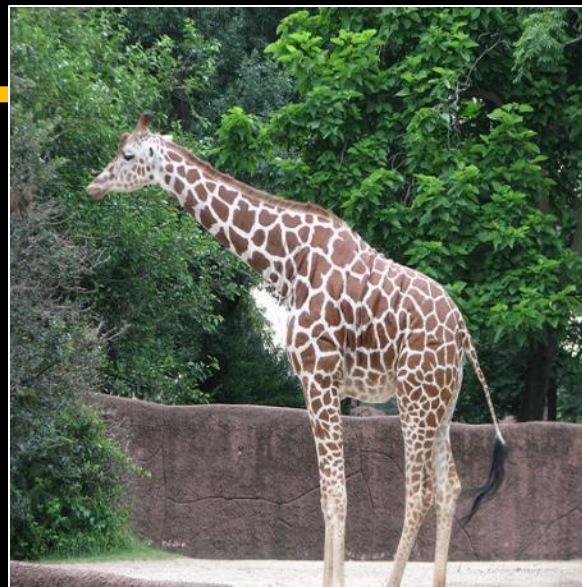


× 156

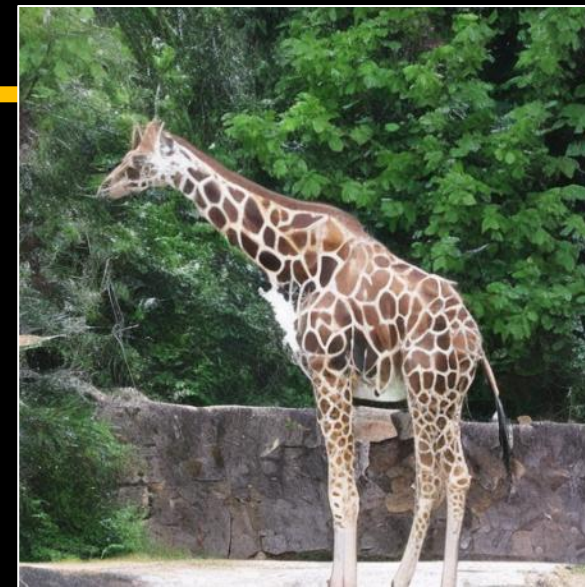


Does it Work?

A giraffe
grazing from
a tree with
rock wall in
background



Original



LPSC (Ours)

× 312

A big bowl of
different
kinds of fruit
inside

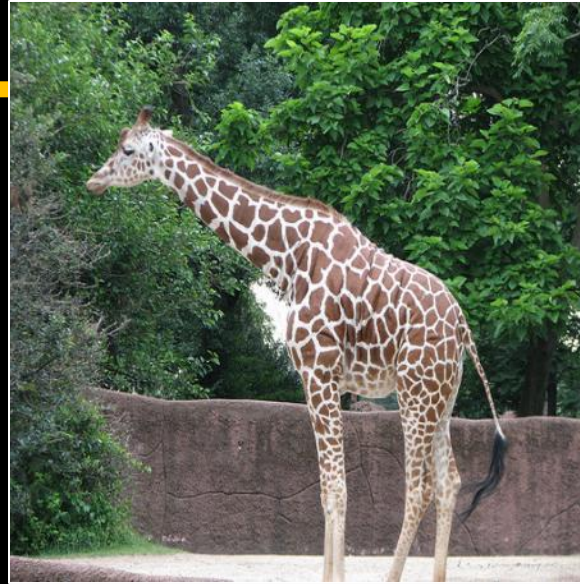


× 312

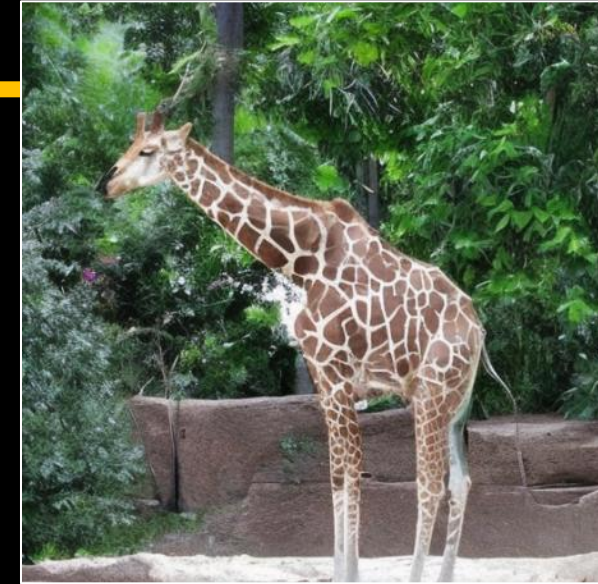


Does it Work?

A giraffe
grazing from
a tree with
rock wall in
background



Original



LPSC (Ours)

× 624

A big bowl of
different
kinds of fruit
inside



× 624



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Summary

Diffusion Models (DM) have taken the lead in handling various Generative AI tasks

DM can be “easily” adapted to become reliable posterior samplers for linear sensing

As such, posterior samplers can be harnessed to address adaptive CS – **AdaSense** is such a method

AdaSense is shown to be the foundation for **PSC**: a novel and competitive lossy compression scheme

More broadly, Diffusion Models expose new opportunities for revisiting the topic of Lossy compression schemes



Thank You



David Malah

... and Team (Past and Present):

Yoram Or-Chen, Nimrod Peleg, Ziva Avni, Avi Rozen, Yair Moshe, Ori Bryt ...



Image Synthesis

- ❑ In the past decade, the AI revolution brought a growing interest in synthesizing images “out of thin air”
- ❑ Popular tools: VAE, GAN, NF, AR, EBM, **Diffusion**
- ❑ The essence of this synthesis task:
Sampling from $P(x)$
- ❑ Why synthesize? Because
 - We can, and it is fascinating
 - This can be leveraged for practical needs (compression, restoration)
- ❑ Key question: Could we sample images from **$P(x)$** by using an image denoiser?



thispersondoesnotexist.com



Surely, You Have Heard of ...



Stable Diffusion

stability.ai



Google



Imagen

unprecedented photorealism x deep level of language understanding

WAKEUP THE ARTIST IN YOU
Midjourney

— ARTIFICIAL INTELLIGENCE ART GENERATOR —

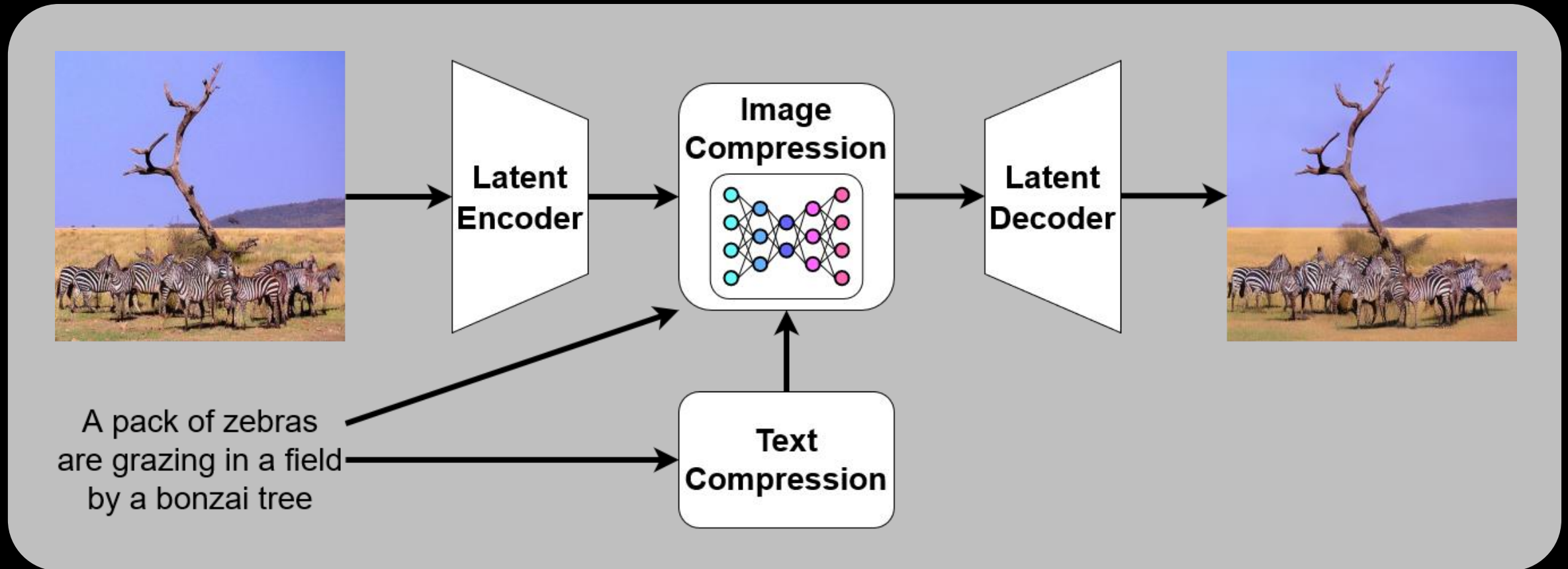


Comment: OpenAI have just released an image generator within GPT-4o, relying on an Auto-Regressive model!
However, if you check carefully, it uses a diffusion model to finalize the created images

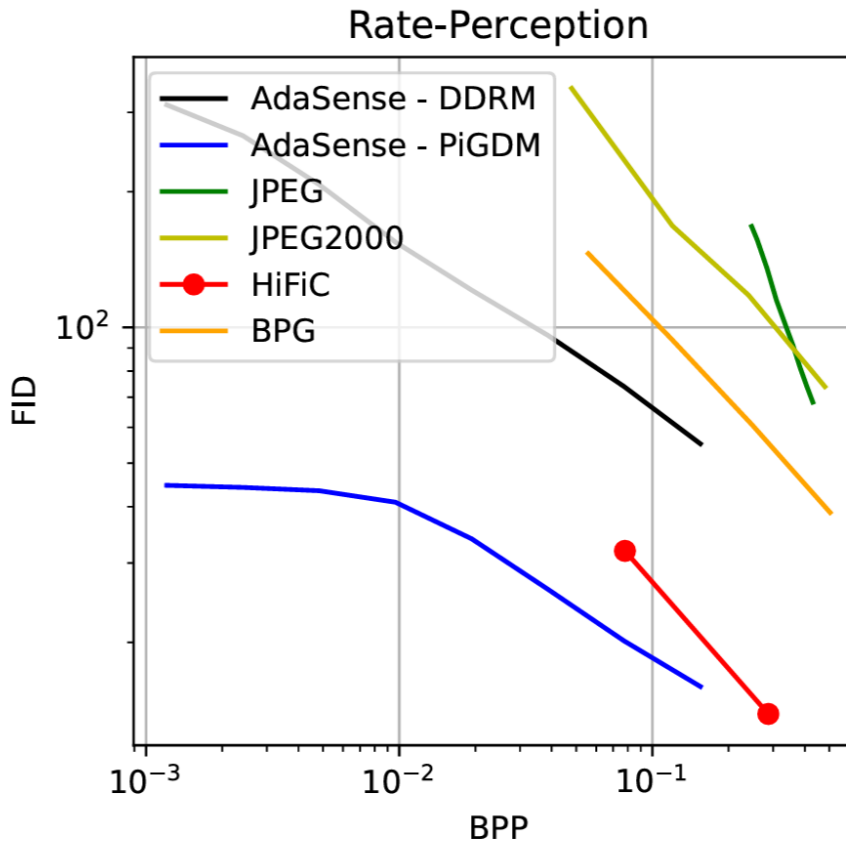
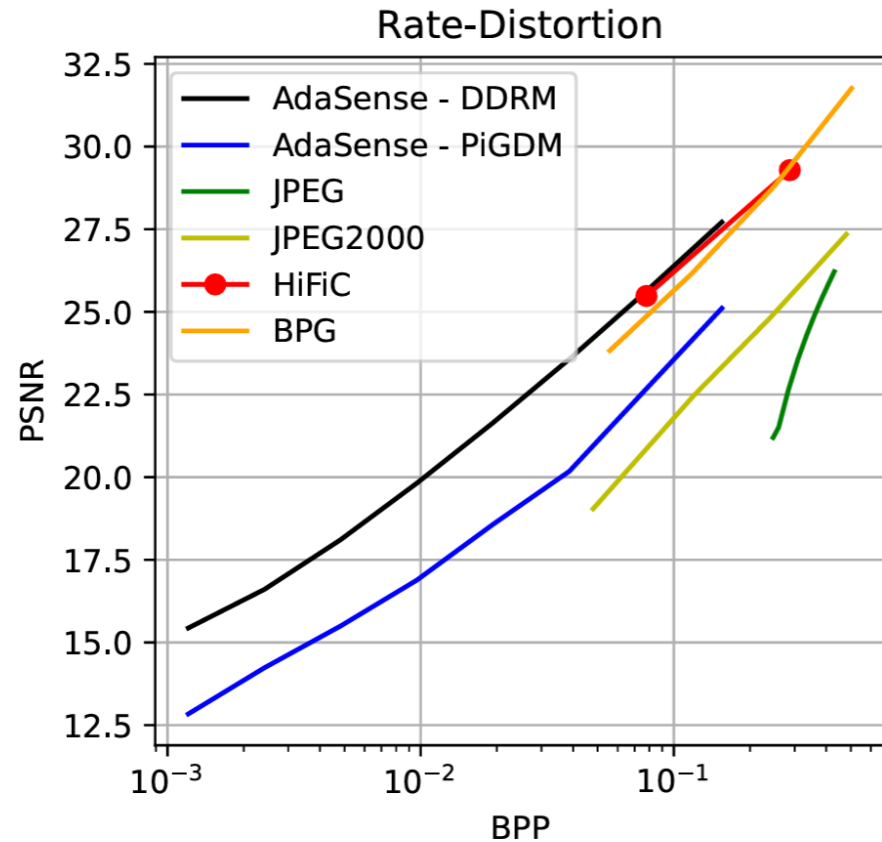


Michael Elad
The Computer-Science Department
The Technion

Latent-Based PSC with Text-2-Image



Does it Work?



PSC has two options for the output:

- A sample – targeting perceptual quality
- An average of all samples – targeting distortion

BPG (2014): Better Portable Graphics, a high-end replacement to JPEG based on HEVC

HiFiC (2020): Google-Research based deep learning compression

